## Arnold Beckmann • Laurent Bienvenu Nataša Jonoska (Eds.)

# I Pursuit of the Universal 

12th Conference on Computability in Europe, CiE 2016 Paris, France, June 27 - July 1, 2016 Proceedings

(1) Springer

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12th Conference on Computability in Europe, CiE 2016 Paris, France, June 27 - July 1, 2016 Proceedings

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Nataša Jonoska<br>Mathematics and Statistics<br>University of South Florida<br>Tampa, FL<br>USA

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## Preface

CiE 2016: Pursuit of the Universal Paris, France, June 27 - July 1, 2016


This year Computability in Europe (CiE) honored the 80th anniversary of A. Turing's paper introducing the Universal Turing Machine. In this context the conference sought better understanding of universal computational frameworks ranging from mathematics, computer science, through various natural sciences such as physics and biology. CiE provides a forum for exchanging ideas on broad aspects of "computability" striving to understand the essence of computation through studies of theoretical models of new paradigms, information processing, encryption, philosophy and history of computing as well as computability in natural and biological systems. This year's CiE conference was held in Paris and through a sequence of tutorials, plenary lectures, and special sessions allowed in-depth discussions and novel approaches in pursuit of the nature of computability. Similarly to previous CiE conferences in this series, CiE 2016 had a broad scope promoting the development of computability-related science.

The conference series is organized under auspices of the Association CiE. The association promotes the development of all areas of mathematics, computer science, as well as natural and engineering sciences that study the notion of "computability," including its philosophical and historical developments. The conference series is a venue where researchers in the field meet and exchange the most novel features of their findings.

CiE 2016 was organized jointly by Université Paris 13 and Université Paris 7, chaired by Paulin de Naurois at Université Paris 13. The previous CiE conferences were held in Amsterdam (The Netherlands) in 2005, Swansea (Wales) in 2006, Siena (Italy) in 2007, Athens (Greece) in 2008, Heidelberg (Germany) in 2009, Ponta Delgada (Portugal) in 2010, Sofia (Bulgaria) in 2011, Cambridge (UK) in 2012, Milan (Italy) in 2013, Budapest (Hungary) in 2014, and Bucharest (Romania) in 2015. The proceedings containing the best submitted papers as well as extended abstracts of invited speakers for all these meetings are published in the Springer series Lecture

Notes in Computer Science. The annual CiE conference has risen to be the largest international meeting focused on computability theory issues. CiE 2017 will be held in Turku, Finland. The leadership of the conference series recognizes that there is under-representation of female researchers in the field of computability and therefore incorporates a special session of Women in Computability (WiC) in every CiE conference. WiC was initiated in 2007, and was first funded by the Elsevier Foundation, later taken over by the publisher Elsevier. This year's program, organized by Liesbeth De Mol, besides the regular workshop also provided travel grants for junior female researchers and a mentorship program.

The 39 -member Program Committee of CiE 2016 was chaired by Laurent Bienvenu (IRIF, CNRS, and Université Paris 7, France), and Nataša Jonoska (University of South Florida, Tampa, USA). The committee selected the plenary speakers and the special session organizers, and ran the reviewing process of all the regular contributions submitted. We received 40 non-invited contributed paper submissions. Each paper received at least three reviews by the Program Committee and additional reviewers. About $45 \%$ of the submitted papers were accepted for publication in this volume. In addition, this volume contains 19 extended abstracts/papers contributed by plenary speakers and speakers of the invited sessions. The production of the volume would have been impossible without the diligent work of all of the Program Committee members and our expert reviewers. We are very grateful to all the Program Committee members and the reviewers for their excellent work.

All authors who contributed to this conference were encouraged to submit significantly extended versions of their papers with unpublished research content to Computability, the journal of the Association CiE.

This year the conference started with a special session honoring the memory of Barry Cooper, one of the initiators and founders of the conference as well as a driving force behind the organization of CiE , including the presidency of the association. The session was organized by Mariya Soskova and the contributors were Theodore Slaman (University of California Berkeley), Andrea Sorbi (University of Siena), Dag Norman (University of Oslo), and Ann Copestake (Cambridge University).

Two tutorials were given by Bernard Chazelle from Princeton University, USA, and Mikolaj Bojanczyk from University of Warsaw, Poland. In addition, the Program Committee invited seven speakers to give plenary lectures: Natasha Alechina (University of Nottingham, UK), Vasco Brattka (Universität der Bundeswehr München, Germany), Delaram Kahrobaei (The City University of New York, USA), Steffen Lempp (University of Wisconsin, USA), André Nies (University of Auckland, New Zealand), Dominique Perrin (Université Paris-Est Marne-la-Vallée, France), and Reed Solomon (University of Connecticut, USA).

Springer generously funded two awards this year, the Best Student Paper Award and Best Paper Award. The winner of the Best Student Paper Award this year was Mikhail Andreev for his contribution "Busy Beavers and Kolmogorov Complexity." The Best Paper Award was given to Olivier Bournez, Nachum Dershowitz and Pierre Neron for their contribution "An Axiomatization of Analog Algorithms."

CiE 2016 has six special sessions: two sessions, Cryptography and Information Theory and Symbolic Dynamics, were organized for the first time in the conference series. The other four special sessions covered new developments in areas previously covered by the conference series: Computable and Constructive Analysis, Computation in Biological Systems, Weak Arithmetic, and History and Philosophy of Computing. Speakers in these special sessions were selected by the special session organizers and were invited to contribute a paper to this volume.

## Computable and Constructive Analysis

Organizers. Daniel Graça and Elvira Mayordomo.
Speakers. Mathieu Hoyrup (Inria and University of Lorraine), Arno Pauly (University of Cambridge), Vela Velupillai (New School for Social Research in New York City and University of Trento), Martin Ziegler (KAIST, Daejeon).

## Computation in Biological Systems

Organizers. Alessandra Carbone and Ion Petre.
Speakers. Daniela Besozzi (University of Milan-Biccocca), Eugen Czeizler (Åbo Akademi University), Vincent Moulton (University of East Anglia), Eric Tannier (Inria and University of Lyon).

## Cryptography and Information Theory

Organizers. Danilo Gligoroski, and Carles Padro.
Speakers. Ludovic Perret (Université Pierre et Marie Curie, France), Ignacio Cascudo (Aarhus University in Denmark), Oriol Farras (Universitat Rovira i Virgili, Spain), Danilo Gligoroski (Norwegian University of Science and Technology Trondheim).

## History and Philosophy of Computing

Organizers. Alberto Naibo and Ksenia Tatarchenko
Speakers. Maël Pégny (IHPST, Paris), Pierre Mounier-Khun (CNRS), Simone Martini (University of Bologna), Walter Dean (University of Warwick).

## Symbolic Dynamics

Organizers. Jarkko Kari and Reem Yassawi.
Speakers. Valérie Berthé (University of Paris 7), Emmanuel Jeandel (University of Lorraine), Irène Marcovici (University of Lorraine), Ronnie Pavlov (Denver University).

## Weak Arithmetic

Organizers. Lev Beklemishev and Stanislav Speranski.
Speakers. Pavel Pudlák (Academy of Sciences of the Czech Republic), Alexis Bès (University of Paris 12), Leszek Kołodziejczyk (University of Warsaw), Albert Visser (University of Utrecht).

The organizers of CiE 2016 would like to acknowledge and thank the following for their financial support (in alphabetical order): the Association for Symbolic Logic (ASL), the European Association for Theoretical Computer Science (EATCS), and Springer. We would also like to acknowledge the support of our non-financial sponsor, the Association Computability in Europe (CiE).

April 2016
Arnold Beckmann
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## S. Barry Cooper 1943-2015



Barry Cooper at the opening of CiE 2009 in Heidelberg. Photo taken by Peter van Emde Boas, July 2009.

Barry Cooper, founding member and former president of the Association Computability in Europe, died on October 26, 2015, shortly after his 72 nd birthday. Born on October 9, 1943, Barry was a leading figure in the UK logic scene all of his academic life, a major figure in computability theory, and in particular degree theory. Most relevant in the context of CiE 2016 is of course that Barry was the driving force of Computability in Europe and without him, our association would not exist. This text is focused on Barry in relation to the Association Computability in Europe, it is based on a short obituary by Benedikt Löwe and Dag Normann [1], and borrows from it with due permission of the authors.

Barry retired from the office of President of the Association CiE in summer 2015, and had the chance to close the association AGM in Bucharest in July 2015 with a speech reminiscing about the history of the association. Barry was very fond of telling the ironic tale of how our association comprising more than a thousand members grew out of a rejected application for European funding.

In order to discuss the negative feedback of the referees, it was decided to have a conference in Amsterdam, which became the first CiE conference. Barry's vision and guidance pushed us along the way, to subsequent CiE conferences and finally to the
formal formation of this association in 2008. In 2007 and 2012, he personally co-chaired the Program Committees of the CiE conferences in Siena and Cambridge; the fact that these two events were the two largest CiE conferences to date is a testament to Barry's infectious enthusiasm and inclusive attitude. Barry also realized the potential of the Turing Centenary and made sure that the 100th birthday of Alan Turing was appropriately celebrated during the Alan Turing Year, not just in the UK, but all across the globe; at Turing's alma mater in Cambridge, Barry was one of the organizers of a six-month Turing-related research program at the Isaac Newton Institute for Mathematical Sciences culminating on Turing's 100th birthday, June 23, 2012, on the lawn in front of King's College. In the years after the centenary, Barry renamed the Alan Turing Year to Alan Turing Years. As the media attention to Alan Turing grew, partly due to the Academy-award winning movie The Imitation Game, Barry became one of Alan Turing's spokespeople on Twitter and in opinion pieces for The Guardian.

A comprehensive account of Barry's impact on Computability in Europe by Benedikt Löwe [2] has been published in the association's journal Computability.

Barry was very influential in shaping our thinking about computability in much broader, interdisciplinary terms, which was key to the success of the movement Computability in Europe. His vision will continue to live in us; his stimulating remarks and kindness will be very much missed.

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[1] Löwe, B., Normann, D.: Barry Cooper (1943-2015). Obituary published on the website of the Association CiE, 28 October 2015
[2] Löwe, B.: Barry Cooper (1943-2015): The Engine of Computability in Europe. Computability 5(1), 3-11 (2016)

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## Invited Papers

# Verifying Systems of Resource-Bounded Agents 

Natasha Alechina ${ }^{(\boxtimes)}$ and Brian Logan<br>University of Nottingham, Nottingham, UK<br>\{nza,bsl\}@cs.nott.ac.uk


#### Abstract

Approaches to the verification of multi-agent systems are typically based on games or transition systems defined in terms of states and actions. However such approaches often ignore a key aspect of multiagent systems, namely that the agents' actions require (and sometimes produce) resources. We briefly survey previous work on the verification of multi-agent systems that takes resources into account, and outline some key challenges for future work.


## 1 Verifying Autonomous Systems

A multi-agent system (MAS) is a system that is composed of multiple interacting agents. An agent is an autonomous entity that has the ability to collect information, reason about it, and perform actions based on it in pursuit of its own goals or on behalf of others. Examples of agents are controllers for satellites, non-driver transport systems such as UAVs, health care systems, and even nodes in sensor networks.

Multi-agent systems are ubiquitous. Many distributed software and hardware systems can be naturally modelled as multi-agent systems. Such systems are by the nature of their components extremely complex, and the interaction between components and their environment can lead to undesired behaviours that are difficult to predict in advance. With the increasing use of autonomous agents in safety critical systems, there is a growing need to verify that their behaviour conforms to the desired system specification, and over the last decade verification of multi-agent systems has become a thriving research area [24].

A key approach to the verification of MAS is model checking. Model checking involves checking whether a model of the system satisfies a temporal logic formula corresponding to some aspect of the system specification. Model checking has the advantage that it is a fully automated technique, which facilitates its use in the MAS development process. ${ }^{1}$ A wide range of approaches to model-checking MAS have been proposed in the literature, ranging from the adaptation of standard model-checking tools, e.g., $[12,13]$ to the development of special-purpose model checkers for multi-agent systems, e.g., [22,27].

[^0]
## 2 Resource-Bounded Agents

In many multi-agent systems, agents are resource-bounded, in the sense that they require resources in order to act. Actions require time to complete and typically require additional resources depending on the application domain, for example energy or money. For many applications, the availability or otherwise of resources is critical to the properties we want to verify: a multi-agent system will have very different behaviours depending on the resource endowment of the agents that comprise it. For example, an agent with insufficient energy may be unable to complete a task in the time assumed by a team plan, if it has to recharge its battery before performing the task.

However, with a few exceptions which we discuss below, previous work on verification of MAS abstracts away from the fact that many multi-agent systems consist of agents that need resources to operate and that those resources are limited. In particular, current state-of-the-art verification techniques and tools for MAS are unable to verify system properties that depend on the resource production and consumption of the agents comprising the MAS.

In this paper we survey recent work in the emerging field of verification of resource-bounded agents, and highlight a number of challenges that must be overcome to allow practical verification of resource-bounded MAS. We argue that recent work on the complexity of model-checking for logics of strategic ability with resources offers the possibility of significant progress in the field, new verification approaches and tools, and the ability to verify the properties of a large, important class of autonomous system that were previously out of reach.

## 3 Model-Checking with Resources

In this section we give a brief introduction to model-checking multi-agent systems and explain how standard model checking approaches can be extended with resources.

In a model-checking approach to the verification of multi-agent systems, a MAS is represented by a finite state transition system. ${ }^{2}$ A state transition system consists of a set of states and transitions between them. Intuitively, each state of a MAS corresponds to a tuple of states of the agents and of the environment, and each transition corresponds to actions performed by the agents. Each state is labelled with atomic propositions that are true in that state. A standard assumption is that each state in the system has at least one outgoing transition (if a state is a deadlock state in the original MAS, we can model this by adding a transition to itself by some null action and labelling it with a 'deadlock' proposition). Properties of the system to be verified are expressed in an appropriate temporal logic $L$. The model-checking problem for $L$ is, given a state transition system $M$ (and possibly a state $s$ ) and an $L$ formula $\phi$, check whether $\phi$ is true in $M$ (at state $s$ ).

[^1]For multi-agent systems, a temporal logic of particular interest is Alternating Time Temporal Logic (ATL) [9]. ATL generalises other temporal logics such as Computation Tree Logic (CTL) [19] (which can be seen as a one-agent ATL) by introducing notions of strategic ability. ATL is interpreted over concurrent game structures (transition systems where edges correspond to a tuple of actions performed simultaneously by all the agents, see the example below). The language of ATL contains atomic propositions, boolean connectives $\neg, \wedge$, etc. and modalities $\langle\langle A\rangle\rangle \bigcirc,\langle\langle A\rangle\rangle \square$ and $\langle\langle A\rangle\rangle \mathcal{U}$ for each subset (or coalition in ATL terms) $A$ of the set of all agents, which express the strategic ability of the coalition $A$. $\langle\langle A\rangle\rangle \bigcirc \phi$ means that the coalition of agents $A$ has a choice of actions such that, regardless of what the other agents in the system do, $\phi$ will hold in the next state. $\langle\langle A\rangle\rangle \square \phi$ means that coalition $A$ has a strategy to keep $\phi$ true forever, regardless of what the other agents do. A strategy is a choice of actions which either only depends on the current state (memoryless strategy) or on the finite history of the current state (perfect recall strategy). Finally, $\langle\langle A\rangle\rangle \phi \mathcal{U} \psi$ means that $A$ has a strategy to ensure that after finitely many steps $\psi$ holds, and in all the states before that, $\phi$ holds. The model-checking problem for ATL can be solved in time polynomial in the size of the transition system and the property [9], and there exist model-checking tools for ATL, for example, MOCHA [10] and MCMAS [27].

Example. Fig. 1 illustrates a simple ATL model of a system with two agents, 1 and 2 , and actions $\alpha, \beta, \gamma$ and $i d l e$. Action tuples on the edges show the actions of each agent, for example, in the transition from state $s_{I}$ to $s$, agent 1 performs action $\alpha$ and agent 2 performs $i d l e$. In this system, in state $s_{I}$, agent 1 has a (memoryless) strategy to enforce that $p$ holds eventually in the future no matter what agent 2 does, which can be expressed in ATL as $\langle\langle\{1\}\rangle\rangle \mathcal{U} p$. Similarly, in $s_{I}$ agent 1 has a memoryless strategy to keep $\neg p$ true forever, so $\langle\langle\{1\}\rangle\rangle \square \neg p$ holds in $s_{I}$.

### 3.1 Adding Resources

In order to model multi-agent systems where agents' actions produce and consume resources, it is necessary to modify the approach above in two ways. One is to add resource annotations to the actions in the transition system: for each individual action and each resource type, we need to specify how many units of this resource type the action produces or consumes. For example, suppose that there are two resource types, $r_{1}$ and $r_{2}$ (e.g., energy and money). Then we can specify that action $\alpha$ in Fig. 1 produces two units of $r_{1}$ and consumes one unit of $r_{2}$, action $\beta$ consumes one unit of $r_{1}$ and produces one unit of $r_{2}$, action $\gamma$ consumes five units of $r_{1}$, and action idle does not produce or consume any resources.

The second modification is to extend the temporal logic so that we can express properties related to resources. For example, we may want to express a property that a group of agents $A$ can eventually reach a state satisfying $\phi$ or can maintain the truth of $\psi$ forever, provided that they have available $n_{1}$ units
of resource type $r_{1}$ and $n_{2}$ units of resource type $r_{2}$. Such statements about coalitional ability under resource bounds can be expressed in an extension of ATL where coalitional modalities are annotated with a resource bound on the strategies available to the coalition. We call logics where every action is associated with produced and consumed resources and the syntax reflects resource requirements on agents, resource logics.

To illustrate the properties resource logics allow us to express, consider the model in Fig. 1 with the production and consumption of resources by actions specified above. In this setting, we can verify if agent 1 can eventually enforce $p$ provided that it has one unit of $r_{2}$ in state $s_{I}$, or whether the coalition of agents $\{1,2\}$ can achieve $p$ under this resource bound by working together. There are surprisingly many different ways of measuring costs of strategies and deciding which actions are executable by the agents given the resources available to them, but under at least one possible semantics, the answer to the first question is no and to the second one yes, but the latter requires a perfect recall strategy (the two agents should loop between states $s_{I}$ and $s$ until they produce a sufficient amount of resource $r_{1}$, and then execute actions corresponding to the $\langle\gamma, i d l e\rangle$ transition from $s$ to $s^{\prime}$ ).


Fig. 1. State transition system.

Clearly, the model-checking problem for temporal logics is a special case of the model-checking problem for the corresponding resource logics. The question is, how much harder does the model-checking problem become when resources are added?

## 4 A Brief Survey of Resource Logics

In this section, we briefly review recent theoretical work on the development of resource logics. We focus on expressiveness and model-checking complexity, as these features determine the suitability of a particular logic for practical verification.

### 4.1 Consumption of Resources

Early work on resource logics considered only consumption of resources (no action produced resources), and initial results were encouraging.

One of the first logics capable of expressing resource requirements of agents was a version of Coalition Logic (CL) ${ }^{3}$, called Resource-Bounded Coalition Logic (RBCL), where actions only consume (and don't produce) resources. It was introduced in [1] with the primary motivation of modelling systems of resourcebounded reasoners, however the framework is sufficiently general to model any kind of action. The model-checking problem for this logic was shown to be decidable in [5] in polynomial time in the transition system and the property and exponential in the number of resource types.

A resource-bounded version of ATL, RB-ATL, where again actions only consume (and not produce) resources was introduced in [2]. It was also shown that the model-checking problem for this logic is decidable in time polynomial in the size of the transition system and exponential in the number of resource types. (For a single resource type, e.g., energy, the model-checking problem is no harder than for ATL.)

Practical work on model-checking standard computer science transition systems (not multi-agent systems) with resources also falls in the category of consumption-only systems, for example probabilistic model-checking of systems with numerical resources as done using PRISM model-checker [26] assumes costs monotonically increasing with time.

### 4.2 Adding Production

However, when resource production is considered in addition to consumption, the situation changes. In a separate strand of work, a range of different formalisms for reasoning about resources was introduced in $[14,16]$. In those formalisms, both consumption and production of resources was considered. In [15] it was shown that the problem of halting on empty input for two-counter automata [25] can be reduced to the model-checking problem for several of their resource logics. Since the halting problem for two-counter automata is undecidable, the modelchecking problem for a variety of resource logic with production of resources is undecidable. The reduction uses two resource types (to represent the values of the two counters) and either one or two agents depending on the version of the logic (whether the agents have perfect recall, whether the formula talking about coalition $A$ can also specify resource availability for remaining agents, and whether nested operators 'remember' initial allocation of resources or can be evaluated independently of such initial allocation).

The only decidable cases considered in [14] are an extension of CTL with resources (essentially one-agent ATL) and a version where on every path only a fixed finite amount of resources can be produced. In [14], the models satisfying this property are called bounded, and it is pointed out that RBCL and RBATL are logics over a special kind of bounded models (where no resources are produced at all). Other decidability results for bounded resource logics have also been reported in the literature. For example, [20] define a decidable logic, PRB-ATL (Priced Resource-Bounded ATL), where the total amount of resources

[^2]in the system has a fixed bound. The model-checking algorithm for PRB-ATL requires time polynomial in the size of the model and exponential in the number of resource types and the resource bound on the system. In [21] an EXPTIME lower bound in the number of resource types for the PRB-ATL model-checking problem is shown.

A general logic over systems with numerical constraints called QATL* was introduced in [17]. In that paper, more undecidability results for the modelchecking problem of QATL* and its fragments were shown. For example, QATL (Quantitative ATL) is undecidable even if no nestings of cooperation modalities is allowed. The main proposals for restoring decidability to the model-checking problem for QATL in [17] are removing negative payoffs (similar to removing resource production) and also introducing memoryless strategies. Shared resources were considered in [18]; most of the cases considered there have undecidable model-checking (apart from the case of a single shared resource, which has decidable model-checking).

This brief survey of work to date suggest that the main approach until recently to dealing with both resource production and consumption was to bound the amount of produced resources globally in the model. For some systems of resource-bounded agents, this is a reasonable restriction. For example, agents that need energy to function and are able to charge their battery, can never 'produce' more energy than the capacity of their battery. This is a typical bounded system. However, in some cases, although every single application of the agent's actions produces a fixed amount of some resource, repeating this action arbitrarily often will produce arbitrarily large amounts of the resource. This may apply to energy stored in unbounded storage, or to money, or many other natural situations. Recent work suggests that verification of such systems may still be possible.

## 5 Decidable Unbounded Production

In [6] a version of ATL, RB $\pm$ ATL, was introduced where actions both produce and consume resources. The models of the logic do not impose bounds on the overall production of resources, and the agents have perfect recall. The syntax of $\mathrm{RB} \pm$ ATL is very similar to that of ATL, but coalition modalities have superscripts which represent resource allocation to agents. Instead of stating the existence of some strategy, they state the existence of a strategy such that every computation generated by following this strategy consumes at most the given amount of resources. Coming back to the example, the property that agent 1 can eventually enforce $p$ provided that it has one unit of $r_{2}$ can be expressed as $\left\langle\left\langle\{1\}^{(0,1)}\right\rangle\right\rangle \top \mathcal{U} p^{4}$ The model-checking problem for $\mathrm{RB} \pm \mathrm{ATL}$ is decidable ( $\mathrm{RB} \pm$ ATL is very similar to one of the resource logics introduced

[^3]in [14] for which the decidability of the model-checking problem was left open). The existence of a decidable resource logic with unbounded production was surprising, as it was the first indication that it is possible to automatically verify properties of this important class of resource-bounded multi-agent system.

However, although this result is encouraging, we are not yet at the point of practical verification of such systems. In [6] the lower bound on the complexity of the model-checking problem for $\mathrm{RB} \pm$ ATL is shown to be EXPSPACE. The proof of EXPSPACE-hardness is by reduction of the reachability problem for Petri Nets to the model-checking problem for $\mathrm{RB} \pm$ ATL. Although the Petri Net reachability problem is decidable, the upper bound on its complexity is still unknown; similarly we do not know the upper bound on the RB $\pm$ ATL complexity. The complexity of the model-checking problem RB $\pm$ ATL is thus much higher than that for ATL without resources and the consumption-only resource logics surveyed above. This high complexity makes it difficult to develop practical verification approaches. The only exception is $1-\mathrm{RB} \pm \mathrm{ATL}, \mathrm{RB} \pm \mathrm{ATL}$ with a single resource type, where the complexity is PSPACE.

In [3], a new syntactic fragment FRAL of resource logic RAL with decidable model-checking has been identified. It restricts the occurrences of coalitional modalities on the left of the Until formulas; on the other hand, it allows nested modalities to refer to resource allocation at the time of evaluation, rather than always consider a fresh resource allocation, as in $\mathrm{RB} \pm$ ATL. More precisely, a formula $\left\langle\left\langle A^{b}\right\rangle\right\rangle \phi \mathcal{U}\left\langle\left\langle A^{\downarrow}\right\rangle\right\rangle \psi_{1} \mathcal{U} \psi_{2}$ says that given resource allocation $b$, coalition $A$ can always reach a state (maintaining $\phi$ ) where with the remaining resources, it can reach $\psi_{2}$ while maintaining $\psi_{1}$. The complexity of model-checking for this fragment is also open, and is also likely to be high.

Although model-checking of ATL with perfect recall and uniform strategies is undecidable, replacing uniformity with a weaker notion, for example defining it using distributed knowledge, is decidable [23]. Similar results hold for $R B \pm$ ATL with syntactic epistemic knowledge and weaker notions of uniformity, $\mathrm{RB} \pm$ ATSEL [4].

Below is a summary of resource logics with decidable model-checking problem. In all of them, the semantics assume that in every state each agent has an available action to do nothing, which produces and consumes no resources (see Table 1).

## 6 Future Challenges

The $\mathrm{RB} \pm$ ATL results offer the possibility of significant progress in the verification of resource-bounded multi-agent systems. However many challenges remain for future research. Below we list three of the most important.

Understanding Sources of Undecidability. Developing a better understanding of the sources of decidability and undecidability (beyond boundedness) will be critical to future progress. As observed in [14], subtle differences in truth conditions for resource logics result in the difference between decidability and

Table 1. Resource logics with decidable model-checking problem

| Logic | Resource production | Complexity of model-checking |
| :--- | :--- | :--- |
| RBCL | no | EXPTIME (PTIME in model) |
| RB-ATL | no | EXPTIME (PTIME in model) |
| PRB-ATL | bounded | EXPTIME |
| RB $\pm$ ATL | yes | EXPSPACE-hard |
| 1-RB $\pm$ ATL | yes | PSPACE |
| FRAL | yes | $?$ |
| RB $\pm$ ATSEL | yes | EXPSPACE-hard |

undecidability of the model-checking problem. Preliminary work in this direction is reported in [3].

Lower Complexity. It is useful to discover sources of undecidability and how to construct expressive logics for which the model-checking problem is decidable. However, it is even more important to be able to develop logics, or fragments of existing logics such as $\mathrm{RB} \pm$ ATL, that are sufficiently expressive for practical problems, but where the model-checking problem has tractable complexity (ideally polynomial in the size of the transition system, as in the case of bounded production logics). Only then would we be able to implement practical modelchecking tools for systems of resource-bounded agents.

Practical Tools. Although model checking algorithms have been proposed for several of the logics surveyed, work on implementation is only beginning. We aim to develop practical model-checking tools for verifying resource-bounded MAS by extending the MCMAS model checker [27] to allow the modelling of multiagent systems in which agents can both consume and produce resources. Work on symbolic encoding of RB-ATL model-checking is reported in [8] and work on symbolic encoding of $\mathrm{RB} \pm$ ATL model-checking are reported in [7].

Addressing these challenges will allow practical model-checking of resource logics and constitute a major break-through in multi-agent system verification.

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# Effective $S$-adic Symbolic Dynamical Systems 

Valérie Berthé ${ }^{1(\boxtimes)}$, Thomas Fernique ${ }^{2}$, and Mathieu Sablik ${ }^{3}$<br>${ }^{1}$ IRIF, CNRS UMR 8243, Univ. Paris Diderot, Paris, France<br>berthe@liafa.univ-paris-diderot.fr<br>${ }^{2}$ LIPN, CNRS UMR 7030, Univ. Paris 13, Villetaneuse, France fernique@lipn.fr<br>${ }^{3}$ I2M UMR 7373, Aix Marseille Univ., Marseille, France<br>mathieu.sablik@univ-amu.fr


#### Abstract

We focus in this survey on effectiveness issues for $S$-adic subshifts and tilings. An $S$-adic subshift or tiling space is a dynamical system obtained by iterating an infinite composition of substitutions, where a substitution is a rule that replaces a letter by a word (that might be multi-dimensional), or a tile by a finite union of tiles. Several notions of effectiveness exist concerning $S$-adic subshifts and tiling spaces, such as the computability of the sequence of iterated substitutions, or the effectiveness of the language. We compare these notions and discuss effectiveness issues concerning classical properties of the associated subshifts and tiling spaces, such as the computability of shift-invariant measures and the existence of local rules (soficity). We also focus on planar tilings.


Keywords: Symbolic dynamics • Adic map • Substitution • $S$-adic system $\cdot$ Planar tiling $\cdot$ Local rules $\cdot$ Sofic subshift $\cdot$ Subshift of finite type $\cdot$ Computable invariant measure $\cdot$ Effective language

## 1 Introduction

Decidability in symbolic dynamics and ergodic theory has already a long history. Let us quote as an illustration the undecidability of the emptiness problem (the domino problem) for multi-dimensional subshifts of finite type (SFT) [8, 40], or else the connections between effective ergodic theory, computable analysis and effective randomness (see for instance $[14,35,44]$ ). Computability is a notion that has also appeared as a major understanding tool in the study of multidimensional subshifts of finite type with the breakthrough characterization by M. Hochman and T. Meyerovitch of the entropies of multi-dimensional subshifts of finite type as the non-negative right recursively enumerable numbers [34] (see also [30] in the one-dimensional case). Let us mention also the realization of effective subshifts (with factor and projective subaction operation) from higher-dimensional subshifts of finite type $[3,19,33]$. It is now clear that

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sofic and effective subshifts are closely related, in particular for substitutive subshifts and tilings. Indeed, contrarily to the one-dimensional case, substitution subshifts are known to have (colored) local rules (they are SFT or sofic) in the higher-dimensional framework $[26,31,38]$.

We focus here on effectiveness issues for $S$-adic subshifts (and tilings). An $S$-adic expansion is a way to represent (or to generate) words (one-dimensional and multi-dimensional ones), or tilings, by composing infinitely many substitutions. A (word) substitution is a morphism of the free monoid: it replaces letters by words. Substitutions can also be defined in the higher-dimensional framework: they replace letters by multi-dimensional patterns, and act on multi-dimensional words (configurations) in $\mathbb{Z}^{d}$; more generally, substitutions can also generate and act on tilings, by replacing a tile by a finite union of tiles. An infinite word $u$ (or a $d$-dimensional configuration, or a tiling) admits an $S$-adic expansion if

$$
u=\lim _{n \rightarrow \infty} \sigma_{0} \sigma_{1} \cdots \sigma_{n-1}\left(a_{n}\right),
$$

where $\left(\sigma_{n}\right)_{n \in \mathbb{N}}$ is a sequence of substitutions, and $\left(a_{n}\right)_{n \in \mathbb{N}}$ a sequence of letters. For more on substitutions, see e.g. [28], and for more on $S$-adic words and tilings, see $[9,29]$. There is a deep parallelism between subshifts associated with such expansions (under natural assumptions like primitivity) and Bratteli-Vershik systems endowed with adic transformations, hence the terminology 'adic', with the letter $S$ referring to 'substitution'. This connection between adic models and substitutions has been widely investigated; see e.g. [24], or [10] and the references therein. Recall also that any Cantor minimal system admits a Bratteli-Vershik representation [32], which illustrates the representation power of this notion.

Without any further assumption on the $S$-adic representation, every infinite word admits an $S$-adic expansion (according to Cassaigne's construction, see e.g. [9, Remark 3]). One thus needs to introduce suitable assumptions on these $S$-adic representations in order to find a good balance between the expressive power of such representations and the information provided by their existence. Let us illustrate this with [2] where it is proved that multi-dimensional $S$-adic subshifts, obtained by applying an effective sequence of substitutions chosen among a finite set of substitutions, are sofic subshifts.

Basic notions and definitions on substitutions and $S$-adic subshifts and tilings are recalled in Sect.2. We discuss some decidability results in the onedimensional setting for substitutive words in Sect. 3. Section 4 focuses on effectiveness for $S$-adic subshifts. Lastly, multi-dimensional Sturmian words and planar tilings are considered in Sect. 5.

## 2 Definitions

### 2.1 Subshifts

Let $\mathcal{A}$ be finite alphabet and $d \geq 1$. A configuration $u$ is an element of $\mathcal{A}^{\mathbb{Z}^{d}}$. A pattern $p$ is an element of $\mathcal{A}^{D}$, where $D \subset \mathbb{Z}^{d}$ is a finite set, called its support.

Denote $\mathcal{A}^{*}$ the set of patterns. A translate of the pattern $p$ by $\mathbf{m} \in \mathbb{Z}^{d}$ is denoted $p+\mathbf{m}$ and has $D+\mathbf{m}$ for support. A pattern $p \in \mathcal{A}^{D}$ is a factor of a configuration $u=\left(u_{\mathbf{n}}\right)_{\mathbf{n} \in \mathbb{Z}^{d}}$ if there exists $\mathbf{m} \in \mathbb{Z}^{d}$ such that the restriction of $u$ to $D+\mathbf{m}$ coincides with $p+\mathbf{m}$. The set of factors (up to translation) of $u$ is called its language.

The set $\mathcal{A}^{\mathbb{Z}^{d}}$ endowed with the product topology is a compact metric space. A $d$-dimensional subshift $X \subset \mathcal{A}^{\mathbb{Z}^{d}}$ is a closed and shift-invariant set of configurations in $\mathcal{A}^{\mathbb{Z}^{d}}$, where the shifts $\sigma_{\mathbf{m}}$ with $\mathbf{m} \in \mathbb{Z}^{d}$ are defined as $\sigma_{\mathbf{m}}: \mathcal{A}^{\mathbb{Z}^{d}} \rightarrow \mathcal{A}^{\mathbb{Z}^{d}}$, $\left(u_{\mathbf{n}}\right)_{\mathbf{n} \in \mathbb{Z}^{d}} \mapsto\left(u_{\mathbf{n}+\mathbf{m}}\right)_{n \in \mathbb{Z}^{d}}$. The shifts provide a natural action of $\mathbb{Z}^{d}$.

A subshift $X$ can be defined by providing its language, that is, the set of patterns (up to translation) that occur in configurations in $X$. It can be defined equivalently by providing the set of forbidden patterns. Subshifts of finite type (also called SFT) are the subshifts such that the set of their forbidden patterns is finite. Sofic subshifts are images of SFT under a factor map, where a factor map $\pi: X \rightarrow Y$ between two subshifts $X$ and $Y$ is a continuous, surjective map such that $\pi \circ \sigma_{\mathbf{m}}=\sigma_{\mathbf{m}} \circ \pi$, for all $\mathbf{m} \in \mathbb{Z}^{d}$.

Definition 1 (Computable subshift). A subshift is said to be

- $\Pi_{1}$-computable or effective if its language is co-recursively enumerable;
- $\Sigma_{1}$-computable if its language is recursively enumerable;
- $\Delta_{1}$-computable or decidable if its language is recursive.

A subshift is said to be linearly recurrent if there exists $C>0$ such that every pattern whose support is a translate of $[-C n, C n]^{d}$ contains every factor whose support is a translate of $[-n, n]^{d}$. The frequency $f(p)$ of a pattern in a $d$-dimensional configuration $u$ is defined as $\limsup _{n}\left|x_{n}\right|_{p} /(2 n+1)^{d}$, where $x_{n}$ is the restriction of $u$ to $[-n, n]^{d}$, and $\left|x_{n}\right|_{p}$ stands for the number of occurrences of $p$ in $x_{n}$. If the lim sup is in fact a limit, then the frequency is said to exist. A subshift is said to be uniquely ergodic if it admits a unique shift-invariant measure; in this case, pattern frequencies do exist. A subshift is said to be minimal if every non-empty closed shift-invariant subset is equal to the whole set. A minimal and uniquely ergodic subshift is said strictly ergodic. Any pattern which appears in a strictly ergodic subshift has a positive frequency. For more on multidimensional subshifts, see e.g. [11, Chap. 8, 9].

### 2.2 Substitutions and $S$-adic Subshifts

A substitution $s$ over the alphabet $\mathcal{A}$ is a map $s: \mathcal{A} \longrightarrow \mathcal{A}^{*}$. Let $\mathscr{S}$ be a finite set of substitutions; we want to define how a pattern of substitutions $\mathbf{s} \in \mathscr{S}^{D}$ acts on a pattern $p \in \mathcal{A}^{D}$, with $D \subset \mathbb{Z}^{d}$ finite. This general definition allows us to apply simultaneously different substitutions; we are in the non-deterministic case of [38]. We thus introduce concatenation rules which specify how the respective images of two adjacent tiles must be glued. A pattern of substitutions $\mathbf{s} \in \mathscr{S}^{D}$ is said to be compatible with a pattern $p \in \mathcal{A}^{D}$ (made of cells) if it is consistent (the image of a cell does not depend on the sequence of concatenation rules that
are used, patterns have a unique image) and non-overlapping (the images of two cells do not overlap).

When $\mathbf{s}$ and $p$ are compatible, the (unique) image of $p$ by $\mathbf{s}$ is denoted as $\mathbf{s}(p)$. If all letters of $\mathbf{s}$ are equal to the same substitution $s \in \mathscr{S}$, the $\mathscr{S}$-pattern is said to be s-constant. This corresponds to the classical case of the action of one substitution (the deterministic case in [38]). An $\mathscr{S}$-super-tile of order $n$ corresponds to $n$ iterations of (compatible) patterns of substitutions applied to a letter. We define the $\mathscr{S}$-adic subshift $X_{\mathscr{S}}$ as the set of the configurations for which every pattern appears in an $\mathscr{S}$-super-tile. Here we can compose all substitutions in $\mathscr{S}$. This notion plays a role for the description of planar tilings introduced in Sect. 5.

We now introduce the usual $S$-adic setting by applying only constant patterns of substitutions. We take a sequence of substitutions $S=\left(s_{n}\right)_{n \in \mathbb{N}} \in \mathscr{S}^{\mathbb{N}}$; the shift acting on $S$ is denoted as $\sigma\left(\sigma(S)=\left(s_{n+1}\right)_{n \in \mathbb{N}} \in \mathscr{S}^{\mathbb{N}}\right)$. The $S$-super-tile of order 0 and type $a \in \mathcal{A}$ is defined as the letter $a$, whereas the $S$-super-tile of order $n+1$ and type $a$ is the image of the $\sigma(S)$-super-tile of order $n$ and type $a$ by a $s_{0}$-constant $\mathscr{S}$-pattern. A super-tile of order $n$ can thus be defined by a word of size $n$ in $\mathscr{S}^{*}$ together with a letter. The sequence $S$ is said to be a directive sequence. We then define the $S$-adic subshift $X_{S}$ as the set of the configurations for which every pattern appears in an $S$-super-tile. For a closed subset $\mathbf{S} \subset \mathscr{S}^{\mathbb{N}}$, we also define the $\mathbf{S}$-adic subshift $X_{\mathbf{S}}=\bigcup_{S \in \mathbf{S}} X_{S}$.

For $d=1$, a natural way to define concatenation rules is to consider that the image of two consecutive letters is obtained as the concatenation of the two image words. Thus a substitution can be viewed as a non-erasing endomorphism of the free monoid $\mathcal{A}^{*}$. For example the Fibonacci substitution on the alphabet $\{a, b\}$ is defined by $\sigma: a \mapsto a b, b \mapsto a$. For $d=2$, if all the supports of the images by an element of $\mathscr{S}$ are rectangular, concatenation rules of two adjacent letters consist in the concatenation of the two image patterns as long as the two glued edges have the same size. Rectangular substitutions are considered e.g. in [38] (see also [16] for the notion of shape-symmetric rectangular substitutions).

It is possible to extend the notion of substitution to geometric tilings. A tiling of $\mathbb{R}^{d}$ is a collection of compact sets which cover topologically $\mathbb{R}^{d}$, that is, with the interiors of the tiles being pairwise disjoint. In general, a tile-substitution in $\mathbb{R}^{d}$ is given by a set of prototiles $T_{1}, \ldots, T_{m} \subset \mathbb{R}^{d}$, an expanding map and a rule how to dissect each expanded prototile into translated copies of some prototiles $T_{i}$. These geometric tiling substitutions are considered e.g. in [31]. It is also possible to define $S$-adic tilings in this context (see [29]).

There are further strategies for defining substitutions such as described in [28]. For instance, one can also use a global information; see the formalism introduced in [1] that allows the generation of multi-dimensional Sturmian words considered in Sect. 5; this formalism also provides concatenation rules [25].

## 3 Some Decisions Problems for Substitutions

In the one-dimensional case, numerous decidability results exist for fixed points of substitutions (D0L words), and their images by general morphisms (HD0L
words). More precisely, let $\mathcal{A}, \mathcal{B}$, be finite alphabets. We consider two morphisms $\sigma: \mathcal{A}^{*} \rightarrow \mathcal{A}^{*}, \phi: \mathcal{A}^{*} \rightarrow \mathcal{B}^{*}$; an infinite word of the form $\lim _{n} \sigma^{n}(u)$ (respectively $\phi\left(\lim _{n} \sigma^{n}(u)\right)$ ) is a D0L (respectively an HD0L word or morphic word), for $u$ finite word.

We focus here on some decision problems that can be solved using the notion of return word and derived sequence (see e.g. [20]). Let $\sigma$ be a primitive substitution. It generates a minimal subshift $X_{\sigma}$. A return word to a word $u$ of its language is a word $w$ of the language such that $u w$ admits exactly two occurrences of $u$, with the second occurrence of $u$ being a suffix of $u w$. One can recode sequences of the subshift via return words, obtaining derived sequence (see e.g. [20]). Note that even if analogous notions exist in the higher-dimensional case and for tilings [39], this is not sufficient to yield a direct generalization of the results described below.

The HD0L $\omega$-equivalence problem (which has been open for more than 30 years) is solved in [21] for primitive morphisms: it is decidable to know whether two HD0L words are equal (see the references in [21] for the D0L case). The decidability of the ultimate periodicity of HD0L infinite sequences has also been a long-standing problem: it is decidable to know whether an HD0L word is ultimately periodic. See [21] for the primitive case, and [22] for the general case. See also the references in [22] for the D0L case. This problem is closely related to the decidability of the ultimate periodicity of recognizable sets of integers in some abstract numeration systems [7]. It is also proved in [23] that the uniform recurrence of morphic sequences is decidable.

The particular case of constant-length substitutions (automatic sequences) has also been widely studied; see e.g. [17,42] where decision procedures are produced based on the connections between first-order logic and automata such as developed in [15] where the equivalence between being $p$-recognizable and $p$-definability is developed. For more references, see also the book [41].

## 4 Effectiveness for S-adic Subshifts and Local Rules

We discuss here several effectivity notions for $S$-adic subshifts concerning their directive sequences, pattern frequencies, or else their language. We also focus on the existence of local rules. We only consider here iterations by constant $\mathscr{S}$-patterns. We recall that $\mathscr{S}$ is finite.

A closed subset $\mathbf{S} \subset \mathscr{S}^{\mathbb{N}}$ is effectively closed if the set of (finite) words which do not appear as prefixes of elements of $\mathbf{S}$ is recursively enumerable (one enumerates forbidden prefixes). An effectively closed set is not necessarily a subshift.

A set of substitutions $\mathscr{S}$ has a good growing property if there are finitely many ways of gluing super-tiles, and if the size of the super-tiles of order $n$ grows with $n$ : there exists a finite set of patterns $\mathcal{P} \subset \mathcal{A}^{*}$ such if a pattern formed by a super-tile of order $n$ surrounded by super-tiles of order $n$ is in the language of $X_{\mathscr{S}}$, then it appears as the $n$-iteration by a constant $\mathscr{S}$-pattern of a pattern of $\mathcal{P}$, and, moreover, if for every ball of radius $R$, there exist $n \in \mathbb{N}$ such a
translate of this ball is contained in all the supports of super-tiles of order $n$. Clearly non-trivial rectangular substitutions or geometrical substitutions (such as defined in [31]) verify this property.

Proposition 1. Let $\mathbf{S} \subset \mathscr{S}^{\mathbb{N}}$ be a closed subset. If $X_{\mathbf{S}}$ is effective, then there exists an effective closed subset $\mathbf{S}^{\prime} \subset \mathscr{S}^{\mathbb{N}}$ such that $X_{\mathbf{S}}=X_{\mathbf{S}^{\prime}}$. The reciprocal is true if $\mathscr{S}$ has the good growing property.

Proof. Assume that $X_{\mathbf{S}}$ is effective. The complement of its language is recursively enumerable. Let $\mathbf{S}^{\prime}$ be the effective closed set such that a word in $\mathscr{S}^{*}$ is a forbidden prefix if the associated super-tiles are in the complement of the language of $X_{\mathbf{S}}$. Clearly $X_{\mathbf{S}^{\prime}}=X_{\mathbf{S}}$.

Conversely, consider the $\mathbf{S}$-adic subshift $X_{\mathbf{S}}$ where $\mathbf{S}$ is effectively closed and let $\mathcal{P}$ be the set of patterns given by the good growing property. Let $\mathcal{P}^{\prime}$ be the set of patterns in $\mathcal{P}$ that occur in $X_{\sigma^{n}(\mathbf{S})}$ for infinitely many $n$. A pattern $p$ is in the language of $X_{\mathbf{S}}$ if it appears in the image by an $n$-iteration of a pattern of $\mathcal{P}^{\prime}$, where $n$ is the first order where the support of $p$ is included in all super-tiles of order $n$. Since $\mathcal{P}^{\prime}$ is finite and the prefixes of $\mathbf{S}$ are co-recursively enumerable, the same holds for the language of $X_{\mathbf{S}}$.

Definition 2 (Computable frequencies and measure). Let $X$ be a subshift. $X$ is said to have computable frequencies if the frequencies of patterns exist and are uniformly computable. A shift-invariant measure is said to be computable if the measure of any cylinder is uniformly computable.

Remark 1. Computability of letter frequencies does not say much on the algorithmic complexity of a subshift: take a subshift $X \subset\{0,1\}^{\mathbb{Z}}$ and consider the subshift $Y$ obtained by applying to each configuration of $X$ the substitution $0 \mapsto 01,1 \mapsto 10$. The subshift $Y$ admits letter frequencies (they are both equal to $1 / 2$ ), and it has the same algorithmic complexity as $X$.

Proposition 2. Let $X$ be a subshift. If $X$ is effective and uniquely ergodic, then its invariant measure is computable and it is decidable. If $X$ is minimal and its frequencies are computable, then its language is recursively enumerable. If $X$ is minimal and effective, then it is decidable.

Proof. Let $X$ be a $d$-dimensional subshift. We assume $X$ effective and uniquely ergodic. Let us prove that the frequency of any pattern is computable. We use the following algorithm that takes as an argument the parameter $e$ that stands for the precision. We consider a finite pattern $p$. At step $n$, one produces all 'square' patterns of size $n$ with support being a translate of $[-n, n]^{d}$ that do not contain the $n$ first forbidden patterns (they do not need to belong to the language of $X$ ). For each of these square patterns of size $n$, one computes the number of occurrences of $p$ in it, divided by $(2 n+1)^{d}$. We continue until these quantities belong to an interval of length $e$. This algorithm then stops, and taking an element of the interval provides an approximation of the frequency of $p$ up to precision $e$. Indeed, the square patterns of size $n$ contain the square patterns of size $n$ of $X$. It remains to prove that the algorithm stops. Suppose it does not,
then, for all $n$, one can find two patterns of size $n, x_{n}$ and $x_{n}^{\prime}$, that do not contain the $n$ first forbidden patterns and such that $\left|\left|x_{n}\right|_{p} /(2 n+1)^{d}-\left|x_{n}^{\prime}\right|_{p} /(2 n+1)^{d}\right|>e$. By compactness, we can extract two configurations $x$ and $x^{\prime}$ that do not contain forbidden patterns (they thus belong to the subshift $X$ ) such that the frequency of $p$ in $x$ is distinct from the frequency of $p$ in $x^{\prime}$. This contradicts the unique ergodicity of $X$.

We now assume $X$ minimal with computable pattern frequencies (frequencies are positive). One can decide whether the frequency of a pattern is larger than a given value. This thus implies that the language is recursively enumerable.

Assume that $X$ is minimal and effective. Consider the square patterns of size $n$ that do not contain the $n$ first forbidden patterns. If a pattern belongs to all these patterns for some $n$, then it belongs to the language. Otherwise, consider size $n+1$. The algorithm stops if a pattern is in the language by minimality.

Corollary 1. Let $X_{S}$ be a strictly ergodic $S$-adic subshift defined with respect to a directive sequence $S \in \mathscr{S}^{\mathbb{N}}$ such that $\mathscr{S}$ satisfies the good growing property. The following conditions are equivalent:

1. There exists a computable sequence $S^{\prime}$ such that $X_{S}=X_{S^{\prime}}$;
2. The unique invariant measure of $X_{S}$ is computable;
3. The subshift $X_{S}$ is decidable.

Proof. We first assume (1). By Proposition 1, $X_{S}$ is effective and Proposition 2 yields that its unique measure is computable and that $X_{S}$ is decidable.

We now assume (2). Let $d$ stand for the cardinality of the alphabet of the substitutions in $\mathscr{S}$. The letter frequency vector is in the cone defined by the product of the incidence matrices of the directive sequence. The incidence matrix of a substitution $s$ is a square matrix whose entry of index $(i, j)$ counts the number occurrences of the letter $i$ in $s(j)$. Let $M_{n}$ stand for the incidence matrix of the substitution $s_{n}$. The letter frequency belongs to the cone $\bigcap_{n} M_{1} \cdots M_{n} \mathbb{R}_{+}^{d}$, which is one-dimensional by unique ergodicity. Given a precision $e$, one can compute $n$ such that the columns of $M_{1} \cdots M_{n}$ are expected to be at a distance less than $e$ from the letter frequency vector. We fix a cylinder around the direction provided by the letter frequency vector with precision $e$. Now we test finite products of $n$ substitutions in $\mathscr{S}$. We consider the cone obtained by taking the product of the incidence matrices, and check whether it intersects the cylinder. If it does not intersect the cylinder, one gets a forbidden product of substitutions, which proves that $\{S\}$ is effectively closed.

It remains to prove that (3) implies (1). There exists a closed effective set such that $X_{S}=X_{\mathbf{S}}$, by Proposition 1. For every $S^{\prime} \in \mathbf{S}$, one has $X_{S^{\prime}}=X_{S}$, by minimality. We then exhibit a computable $S^{\prime}$ in $\mathbf{S}$ as follows: for any $n$, take the first prefix for the lexicographic order among the prefixes of elements in $\mathbf{S}$ such that $s_{0} s_{1} \cdots s_{n}(a)$ is in the language of $X_{S}$.

Existence of Local Rules. A natural question in tiling theory is to find local rules which only produce aperiodic tilings. The first examples of aperiodic subshifts of finite type were based on hierarchical structures $[8,40]$ : substitutive
structures are known to be able to force aperiodicity. Note that a non-trivial substitutive subshift cannot be sofic in dimension 1: it has zero topological entropy whereas non-trivial sofic subshifts have positive entropy. In dimension $d \geq 2$, under natural assumptions, it is known for different types of substitutions that substitutive tilings can be enforced with (colored) local rules. The ideas is always to force a hierarchical structure, as in Robison's tiling, where each change of level is marked by the type of the super-tile of this level, and the rule used is transmitted for super-tiles of lower order. For rectangular substitutions, the result is proved in [38] (with the result being more general since the substitutions are non-deterministic). The case of geometrical substitutions is handled in [31] and the result is also true in a more combinatorial way [26].

In the case of rectangular substitutions it is shown in [2] that the $\mathbf{S}$-adic subshift $X_{\mathbf{S}}$ is sofic if and only if it can be defined by a set of directive sequences $\mathbf{S}$ which is effectively closed. A similar result for more general substitutions is expected; the difficulty relies in the ability to exhibit a rectangular grid to use the simulation (see $[3,19]$ ) of a one-dimensional effective subshift by a twodimensional sofic subshift. One can also ask whether linearly recurrent effective subshifts are sofic. Note that such a statement cannot hold for computability reasons (there are uncountably many linearly recurrent subshifts) without any effectivity assumption. Note also that in the one-dimensional case, linearly recurrent subshifts are primitive $S$-adic [20].

## 5 An Application: Planar Tilings

As an example of $S$-adic configurations, we consider multi-dimensional Sturmian words. The associated tilings belong to the more general class of planar tilings. A (canonical) planar tiling is an approximation of an affine $d$-plane $E$ in $\mathbb{R}^{n}$, via the cut-and-project method (see e.g. [4]). Such a tiling can be lifted into the tube $E+[0, t]^{n}$ : the space $E$ is called the slope and the smallest possible $t$ the thickness. Planar tilings are closely related to discrete planes in discrete geometry and provide models of quasicrystals. The case $t=1$ and $d=n-1$ corresponds to the multi-dimensional Sturmian case. In terms of configurations, a multidimensional Sturmian word is defined as the coding of a $\mathbb{Z}^{d}$-action by $d$ rotations $R_{\alpha_{i}}: \mathbb{R} / \mathbb{Z} \rightarrow \mathbb{R} / \mathbb{Z}, x \mapsto x+\alpha_{i}(1 \leq i \leq d)$, where the $\alpha_{i}$ are positive real numbers. We assume $1, \alpha_{1}, \cdots, \alpha_{d}$ rationally independent and $\sum_{i} \alpha_{i}<1$. A multidimensional Sturmian word $u \in\{1,2, \cdots, d+1\}^{\mathbb{Z}^{d}}$ is defined as follows: there exists $\rho$, a partition of $\mathbb{R} / \mathbb{Z}$ into $d+1$ semi-open intervals, $d$ of lengths $\alpha_{i}$, and one of length $1-\sum \alpha_{i}$, such that $u_{\mathbf{n}}=i$ if and only if $n_{1} \alpha_{1}+\cdots+n_{d} \alpha_{d}+\rho \in I_{i}$ [12]. The cut-and-project framework is larger than the $S$-adic framework but multidimensional Sturmian words are $S$-adic [13] (via the formalism of [1]).

The study of the connections between the existence of local rules for a planar tiling and the parameters of its slope started with [18, 36, 37, 43]. In particular, it was proven in [36] that a slope enforced by undecorated local rules is necessarily algebraic (this is however not sufficient, see e.g. [5,6]). However, computability comes into play when the tiles can be decorated. Decorations indeed allow the
transfer of information through the tiling, and this was used in [27] to prove that a slope can be enforced by such rules if and only if it is computable.

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# Reaction-Based Models of Biochemical Networks 

Daniela Besozzi ${ }^{1,2(\boxtimes)}$<br>${ }^{1}$ Department of Informatics, Systems and Communication, University of Milano-Bicocca, Viale Sarca 336, 20126 Milano, Italy<br>daniela.besozzi@unimib.it<br>${ }^{2}$ SYSBIO.IT, Centre of Systems Biology, Milano, Italy


#### Abstract

Mathematical modeling and computational analyses of biological systems generally pose to modelers questions like: "Which modeling approach is suitable to describe the system we are interested in? Which computational tools do we need to simulate and analyze this system? What kind of predictions the model is expected to give?". To answer these questions, some general tips are here suggested to choose the proper modeling approach according to the size of the system, the desired level of detail for the system description, the availability of experimental data and the computational costs of the analyses that the model will require. The attention is then focused on the numerous advantages of reaction-based modeling, such as its high level of detail and easy understandability, or the possibility to run both deterministic and stochastic simulations exploiting the same model. Some notes on the computational methods required to analyze reaction-based models, as well as their parallelization on Graphics Processing Units, are finally provided.


## 1 Introduction

In many fields of life sciences, researchers are taking more and more frequently advantage of mathematical modeling and computational analysis as complementary tools to experimental laboratory methods. In the recent years, plenty of multidisciplinary works largely proved the increased capability of the synergistic integration between computational and experimental research, to achieve faster and better comprehension of biological systems. In this context, many frameworks and methodologies that were originally developed in computer science and engineering have been applied to study a variety of biological systems [1,2]. Anyway, the choice of the most adequate modeling and computational approach is not a straightforward process, which should always be guided by the nature of the system of interest, by the experimental data in hand and, above all, by the scientific question that motivates the development of the model.

Many times, laboratory experiments simply aim at identifying the existence of some sort of "difference" between what is happening in the biological system before/after a perturbation, or with/without a key component of that system (for instance, the response to a stress state vs. the physiological state, the phenotype of a mutant organism vs. the wild type, etc.). In these cases, statistical methods can be used to establish the significance of such a difference by using experimental
data that are generally easy to measure [3]. When the scope of the modeling is to make predictions about the emergent behaviours of the system in conditions that were not previously analyzed, then also the complexity and the technical challenges of laboratory experiments increase up to the measurement of quantitative timeseries, which most of the time require time-consuming and expensive procedures $[4,5]$. In these situations, the available biotechnological equipments or experimental methodologies could represent a bottleneck to the measurement of biological data at the level of precision or sampling frequency that are beneficial for mathematical modeling, therefore limiting or hampering the effective application of many computational approaches. Albeit the drawbacks that characterize these complex experiments, the availability of reliable time-series data is an indispensable mean to calibrate mechanistic parametric models, which are usually assumed to be the most likely candidate to achieve an in-depth comprehension of biological systems. Indeed, these models can lead to quantitative predictions of the system dynamics: by considering in details the mutual regulations among the system components, they allow to infer the possible temporal evolution of the system in both physiological and perturbed conditions.

Although some general guidelines can be suggested to facilitate the choice of the modeling approach, the most widespread strategies-that is, interactionbased, constraint-based, mechanism-based and logic-based models-are characterized by peculiar limits and strengths, as discussed hereby. One of the most challenging problem to date, anyway, is how to define integrated models having a unified mathematical formalism and able to bridge different layers of the cell functioning (gene expression, signal transduction, metabolism, etc.) $[6,7]$. Although some striking examples of this integration exist [8], the problem is still far from a general-purpose solution.

The aim of this paper is to provide a brief (non-exhaustive) overview of existing modeling approaches, together with some general tips for choosing the proper formalism for any given biological system under investigation. In particular, the attention is here focused on the advantages of reaction-based modeling, a mechanistic and parametric approach especially useful to describe and analyze networks of biochemical reactions [9]. This modeling approach was successfully used to investigate various signal transduction pathways and to predict their emergent behaviour under different conditions (see, e.g., [10-16]). The analysis of reaction-based models typically involves the execution of stochastic or deterministic simulations, and most often requires to face challenging computational problems necessary for model definition and calibration. The benefits of using high-performance computing solutions to parallelize the simulations and reduce the computational burden of these tasks are concisely sketched as conclusive remarks to this work.

## 2 A Few Tips for the Modeling of Biological Systems

The starting point for the definition of the model of a biological system is, in general, a diagram or a conceptual map, graphically representing the current biological knowledge about that system. In general, these diagrams show


|  | No. | Reaction |
| :---: | :---: | :---: |
|  | $\mathrm{r}_{1}$ | Ras2 $\bullet$ GDP + Cdc25 $\rightarrow$ Ras $2 \bullet$ GDP $\bullet$ Cdc 25 |
|  | $\mathrm{r}_{2}$ | Ras2 $\cdot \mathrm{GDP} \cdot \mathrm{Cdc} 25 \rightarrow$ Ras2 $\bullet$ GDP +Cdc 25 |
|  | $\mathrm{r}_{3}$ | Ras2 $\bullet$ GDP • Cdc25 $\rightarrow$ Ras2 $\bullet$ Cdc25 + GDP |
|  | $\mathrm{r}_{4}$ | Ras2 $\bullet$ Cdc $25+\mathrm{GDP} \rightarrow$ Ras2 $\bullet$ GDP $\bullet$ Cdc 25 |
|  | $\mathrm{r}_{5}$ | Ras2 $\bullet$ Cdc $25+\mathrm{GTP} \rightarrow$ Ras2 $\bullet$ GTP $\bullet$ Cdc 25 |
|  | $\mathrm{r}_{6}$ | Ras2 $\cdot$ GTP $\cdot$ Cdc25 $\rightarrow$ Ras2 $\bullet$ Cdc $25+$ GTP |
|  | $\mathrm{r}_{7}$ | Ras2 $\cdot \mathrm{GTP} \cdot \mathrm{Cdc} 25 \rightarrow$ Ras2 $\bullet$ GTP + Cdc 25 |
|  | $\mathrm{r}_{8}$ | Ras2 $\bullet$ GTP + Cdc $25 \rightarrow$ Ras $2 \cdot \mathrm{GTP} \cdot \mathrm{Cdc} 25$ |
|  | $\mathrm{r}_{9}$ | Ras2 $\bullet$ GTP + Ira2 $\rightarrow$ Ras2 $\bullet$ GTP •Ira2 |
|  | $\mathrm{r}_{10}$ |  |

Fig. 1. The switch cycle of protein Ras2: from diagrams to reaction-based model. Panels A and B represent the interactions among proteins Ras2, Cdc25, Ira2, and nucleotides GTP and GDP. Briefly, Ras2 is a GTPase - that is, an enzyme that catalyzes the hydrolysis of guanosine triphosphate (GTP) into guanosine diphosphate (GDP) - which cycles between an active state, when bound to GTP, and an inactive state, when bound to GDP. The active state of Ras2 is positively regulated by Cdc25, a protein that stimulates the GDP to GTP exchange, and negatively regulated by Ira2, a protein that stimulates the GTPase activity of Ras2. The intracellular ratio of GTP and GDP is also involved in the regulation of Ras2 activity, since Cdc25 stimulates the exchange of these nucleotides according to their relative concentration. Panels $\mathbf{C}$ and $\mathbf{D}$ formally describe the cascade of biochemical reactions and the formation of molecular complexes among proteins and nucleotides.
the molecular components involved in the system and their mutual relations, usually specifying also the formation of molecular complexes, and they possibly highlight the activation/inhibition of some component or the presence of positive/negative feedback regulations. For instance, panels A and B in Fig. 1 show two similar diagrammatic representations of the switch cycle of Ras2 proteins (see caption for more details). Although both diagrams in Fig. 1 represent the same biochemical process, panel B better characterizes the roles of GTP and GDP in their interaction with proteins Ras2 and Cdc25, which is instead not inferable by panel A. This example emphasizes that diagrams of cellular processes could be subject to ambiguous interpretation, or might not explicitly represent all biological knowledge about that process.

In order to avoid any possible misinterpretations, mathematical modeling should be used to precisely and unambiguously describe what is known to occur within the biological system of interest, and possibly include any hypothesis that should then be tested by further laboratory experiments. In the case of the switch cycle of Ras2 protein, an example of precise formalization of these diagrams is given in panel C, which shows a typical representation of reactions scheme used
in biochemistry, while panel D shows the reaction-based model that explicitly describes all direct and inverse biochemical reactions occurring in Ras2 switch cycle (the interested reader is referred to $[12,15]$ for more details about the modeling of this process in the context of the Ras/cAMP/PKA signal transduction pathway in yeast). Reaction-based models belong to the family of mechanismbased modeling approach, which is hereby discussed in the context of the choice of the most adequate formalism for the biological system of interest.

Considering the available knowledge about the biochemical, physical and regulatory properties of all system components and their mutual interactions, the very first step in the definition of a mathematical model consists in the establishment of the purpose (or scientific questions) of the model. This step is fundamental in order to recognize what kind of novel insights the modeling process should add to the experimental investigation. Although some general guidelines can be suggested to facilitate the choice of the modeling approach, the four features schematized in Fig. 2 emphasize the limits and the strengths of the most widespread strategies, that is, interaction-based, constraint-based, mechanism-based and logic-based models (see figure caption for more details).

Unfortunately, there exists no biunivocal correspondence between each of these modeling approaches and any specific biological system, that allows to define a crisp separation among the variety of mathematical formalisms and to select the most suitable strategy for the scientific question under examination. Actually, this is a game with no winners: Fig. 2 shows that mechanism-based models are characterized by the highest level of detail in the formalization of the system, so that they can achieve quantitative predictions of the system dynamics, but they require a lot of data measurements and usually have the highest computational costs. On the other side of the spectrum, interaction-based models can be analyzed with generally low execution times, but they are characterized by the lowest level of detail and hence, in general, they are only used to investigate topological properties of biological networks. Constraint-based models are used to investigate flux distributions in metabolic networks, up to genome-wide size systems, while Boolean/fuzzy logic models have been exploited to study gene regulatory networks or signal transduction pathways having different system sizes. Because of space limits, the interested reader is referred to [17-23] for more information on these modeling approaches and examples of their application.

Despite their highest predictive power with respect to other modeling approaches, mechanism-based models are generally subject to a relevant criticism: usually, they are not exploited to formally describe systems with a large number of components and interactions, so that they cannot be used to explain the functioning of large-scale systems or whole cells. Anyway, even a small-scale quantitative model can be useful to better understand a complex cellular process, together with its feedback/feedforward regulations and non-linear behaviour, and predict its dynamics in physiological and perturbed conditions. To this aim, the model should be as detailed as possible, but should also retain a simple formalization, so that also experimentalists can easily intervene in the discussion and in the development of the model itself. Reaction-based models satisfy these conditions: some of their peculiar strengths are described in the following section.


Fig. 2. Given a biological system and the purpose of the model, the modeling approach should be chosen by taking into account four aspects. First, the size of the system, defined in terms of the number of components and interactions, which can be in the order of a few units or tens for small-scale models, up to hundreds and thousands for large-scale models. Second, the necessary level of detail has to be chosen to fully describe the components of the system and their mutual interactions. Third, the type and the quality of experimental data that are already available or could be measured with proper protocols often constrain the model definition according to a specific approach. Fourth, the computational burden that the chosen approach will bring about in the simulation and analysis of the model, and that sometimes might represent the trade-off in the choice between quantitative or qualitative models. These features usually (but not strictly) lead to model signal transduction or small metabolic pathways with the mechanism-based approach; protein-protein interaction networks with the interaction-based approach; gene regulation networks with the interaction/logicbased approach; whole-cell metabolic networks with the constraint-based approach. The modeling approaches are graphically represented with an increasing intensity scale, according to their quantitative feature: for instance, interaction-based models are characterized by a simplified representation of the system and can only derive qualitative (topological) outcomes from computational analyses, while mechanism-based models are highly detailed and fully parameterized, and have the highest quantitative predictive capability. In order to be predictive, models should be defined in such a way that both the expected and the unknown dynamical patterns are not directly "hard-wired" into the mathematical description, but they globally appear as the consequence of the chemico-physical interactions between the species included in the model.

## 3 The Advantages of Reaction-Based Models

Given a biological system $\Omega$, a reaction-based model of $\Omega$ formally consists in a set of molecular species $\mathcal{S}=\left\{S_{1}, \ldots, S_{N}\right\}$ and a set of biochemical reactions $\mathcal{R}=\left\{r_{1}, \ldots, r_{M}\right\}$. A generic biochemical reaction is given in the form $r_{j}$ : $\sum_{i=1}^{N} \alpha_{j i} S_{i} \rightarrow \sum_{i=1}^{N} \beta_{j i} S_{i}$, where $\alpha_{j i}, \beta_{j i} \in \mathbb{N}$ are the stoichiometric coefficients
of the $i$-th reactant and the $i$-th product of the $j$-th reaction, respectively, which specify how many molecules of each species are involved in the reaction.

It is worth highlighting that reaction-based models are quantitative parametric models, which means that each species $S_{i}$ is characterized by an amount $X_{i}$ (given either as an integer number of molecules or a real-value concentration), and each reaction $r_{j}$ is characterized by a kinetic constant $k_{j} \in \mathbb{R}^{+}$, encompassing its physical and chemical properties. The temporal evolution of $\Omega$ can be simulated by updating the state of the system at consecutive time steps, where the state at some time $t$ is a vector $\mathbf{X}(t)=\left(X_{1}(t), \ldots, X_{N}(t)\right)$, having as components the amounts of all species at time $t$. Usually, the initial state of the system is given by specifying the amount of each species at some fixed $t=0$, chosen as the beginning time point of both experimental measurement and computational analysis.

Since it relies on syntactic and semantic features similar to the language of biochemistry, reaction-based modeling represents for experimentalists the most easily comprehensible formalization of a biological system. Indeed, it does not require any expertise in mathematical or computational modeling. This is a key strength of reaction-based models, since their correctness and biological plausibility can be easily assessed by experimentalists, without the need for modelers to translate the formal (mathematical) system into a "narrative" (and else possibly ambiguous) description of the model itself. Therefore, reaction-based models can largely facilitate the communication and the crosstalk between modelers and experimentalists, and possibly step up interdisciplinary analysis of biological systems.

Reaction-based modeling does not rely on the use of approximate kinetic functions, such as Michaelis-Menten rate law for enzymatic processes or Hill functions for cooperative binding, that are usually exploited in systems of differential equations. On the contrary, it explicitly provides a detailed and accurate description of the molecular interactions and control mechanisms (including feedback or feedforward regulation) that take place in cellular processes. In addition, since all molecular species and their mutual biochemical reactions appear as atomic entities in the model, they can be analyzed either independently from each other or in combination with other components, in order to determine the corresponding influence on the system behaviour. This feature allows modelers to measure all system components and, above all, to precisely determine the role of each species and reaction in the overall functioning of the system.

The reaction-based formalization is general enough to describe any kind of process determined by interacting components, by simply assigning the appropriate semantics to the set of species and to the set of reactions. For instance, it can be easily used to model ecological systems: in this case, the "species" correspond to the individuals of different populations, while the "reactions" describe the type of relationships occurring between these populations. Assuming, e.g., prey-predator interactions between two populations, as in the LotkaVolterra equations, a reaction-based model can be specified by considering (1) the set $\mathcal{S}=\{A, X, Y\}$, where $A, X, Y$ represent, respectively, the food resource, the prey and the predator individuals, and (2) the set of reactions $\mathcal{R}=\left\{r_{1}: A+X \rightarrow X+X ; r_{2}: X+Y \rightarrow Y+Y ; r_{3}: Y \rightarrow \lambda\right\}$, where $r_{1}$ describes the growth of preys, according to available resources, $r_{2}$ describes the
interaction between preys and predators and the consequent growth of predators, while $r_{3}$ describes the natural death of predators (symbol $\lambda$ here denotes the die-off of individual $Y$ ).

In contrast to other formalisms, any reaction-based model can be easily refined or extended without any labored adjustment in the formalization of the former model (see also [24]). Adding new species or reactions to a model formalized as a system of differential equations would instead require, in general, the modification of many of its differential equations. Thanks to this feature, reaction-based modeling is suitable to modular construction of larger and larger models, whereby an initial core of species and reactions can be extended to take into account other cellular processes (e.g., to introduce new species or new reactions after the former model has been validated with laboratory experiments).

Reaction-based models can be exploited to run both stochastic and deterministic simulations of the temporal evolution of the system. On the one hand, according to the stochastic formulation of chemical kinetics [25], the formalization of a biochemical network into a set of species and a set of reactions can be straightforwardly used to run stochastic simulation algorithms [26]. On the other hand, any reaction-based model can be automatically converted into a corresponding system of ordinary differential equations, and then simulated by means of some numerical integration algorithm [27]. This automatic conversion can be done by considering the law of mass action [28], an empirical chemical law stating that the rate of a reaction is proportional to the product of the concentration of its reactants, each one raised to the power of the corresponding stoichiometric coefficient. This allows to derive a rate equation for each species appearing in the reaction-based model, formally describing how its concentration changes in time according to all reactions where it appears either as reactant or product. Therefore, any biochemical system defined by means of a reactionbased model can be also formalized as a set of coupled (non-linear) first order ordinary differential equations. The duality of the stochastic and deterministic interpretation of a reaction-based model concerns also reaction constants and molecular amounts, which can be bidirectionally converted into appropriate numerical values-i.e., number of molecules into concentration, stochastic constant into kinetic rate, or viceversa-assuming that the dimension of the volume where reactions take place is known [25,29].

## 4 Conclusive Remarks

In order to gain novel insights into the functioning of a biological system, various computational methods - such as parameter sweep analysis, sensitivity analysis, parameter estimation, reverse engineering (see [30,31] and references therein)can be exploited to make predictions on the way the system behaves both in physiological and perturbed conditions. To this aim, the response of the system is usually analyzed using distinct model parameterizations, that is, different initializations of species amounts and/or reaction constants. These methods require the execution of a large number of simulations, whose computational burden can
rapidly overtake the capabilities of Central Processing Units (CPUs), especially for mechanism-based models. In these cases, high-performance computing solutions can be exploited, such as Graphics Processing Units (GPUs), which recently gained ground in many fields related to life sciences $[32,33]$.

Despite better computational performances, the use of GPUs in scientific applications is weakened by the striking difference between GPU-based and CPU-based computing, which demands specific programming skills to the purpose of fully exploiting the peculiar GPU's architecture. To overcome these drawbacks and to provide user-friendly tools to analyze the dynamics of reactionbased models, coarse-grain GPU implementations of both stochastic and deterministic simulation algorithms were presented in $[34,35]$. These tools allow to simultaneously execute a massive number of simulations, each one characterized by a different model parameterization, and were proven to largely speed up burdensome computational analyses of various cellular processes [13, 34, 35].

An additional benefit is that these parallel simulators can be used as core engines to apply other computational methods, like parameter estimation (PE) or reverse engineering (RE). PE consists in the inference of the unknown values for species amounts or reaction constants, and it is often tackled by means of global optimization techniques $[36,37]$, while RE consists in devising a plausible cascade of spatio-temporal interactions in a network of biochemical reactions. Both problems require the availability of time-series experimental data (e.g., quantitative amount of some molecular species), needed as target to assess the goodness of the inferred model and/or its parameterization.

In the context of reaction-based models, the RE problem was faced in [38], while a PE method was initially introduced in [39]. The latter is based on a Swarm Intelligence heuristics, called Particle Swarm Optimization, which exploits a set of random candidate model parameterizations (each one codified by a particle in the swarm) to the purpose of converging to an optimal solution. In particular, the method in [39] exploits a multi-swarm topology, whereby each swarm is associated to time-series data measured in some experimental condition. The method is inspired by the quite common scenario of experimental research, where experiments are executed to observe the biological system in different conditions, and data are collected in both physiological and perturbed states. In the multi-swarm approach, all swarms cooperate to estimate a unique model parameterization, which is able to generate the system dynamics that better fits all the data that were measured in all tested experimental conditions. The quality of the model parameterization is evaluated by a fitness function defined, e.g., as the distance between the experimental data at each sampled time point, and the corresponding species amounts obtained (at the same time points) by means of dynamic simulations.

This PE method was further refined to fully exploit the parallel acceleration granted by GPUs [40], as well as to take advantage of fuzzy rule-based systems [41] or to test different strategies for the initialization of candidate solutions (i.e., particles' position within the search space) [42]. These refinements were shown to improve both the computational and effectiveness performance of the PE method.

Hopefully, the dissemination of the advantages of reaction-based models and of GPU-accelerated tools for their analysis will provoke a widespread adoption of this formalism by modelers. This could allow a more easy definition and validation of mathematical models, the prediction of unknown behaviours and the identification of critical factors of biochemical networks in reduced times, as well as a faster design of focused laboratory experiments to better comprehend the malfunctioning of cellular processes.

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# Comparative Genomics on Artificial Life 

Priscila Biller ${ }^{1,2}$, Carole Knibbe ${ }^{2,3}$, Guillaume Beslon ${ }^{2,3}$, and Eric Tannier ${ }^{2,4(\boxtimes)}$<br>${ }^{1}$ University of Campinas, São Paulo, Brazil<br>${ }^{2}$ INRIA Grenoble Rhône-Alpes, 38334 Montbonnot, France eric.tannier@inria.fr<br>${ }^{3}$ Université Lyon 1, LIRIS, UMR5205, 69622 Villeurbanne, France<br>${ }^{4}$ Université Lyon 1, LBBE, UMR5558, 69622 Villeurbanne, France


#### Abstract

Molecular evolutionary methods and tools are difficult to validate as we have almost no direct access to ancient molecules. Inference methods may be tested with simulated data, producing full scenarios they can be compared with. But often simulations design is concomitant with the design of a particular method, developed by a same team, based on the same assumptions, when both should be blind to each other. In silico experimental evolution consists in evolving digital organisms with the aim of testing or discovering complex evolutionary processes. Models were not designed with a particular inference method in mind, only with basic biological principles. As such they provide a unique opportunity to blind test the behavior of inference methods. We give a proof of this concept on a comparative genomics problem: inferring the number of inversions separating two genomes. We use Aevol, an in silico experimental evolution platform, to produce benchmarks, and show that most combinatorial or statistical estimators of the number of inversions fail on this dataset while they were behaving perfectly on ad-hoc simulations. We argue that biological data is probably closer to the difficult situation.


Keywords: Comparative genomics • In silico experimental evolution • Benchmark • Rearrangements

## 1 Validation of Evolutionary Inferences

The comparative method in evolutionary biology consists in detecting similarities and differences between extant organisms, and, based on more or less formalized hypotheses on the evolutionary processes, infer ancestral states explaining the similarities and an evolutionary history explaining the differences.

A common concern in all evolutionary studies is the validity of the methods and results. Results concern events that were supposed to occur in a deep past (up to 4 billion years) and they have no other trace today than the present molecules used by the comparative method.

As we cannot travel back in time to verify the results, there are several ways to assess the validity of molecular evolution studies: theoretical considerations about the models and methods (realism, consistency, computational complexity,
model testing, ability to generate a statistical support or a variety of the solutions) [24], coherence with fossil records [26], or ancient DNA [11], or empirical tests when the solution is known, on experimental evolution [17] or simulations. Each method has its caveats. Models for inference have to adopt a compromise between realism, consistency and complexity. Ancient DNA is rarely available, usually not in an assembled shape. Fossils are also rare and provide a biased sampling of ancient diversity. Experimental evolution is expensive, time-consuming and limited in the number of generations it can provide.

Simulation is the most popular validation tool. Genome evolution can be simulated in silico for a much higher number of generations than in experimental evolution, at a lower cost. All the history can be recorded in details, and compared with the inference results. A problem with simulations, however, is that they necessarily oversimplify genome evolution processes. Moreover, very often, even if they are designed to be used by another team for inference $[4,10,14,15,23]$, they encode the same simplifications as the inference methods. For example, only fixed mutations are generated because only these are visible by inference methods, selection is tuned to fit what is visible by the inference methods; genes are evolutionary units in simulations because they are the units taken for inference. Everything is designed thinking of the possibilities of the inference methods, leading to easy unrealistic instances.

This mode of $a d-h o c$ simulation has been widely applied to test estimators of rearrangement distances, and in particular inversion distances [5, 7, 9, 12, 22]. The problem consists in comparing two genomes and estimating the number of inversions (a rearrangement that reverses the reading direction of a genomic segment) that have occurred in the evolutionary lineages separating them. To construct a solution, conserved genes or synteny blocks are detected in the two genomes, and a number of inversions explaining the differences in gene orders is estimated. A lot of work has consisted in finding shortest scenarios [13]. Statistical estimations need a model. The standard and most used model depicts genomes as permutations of genes and assumes that an inversion reverses a segment of the permutation, taken uniformly at random over all segments. When simulators are designed to validate the estimators, they also use permutations as models of gene orders, and inversions on segments of this permutations, chosen uniformly at random. Estimators show good performances on such simulations, but transforming a genome into a permutation of genes is such a simplification from both parts that it means nothing about any ability to estimate a rearrangement distance in biological data [8].

We propose to use simulations that were not designed for validation purposes. It is the case, in artificial life, of in silico experimental evolution [18], and in particular of the Aevol platform [3,19]. Aevol contains, among many other features, all what is needed to test rearrangement inference methods. The genomes have gene sequences and non coding sequences organized in a chromosome, and evolve with inversions, in addition to substitutions, indels, duplications, losses, translocations. Rearrangements are chosen with a uniform random model on the genome, which should fit the goals of the statistical estimators, but is different
from a uniform random model on permutations [8]. We tested 10 different estimators of inversion distance on 18 different datasets generated by Aevol. The difference with ad-hoc simulations is striking. Most estimators completely fail to give a close estimate in a vast majority of conditions. We argue that the reason for this failure lies in realistic features in artificial genomes that are very likely to reproduce the failure on real data.

We first describe the principle of the estimators, then the principles of the simulator, with its goals and its functioning. We will show how to process its results to test statistical estimators of rearrangement distances.

## 2 Comparative Genomics: Estimating an Inversion Distance

We tested 10 estimators of the number of inversions separating two genomes, called ID (the inversion distance) [16], CL for Caprara and Lancia [9], EH for Eriksen and Hultman [12], Badger [20], BD for Berestycki and Durrett [5], LM for Lin and Moret [22], BGT for Biller, Guéguen and Tannier [7], AA for Alexeev and Alekseyev [2], ER1 and ER2 for Erdös-Renyi 1 and 2 [8].

For 8 of them (ID, LM, BGT, Badger, EH, BD, CL, AA), a genome is defined as a signed permutation, $\pi$ over $\{1, \ldots, n\}$, that is, an ordering of the elements of $\{1, \ldots, n\}$ where each element is given a sign, + or $-(+$ usually omitted), representing the reading direction of an element. The elements of the permutation are genes, or solid regions, the ones that are never cut by inversions. All inversions have the same probability. For the two remaining estimators (ER1 and ER2), a genome is made up of two components: the same signed permutation, and in addition a vector $p$ of $n+1$ breakage probabilities, $p_{i}>0,0 \leq i \leq n$, with $\sum_{i} p_{i}=1$. An inversion of the segment $\left[\pi_{i}, \ldots, \pi_{j}\right]$ has probability $p_{i-1} p_{j}$.

Suppose $A$ and $B$ are two signed permutations. We define the breakpoint graph of $A$ and $B$ as the graph with $2 n+2$ vertices and $2 n+2$ edges: for each element $i \in\{1, \ldots, n\}$, define two vertices $i_{t}$ and $i_{h}$, plus two additional vertices $0_{h}$ and $n+1_{t}$; then for any two consecutive numbers $a b$ of $A$, join two extremities by an $A$-edge: first is $a_{h}$ if $a$ is positive, $a_{t}$ otherwise, second is $b_{t}$ if $b$ is positive, $b_{h}$ otherwise. Additionally, if $a$ is the first element of the permutation, join $0_{h}$ and $a_{t}$ if $a$ is positive, $a_{h}$ otherwise, and if $b$ is the last element of the permutation, join $n+1_{t}$ and $b_{h}$ if $b$ is positive, $b_{t}$ otherwise. Do the same for $B$, and call the edges $B$-edges.

An adjacency of a genome $A$ is an $A$-edge in the breakpoint graph. It is a common adjacency with a genome $B$ if it is also a $B$-edge, otherwise it is a breakpoint. Breakpoint graphs have a uniform degree of 2 on all vertices, thus they are sets of disjoint cycles alternating between $A$-edges and $B$-edges. We note $b$ the number of breakpoints, $c$ the number of cycles of the breakpoint graph, and $c_{2}$ the number of cycles with 4 edges.

The parsimony estimator (ID) is the minimum number of inversions necessary to transform $A$ into $B$, which is close to $n+1-c$ [16]. Badger is a Bayesian sampler of inversion scenarios and computes an a posteriori probable distance.

The others all work with the method of moments. This consists in computing an expected value for one or two observable parameters of the breakpoint graph ( $b$, $c_{2}, c$ or a combination of two of them) if $A$ and $B$ are separated by $k$ inversions. It is a function of $k$ and $n: f_{n}(k)$. It is never computed exactly, approximate formulas or computation principles are given. Then $k$ is estimated as $\hat{k}=f_{n}^{-1}(p)$ for the observed value $p$ of the parameter. LM, CL, BGT, ER1 are based on the expected value of $b$. EH and BD are based on the expected value of $c . \mathrm{ER} 2$ is based on expected values for $b$ and $c_{2}$, and AA uses expected values for $b$ and $c$. The two latter use two values because they also consider $n$ as unknown and estimate it as well as $k$.

## 3 Artificial Life: In Silico Experimental Evolution and the Aevol Platform

Unlike many simulators used to validate phylogenetic inference methods [4, 10, 14, 15,23], Aevol does not represent a species by a single lineage undergoing fixed mutations. Like forward-in-time simulators used in population genetics, it explicitly represents all genotypes present in the population and simulates spontaneous mutations, which can be deleterious, neutral or beneficial. An important difference, however, is that the selection coefficients of mutations are not predefined for each locus nor drawn from a random distribution. Instead, an artificial chemistry is used to decode each genome present in the population and compute its phenotype, which is its ability to perform a computational task (see details below). Point mutations or small indels can alter gene sequences and non coding sequences. A local mutation in a gene can have a different effect on phenotype and fitness, depending on the genomic background (other genes). Chromosomal rearrangements like duplications, deletions, translocations or inversions can occur anywhere in the chromosome sequence. They can alter gene number and gene order and disrupt genes.

Figure 1 summarizes the functioning of Aevol. We give a high level description here, and emphasize that the tool has many other possibilities than being used as a bench mark. For a complete description and some of its possibilities, see $[3,19]$. Genomes are circular sequences on a binary alphabet. A population of typically 1000 genomes lives at a given generation. Genes are segments situated on a transcribed sequence (i.e., a sequence starting after a promoter and ending at a terminator sequence) starting after a Ribosome Binding Start and a Start codon and ended by a Stop codon on the same reading frame. Inside a gene, a coding sequence is translated into a protein sequence using a genetic code on size three codons. This protein sequence encodes the parameters of a piecewise linear function that indicates the contribution (in $[-1,1]$ ) of the protein to each abstract "phenotypic trait" in $[0,1]$. All proteins encoded in a genome are summed to produce the phenotype, which is thus a piece-wise linear function indicating the level of each phenotypic trait in $[0,1]$.

This phenotype is then compared with a target function indirectly representing the environment of the individual. The difference between the two is


Fig. 1. Overview of the Aevol model. (A) A population of genomes is simulated. At each generation, all genomes are evaluated and the fittest ones are replicated to produce the next generation. The replication process includes variation operators. The joint actions of selection, genetic drift and variation make the population evolve. (B) Overview of the genome decoding process. Left: Each individual own a circular doublestrand genome with scattered genes. Right: The individual's phenotype is the level of each abstract phenotypic trait in [0, 1]. It is compared a target representing the optimal phenotype given the environment. Middle: Each gene is decoded into a protein that contributes to a small subset of phenotypic traits. More precisely, the sequence of the gene is decoded into three reals that specify the mean, width and height of a triangular kernel function. All the proteins are then summed up to compute the phenotype. The individual displayed here was obtained after 460.000 generations of evolution in Aevol under a mutational pressure of $10^{-6}$ mutations/bp/generation for local events and $10^{-5}$ mutations $/ \mathrm{bp} /$ generation for chromosomal rearrangements (see below). Its genome is 6898 bp long. It encodes 113 genes and 35 RNAs (not shown). $28.4 \%$ of the genome is non-coding. (C) Overview of the replication process. During its replication each genome may undergo chromosomal rearrangements affecting large DNA segments (here an inversion and a translocation) and local mutations (point mutations or small InDels).
used to compute the fitness of the genome. To produce the next generation, genomes with high fitness are replicated in the following generation with higher probability than genomes with low fitness. During replication, local mutations and chromosomal rearrangements are performed on the genomes, at a spontaneous rate fixed by simulation parameters.

The population is initialized with a same random genome containing at least one gene. As generations go by, neutral, deleterious or beneficial mutants appear and their frequencies in the population vary depending on natural selection and genetic drift. The target function is better and better approximated and the genome structure evolves to eventually contain between tens to hundreds of genes (depending on the evolutionary conditions) scattered along the genome.

In silico experimental evolution allows for perfect recording of all mutational events that have occurred in the lineage of any organism. It is thus possible to trace the evolution of a single gene along the generations, and thus to compare genomes from different generations by identifying genes that descend one from the other.

## 4 Inversion Distance Estimators on Artificial Genomes

We propose 18 runs of Aevol to be used as benchmark datasets for comparative genomics studies. All estimators were computed for the 18 experiments, and we show the results for two experiments in Fig. 2. Experiments with 6 different conditions were run 3 times each, with a different seed each time. The conditions concern the allowed mutation types, among: inversions, duplications (where the copied segment is pasted anywhere on the genome), tandem duplications (where it is pasted next to the position of the ancestral segment), losses, translocations, point mutations and small InDels. Mutation rates were set to $5.10^{-6}$ mutation per base per generation for point mutations and InDels, and $10^{-5}$ for larger allowed rearrangements. All runs were stopped after 15000 generations with a genome containing approximately 100 genes. We make accessible, for each of the 18 runs, the input parameters, and for each generation, the list of genes, their coordinates on the genome, and their genealogy (how they relate to each other across generations). Material can be uploaded here: http://aevol.inrialpes. fr/resources/benchmark/cie_2016.

From Aevol output we compute signed permutations of genes without duplicates which model the relative order of genes compared with the last generation. We keep, in each generation, only the genes that have a unique descendant in the last generation, with no duplication event in its history between this generation and the last. Only the last generations can contain such genes, so permutations are only computed for a few hundred generations.

The results for two different runs out of 18 are shown in Fig. 2. The two were chosen for extreme but informative behaviors. The first run allowed for inversions, duplications and small mutations (A), while the second one allowed for translocations and tandem duplications in addition (B). At each generation we keep the genome in the ancestry lineage of the fittest genome at generation 15000 . The true number of inversions is compared with the estimated one, according to 7 estimators (we removed 3 of them because the curves were indistinguishable from others). The results highly depend on the conditions. On the (A) part, all estimators except AA are estimating a rather good number of inversions up to 50 events. On the (B) part, we cut the graph after 100 generations


Fig. 2. The results of 7 estimators: ID, Badger, EH, BGT, ER1, ER2, AA. The other ones gave indistinguishable curves (BD from EH, LM and CL from BGT). $x$ axis is the generation number. $y$ axis is the number of inversions. All generations are compared with the last one, number 15000. The true number is the black solid line, and the others are estimated numbers. These Aevol runs includes (A) inversions, duplications and deletions (B) inversions, duplications, tandem duplication, translocations and deletions. The number of compared genes is from 119 to 109 (A), and from 92 to 49 (B).
because all estimators had lost the signal for a long time. This rapid signal loss is expected because of the accumulation of translocations that blur the inversion signal. On both runs, EH and BD are giving estimates which are almost equal to ID, the parsimony value. BGT, LM, CL and AA are constantly worse than the parsimony value. Only ER1 gives a better estimate than parsimony until generation 600 (after $\sim 200$ inversions) in the first run. On ad-hoc simulators reported in the papers describing the estimators, all 10 estimators gave significantly better results than parsimony for any variation of parameters [5,7,8,12,20, 22].

## 5 Discussion

In our experiments there are many quality differences between estimators. But a general tendency is that after a low true number of events $(\sim n / 3$, where $n$ is the number of genes), most of them significantly underestimate the true value. This highly contrasts with the claimed performances of these estimators. For example ID is supposed to have great chance of giving the right value up to $n / 2$, while LM, EH, BD, BGT all have been tested on simulations and reported to give the right value far above $n$ [5, 7, 12, 22].

We argue that our datasets are not artefactually difficult (nor purposely made difficult), and that the poor results encountered here are susceptible to reflect real results on biological data. One argument for this is the better behavior of ER1 in several situations, including the one depicted on Fig. 2(A). The addition of ER1 compared to the other estimators is that it takes into account the distribution of intergene sizes. It suggests that part of the failure of the other estimators can be explained by this ignorance of intergene sizes. In biological data, intergene sizes influence probabilities of breakages, as it has been shown several times on mammals for example [ 6,21$]$.

Some estimators have been tried on biological data. The inversion distance is often used. Badger has been used several times to reconstruct bacterial or mitocondrial gene orders [20], and AA has been used to estimate distances between Yeast genomes [2], and ER2 on amniote data [7]. The results have to be read in regard of this study on artificial life.

Part of the discrepancy between the true value and the estimated value remains unexplained. The complexity of the real scenarios probably blurs the signal that estimators are able to capture. But again, this complexity is not a specificity of Aevol, and is probably encountered in biological data. So by this simple experiment we can worry that none of the existing estimators of rearrangement distance would be able to produce a plausible value on real genomes.

Future Work. We tested only the estimation of the number of inversions. But only with the runs we have already computed, a lot more can be done: estimation of the proportion of translocations as in [1], or estimating both inversions and duplications as in [25]. Artificial genomes could in principle not only be used by comparative genomics inference methods, but by a larger set of molecular evolution studies. For the moment the sequences are made of 0 s and 1 s , which is
not a problem to study gene order, but can be disturbing for sequence analyses. This way of coding sequences is on another hand a good sign that Aevol was not developed for benchmarking purposes. In a close future nucleotidic and proteic sequences with the biological alphabet will be added to extend the benchmarking possibilities of the model.

Also we work with only one lineage, and compare only two genomes here, because Aevol evolves a single population. A useful addition will be speciation processes, in order to be able to compare several genomes.

On the Blind Multidisciplinarity. This study experiments a singular kind of interdisciplinarity. Obviously communities from comparative genomics and artificial life have to work together in order to make such results possible. But, on the opposite, these results are only possible because both communities first work in relative isolation. If they had defined their working plans together, spoke to each other too often or influenced each other's way of thinking evolutionary biology, the work would have lost some value. Indeed, what makes the difficulty here for comparative genomicists is that they have to infer histories on data for which they have no stranglehold on the processes, just as for biological data, but on which they also have the correct answer, just not as for biological data.

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# Computability and Analysis, a Historical Approach 

Vasco Brattka ${ }^{1,2(\boxtimes)}$<br>${ }^{1}$ Faculty of Computer Science, Universität der Bundeswehr München, München, Germany<br>${ }^{2}$ Deptartment of Mathematics and Applied Mathematics, University of Cape Town, Cape Town, South Africa<br>Vasco.Brattka@cca-net.de


#### Abstract

The history of computability theory and the history of analysis are surprisingly intertwined since the beginning of the twentieth century. For one, Émil Borel discussed his ideas on computable real number functions in his introduction to measure theory. On the other hand, Alan Turing had computable real numbers in mind when he introduced his now famous machine model. Here we want to focus on a particular aspect of computability and analysis, namely on computability properties of theorems from analysis. This is a topic that emerged already in early work of Turing, Specker and other pioneers of computable analysis and eventually leads us to the very recent project of classifying the computational content of theorems in the Weihrauch lattice.


## 1 Introduction

Probably Émil Borel was the first mathematician who had an intuitive understanding of computable real number functions and he anticipated some basic ideas of computable analysis as early as at the beginning of the 20th century. It was in his introduction to measure theory where he felt the need to discuss such concepts and we can find for instance the following crucial observation in $[5,6]$.

Theorem 1 (Borel 1912). Every computable real number function $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$ is continuous.

Strictly speaking, Borel's definition of a computable real number function was a slight variant of the modern definition (see [2] for details and translations) and his definition was informal in the sense that no rigorous notion of computability or of an algorithm was available at Borel's time.

It was only Alan Turing who introduced such a notion with the help of his now famous machine model [43,44]. Interestingly, also Turing was primarily interested in computable real numbers (hence the title of his paper!) and not so much in functions and sets on natural numbers that are the main objects

[^4]of study in modern computability theory. Turing's definition of a computable real number function is also a slight variant of the modern definition (see [2] for details).

We conclude that computability theory was intertwined with analysis since its early years and here we want to focus on a particular aspect of this story that is related to computability properties of theorems in analysis, which are one subject of interest in modern computable analysis [17,27,38, 45].

## 2 Some Theorems from Real Analysis

In his early work [43] Turing already implicitly discussed the computational content of some classical theorems from analysis. Some of his rather informal observations have been made precise later by Specker and others [2]. Ernst Specker was probably the first one who actually gave a definition of computable real number functions that is equivalent to the modern one [41,42]. The following theorem is one of those theorems that are implicitly discussed by Turing in [43].

Theorem 2 (Monotone Convergence Theorem). Every monotone increasing and bounded sequence of real numbers $\left(x_{n}\right)_{n}$ has a least upper bound $\sup _{n \in \mathbb{N}} x_{n}$.

What Turing observed (without proof) is that for a computable sequence $\left(x_{n}\right)_{n}$ of this type the least upper bound is not necessarily computable. A rigorous proof of this result was presented only ten years later by Specker [41].

Proposition 3 (Turing 1937, Specker 1949). There is a computable monotone increasing and bounded sequence $\left(x_{n}\right)_{n}$ of real numbers such that $x=$ $\sup _{n \in \mathbb{N}} x_{n}$ is not computable.

Specker used (an enumeration of) the halting problem $K \subseteq \mathbb{N}$ to construct a corresponding sequence $\left(x_{n}\right)_{n}$ and such sequences are nowadays called Specker sequences. Then the corresponding non-computable least upper bound is $x=$ $\sum_{i \in K} 2^{-i}$. While the Monotone Convergence Theorem is an example of a noncomputable theorem, Turing also discusses a (special case) of the Intermediate Value Theorem [43], which is somewhat better behaved.

Theorem 4 (Intermediate Value Theorem). Every continuous function $f:[0,1] \rightarrow \mathbb{R}$ with $f(0) \cdot f(1)<0$ has a zero $x \in[0,1]$.

And then Turing's observation, which was stated for the general case by Specker [42] could be phrased in modern terms as follows.

Proposition 5 (Turing 1937, Specker 1959). Every computable function $f:[0,1] \rightarrow \mathbb{R}$ with $f(0) \cdot f(1)<0$ has a computable zero $x \in[0,1]$.

A rigorous proof could utilize the trisection method, a constructive variant of the well-known bisection method and can be found in [45]. Hence, while the Monotone Convergence Theorem does not hold computably, the Intermediate Value Theorem does hold computably, at least in a non uniform sense. It was claimed by Specker [42] (without proof) and later proved by Pour-El and Richards [38] that this situation changes if one considers a sequential version of the Intermediate Value Theorem.

Proposition 6 (Specker 1959, Pour-El and Richards 1989). There exists a computable sequence $\left(f_{n}\right)_{n}$ of computable functions $f_{n}:[0,1] \rightarrow \mathbb{R}$ with $f_{n}(0)$. $f_{n}(1)<0$ for all $n \in \mathbb{N}$ and such that there is no computable sequence $\left(x_{n}\right)_{n}$ of real numbers $x_{n} \in[0,1]$ with $f_{n}\left(x_{n}\right)=0$.

For their proof Pour-El and Richards used two c.e. sets that are computably inseparable. Their result indicates that the Intermediate Value Theorem does not hold computably in a uniform sense. In fact, it is known that the Intermediate Value Theorem is not computable in the following fully uniform sense: namely, there is no algorithm that given a program for $f:[0,1] \rightarrow \mathbb{R}$ with $f(0) \cdot f(1)<0$ produces a zero of $f$. Nowadays, we can express this as follows with a partial multi-valued map [45].

Proposition 7 (Weihrauch 2000). IVT: $\subseteq \mathcal{C}[0,1] \rightrightarrows[0,1], f \mapsto f^{-1}\{0\}$ with $\operatorname{dom}(\mathrm{IVT})=\{f \in \mathcal{C}[0,1]: f(0) \cdot f(1)<0\}$ is not computable.

For general represented spaces $X, Y$ we denote by $\mathcal{C}(X, Y)$ the space of continuous functions $f: X \rightarrow Y$ endowed with a suitable representation [45] (and the compact open topology) and we use the abbreviation $\mathcal{C}(X):=\mathcal{C}(X, \mathbb{R})$. In fact, IVT is not even continuous and this observation is related to the fact that the Intermediate Value Theorem has no constructive proof [4].

Another theorem discussed by Specker [42] is the Theorem of the Maximum.
Theorem 8 (Theorem of the Maximum). For every continuous function $f:[0,1] \rightarrow \mathbb{R}$ there exists a point $x \in[0,1]$ such that $f(x)=\max f([0,1])$.

Grzegorczyk [22] raised the question whether every computable function $f:[0,1] \rightarrow \mathbb{R}$ attains its maximum at a computable point. This question was answered in the negative by Lacombe [29] (without proof) and later independent proofs were provided by Lacombe [30, Theorems VI and VII] and Specker [42].

Proposition 9 (Lacombe 1957, Specker 1959). There exists a computable function $f:[0,1] \rightarrow \mathbb{R}$ such that there is no computable $x \in[0,1]$ with $f(x)=$ $\max f([0,1])$.

Similar results have also been derived by Zaslavskii [46]. Specker used a Kleene tree for his construction of a counterexample. A Kleene tree is a computable counterexample to Weak Kőnig's Lemma.

Theorem 10 (Weak Kőnig's Lemma). Every infinite binary tree has an infinite path.

Kleene [26] has proved that such counterexamples exist and (like Proposition 6) this can be easily achieved using two computably inseparable c.e. sets.

Proposition 11 (Kleene 1952). There exists a computable infinite binary tree without computable paths.

It is interesting to note that even though computable infinite binary trees do not need to have computable infinite paths, they do at least have paths that are low, which means that the halting problem relative to this path is not more difficult than the ordinary halting problem. In this sense low paths are "almost computable". The existence of such solutions has been proved by Jockusch and Soare in their now famous Low Basis Theorem [25].

Theorem 12 (Low Basis Theorem of Jockusch and Soare 1972). Every computable infinite binary tree has a low path.

Such low solutions also exist in case of the Theorem of the Maximum 8 for computable instances. The Monotone Convergence Theorem 2 is an example of a theorem where not even low solutions exist in general, e.g., Specker's sequence is already an example of a computable monotone and bounded sequence with a least upper bound that is equivalent to the halting problem and hence not low.

Another case similar to the Theorem of the Maximum is the Brouwer Fixed Point Theorem.

Theorem 13 (Brouwer Fixed Point Theorem). For every continuous function $f:[0,1]^{k} \rightarrow[0,1]^{k}$ there exists a point $x \in[0,1]^{k}$ such that $f(x)=x$.

It is an ironic coincidence that Brouwer, who was a strong proponent of intuitionistic mathematics is most famous for his Fixed Point Theorem that does not admit a constructive proof. It was Orevkov [34] who proved in the sense of Markov's school that there is a computable counterexample and Baigger [3] later proved this result in terms of modern computable analysis.

Proposition 14 (Orevkov 1963, Baigger 1985). There is a computable function $f:[0,1]^{2} \rightarrow[0,1]^{2}$ without a computable $x \in[0,1]^{2}$ with $f(x)=x$.

One can conclude from Proposition 5 that such a counterexample cannot exist in dimension $k=1$. Baigger also used two c.e. sets that are computably inseparable for his construction.

Yet another theorem with interesting computability properties is the Theorem of Bolzano-Weierstraß.

Theorem 15 (Bolzano-Weierstraß). Every sequence $\left(x_{n}\right)_{n}$ in the unit cube $[0,1]^{k}$ has a cluster point.

It was Rice [39] who proved that a straightforward computable version of the Bolzano-Weierstraß Theorem does not hold and Kreisel pointed out in his review of this article for the Mathematical Reviews of the American Mathematical Society that he already proved a more general result [28].

Proposition 16 (Kreisel 1952, Rice 1954). There exists a computable sequence $\left(x_{n}\right)_{n}$ in $[0,1]$ without a computable cluster point.

In fact, this result is not all too surprising and a simple consequence of Proposition 3. What is more interesting is that in case of the Bolzano-Weierstraß Theorem there are even computable bounded sequences without limit computable cluster point. Here a point is called limit computable if it is the limit of a computable sequence. However, this was established only much later by Le Roux and Ziegler [31] (answering a question posed by Giovanni Lagnese on the email list Foundations of Mathematics [fom] in 2006).

Proposition 17 (Le Roux and Ziegler 2008). There exists a computable sequence $\left(x_{n}\right)_{n}$ in $[0,1]$ without a limit computable cluster point.

This is in notable contrast to all aforementioned results that all admit a limit computable solution for computable instances. For instance, the Monotone Convergence Theorem 2 itself implies that every monotone bounded sequence is convergent and hence every computable monotone bounded sequence automatically has a limit computable supremum. Hence, in a certain sense the BolzanoWeierstraß Theorem is even less computable than all the other results mentioned in this section.

In this section we have only discussed a selection of theorems that can illustrate a certain variety of possibilities that occur. The computational content of several other theorems from real analysis has been studied. For instance Aberth [1] constructed a computable counterexample in the Russian sense for the Peano Existence Theorem for solutions of ordinary differential equations. Later Pour-El and Richards [37] constructed another counterexample in the modern sense of computable analysis for this theorem. The Riemann Mapping Theorem is an interesting example of a theorem from complex analysis that was studied by Hertling [23]. Now we turn to functional analysis.

## 3 Some Theorems from Functional Analysis

Starting with the work of Metakides, Nerode and Shore [32,33] theorems from functional analysis were studied from the perspective of computability theory. In particular the aforementioned authors studied the Hahn-Banach Theorem.

Theorem 18 (Hahn-Banach Theorem). Let $X$ be a normed space over the field $\mathbb{R}$ with a linear subspace $Y \subseteq X$. Then every linear bounded functional $f: Y \rightarrow \mathbb{R}$ has a linear bounded extension $g: X \rightarrow \mathbb{R}$ with $\|g\|=\|f\|$.

Here $\|f\|:=\sup _{\|x\| \leq 1}|f(x)|$ denotes the operator norm. The result holds analogously over the field $\mathbb{C}$. Here and in the following a computable metric space $X$ is just a metric space together with a dense sequence such that the distances can be computed on that sequence (as a double sequence of real numbers). If the space has additional properties or ingredients, such as a norm that generates the metric, then it is called a computable normed space or in case of completeness
also a computable Banach space. If, additionally, the norm is generated by an inner product, then the space is called a a computable Hilbert space. A subspace $Y \subseteq X$ is called c.e. closed if there is a computable sequence $\left(x_{n}\right)_{n}$ in $X$ such that one obtains $\overline{\left\{x_{n}: n \in \mathbb{N}\right\}}=Y$ (where the $\bar{A}$ denotes the closure of $A$ ). Metakides, Nerode and Shore [33] constructed a computable counterexample to the Hahn-Banach Theorem.

Proposition 19 (Metakides, Nerode and Shore 1985). There exists a computable Banach space $X$ over $\mathbb{R}$ with a c.e. closed linear subspace $Y \subseteq X$ and a computable linear functional $f: Y \rightarrow \mathbb{R}$ with a computable norm $\|f\|$ such that every linear bounded extension $g: X \rightarrow \mathbb{R}$ with $\|g\|=\|f\|$ is noncomputable.

Similarly, as the computability status of the Brouwer Fixed Point Theorem 13 was dependent on the dimension $k$ of the underlying space $[0,1]^{k}$, the computability status of the Hahn-Banach Theorem 18 is dependent on the dimension and other aspects of the space $X$ [10]. Nerode and Metakides [32] observed that for finite-dimensional $X$ no counterexample as in Proposition 19 exists. However, even in this case the theorem is not uniformly computable [10]. Under all conditions that guarantee that the extension is uniquely determined, the HahnBanach Theorem is fully computable; this includes for instance all computable Hilbert spaces [10].

A number of further theorems from functional analysis were analyzed by the author of this article and these include the Open Mapping Theorem, the Closed Graph Theorem and Banach's Inverse Mapping Theorem [7,11]. Another theorem that falls into this category is the Uniform Boundedness Theorem [9]. These examples are interesting, since they behave differently from all aforementioned examples. We illustrate the situation using Banach's Inverse Mapping Theorem.

Theorem 20 (Banach's Inverse Mapping Theorem). If $T: X \rightarrow Y$ is a bijective, linear and bounded operator on Banach spaces $X, Y$, then its inverse $T^{-1}: Y \rightarrow X$ is bounded too.

Here we obtain the following computable version [11].
Proposition 21 (B. 2009). If $T: X \rightarrow Y$ is a computable, bijective and linear operator on computable Banach spaces $X, Y$, then its inverse $T^{-1}: Y \rightarrow X$ is computable too.

That is, every bijective and linear operator $T$ with a program admits also a program for its inverse $T^{-1}$, but there is not general method to compute such a program for $T^{-1}$ from a program for $T$ in general as the following result shows [11].

Proposition 22 (B. 2009). Inversion $\mathrm{BIM}: \subseteq \mathcal{C}\left(\ell_{2}, \ell_{2}\right) \rightarrow \mathcal{C}\left(\ell_{2}, \ell_{2}\right), T \mapsto T^{-1}$ restricted to bijective, linear and bounded $T: \ell_{2} \rightarrow \ell_{2}$ is not computable (and not even continuous).

Analogously, there is also a sequential counterexample [11].

Proposition 23 (B. 2009). There exists a computable sequence $\left(T_{n}\right)_{n}$ of computable, bijective and linear operators $T_{n}: \ell_{2} \rightarrow \ell_{2}$ such that the sequence $\left(T_{n}^{-1}\right)_{n}$ of their inverses is not computable.

Hence, in several respects the Banach Inverse Mapping Theorem behaves similarly to the Intermediate Value Theorem: it is non-uniformly computable, but not uniformly computable. Yet we will see that the uniform content of both theorems is different.

With our final example of a theorem from functional analysis we want to close the circle and mention a result that behaves similarly to the Monotone Convergence Theorem 2, namely the Fréchet-Riesz Representation Theorem.

Theorem 24 (Fréchet-Riesz Representation Theorem). For every linear bounded functional $f: H \rightarrow \mathbb{R}$ on a Hilbert space $H$ there exists a unique $y \in H$ such that $f=f_{y}$ and $\|f\|=\|y\|$, where $f_{y}: H \rightarrow \mathbb{R}, x \mapsto\langle x, y\rangle$.

Here $\langle$,$\rangle denotes the inner product of the Hilbert space H$. For every computable $y \in H$ the functional $f_{y}$ is computable with norm $\left\|f_{y}\right\|=\|y\|$. Since the norm $\|y\|$ is always computable for a computable $y \in H$, it is immediately clear that it suffices to construct a computable functional $f: H \rightarrow \mathbb{R}$ without a computable norm $\|f\|$ in order to show that this theorem cannot hold computably. In fact, a Specker-like counterexample suffices in this case [20].

Proposition 25 (B. and Yoshikawa 2006). There exists a computable functional $f: \ell_{2} \rightarrow \mathbb{R}$ such that $\|f\|$ is not computable and hence there cannot be a computable $y \in H$ with $f=f_{y}$ and $\|f\|=\|y\|$.

## 4 A Classification Scheme for Theorems

If we look at the different examples of theorems that we have presented in the preceding sections then it becomes clear that theorems can behave quite differently with respect to computability. For one, the uniform and the non-uniform behavior can differ and the levels of computability can be of different complexities (computable, low, limit computable, etc.). On the other hand, certain theorems seem to be quite similar to each other, for instance the Monotone Convergence Theorem is similar in its behavior to the Fréchet-Riesz Theorem.

This naturally leads us to the question whether there is a classification scheme that allows to derive all sorts of computability properties of a theorem once it has been classified according to the corresponding scheme. The best known classification scheme for theorems in logic is Reverse Mathematics, i.e., the project to classify theorems in second order arithmetic according to certain axioms that are required to prove the corresponding theorem [40]. It turns out that this classification scheme is not fine enough for our purposes, because it only captures theorems in a non-uniform sense. In order to preserve computability properties such as lowness that are not closed under product, we also need a classification scheme that is more resource sensitive than reverse mathematics.

Such a classification scheme has been developed over the previous eight years using the concept of Weihrauch reducibility [12-16,18,21,35,36]. If $X, Y, Z, W$ are represented spaces, then $f: \subseteq X \rightrightarrows Y$ is Weihrauch reducible to $g: \subseteq$ $Z \rightrightarrows W$, if there are computable multi-valued functions $H, K$ such that $\emptyset \neq H(x, g K(x)) \subseteq f(x)$ for all $x \in \operatorname{dom}(f)$. In symbols we write $f \leq_{\mathrm{W}} g$ in this situation. If the reduction works in both directions, then we write $f \equiv_{\mathrm{W}} g$. It can be shown that this reducibility induces a lattice structure $[14,36]$.

Now a theorem of logical form $(\forall x \in X)(x \in D \Longrightarrow(\exists y \in Y) P(x, y))$ can be interpreted as a multi-valued function $f: \subseteq X \rightrightarrows Y, x \mapsto\{y \in Y: P(x, y)\}$ with $\operatorname{dom}(f)=D$. For instance, we obtain the following multi-valued functions for the theorems that we have considered in the previous sections (some of which are formulated in greater generality here). Here Tr denotes the set of binary trees and $[T]$ the set of infinite paths of a tree $T$.

- MCT $: \subseteq \mathbb{R}^{\mathbb{N}} \rightarrow \mathbb{R},\left(x_{n}\right)_{n} \mapsto \sup _{n \in \mathbb{N}} x_{n}$ restricted to monotone bounded sequences.
$-\mathrm{IVT}: \subseteq \mathcal{C}[0,1] \rightrightarrows[0,1], f \mapsto f^{-1}\{0\}$ with $\operatorname{dom}(\mathrm{IVT}):=\{f: f(0) \cdot f(1)<0\}$.
$-\operatorname{MAX}_{X}: \subseteq \mathcal{C}(X) \rightrightarrows \mathbb{R}, f \mapsto\{x \in X: f(x)=\max f(X)\}$ for computably compact ${ }^{1}$ computable metric spaces $X$ and, in particular, for $X=[0,1]$.
- WKL : $\subseteq \operatorname{Tr} \rightrightarrows 2^{\mathbb{N}}, T \mapsto[T]$ restricted to infinite binary trees.
$-\mathrm{BFT}_{n}: \mathcal{C}\left([0,1]^{n},[0,1]^{n}\right) \rightrightarrows[0,1]^{n}, f \mapsto\{x: f(x)=x\}$ for $n \geq 1$.
$-\mathrm{BWT}_{X}: \subseteq X^{\mathbb{N}} \rightrightarrows X,\left(x_{n}\right)_{n} \mapsto\left\{x: x\right.$ cluster point of $\left.\left(x_{n}\right)_{n}\right\}$, restricted to sequences that are in a compact subset of $X$.
- $\mathrm{BIM}_{X, Y}: \subseteq \mathcal{C}(X, Y) \rightarrow \mathcal{C}(Y, X), T \mapsto T^{-1}$, restricted to bijective, linear, bounded $T$ and for computable Banach spaces $X, Y$.
- $\mathrm{FRR}_{H}: \subseteq \mathcal{C}(H) \rightarrow H, f_{y} \mapsto y$ for computable Hilbert spaces $H$.
$-\mathrm{Z}_{X}: \subseteq \mathcal{C}(X) \rightarrow \mathbb{R}, f \mapsto f^{-1}\{0\}$ for computable metric spaces $X$.
We have not formalized the Hahn-Banach Theorem here and point the reader to [21]. The last mentioned problem $\mathrm{Z}_{X}$ is the zero problem, which is the problem to find a zero of a continuous function that admits at least one zero. By [19, Theorem 3.10] we obtain $Z_{X} \equiv{ }_{W} C_{X}$ for the choice problem of every computable metric space. We are not going to define $C_{X}$ here, but whenever we use it we will actually take $Z_{X}$ as a substitute for it. The following equivalences were proved in [13]:

Theorem 26 (Choice on the natural numbers). The following are all Weihrauch equivalent to each other and complete among functions that are computable with finitely many mind changes [12]:

1. Choice on natural numbers $\mathrm{C}_{\mathbb{N}}$.
2. The Baire Category Theorem (in appropriate formulation).
3. Banach's Inverse Mapping Theorem $\mathrm{BIM}_{\ell_{2}, \ell_{2}}$.
4. The Open Mapping Theorem for $\ell_{2}$.

[^5]5. The Closed Graph Theorem for $\ell_{2}$.
6. The Uniform Boundedness Theorem for $\ell_{2}$.

Hence the equivalence class of choice $\mathrm{C}_{\mathbb{N}}$ on the natural numbers contains many theorems that are typically proved with and closely related to the Baire Category Theorem.

We prove that the Theorem of the Maximum is equivalent to the zero problem of $[0,1]$.

Theorem 27. $\mathrm{MAX}_{X} \equiv_{\mathrm{W}} \mathrm{Z}_{X}$ for every computably compact computable metric space $X$.

Proof. We prove $\mathrm{Z}_{X} \leq_{\mathrm{W}} \mathrm{MAX}_{X}$. Given a continuous function $f: X \rightarrow \mathbb{R}$ with $A=f^{-1}\{0\} \neq \emptyset$, we can compute the function $g: X \rightarrow \mathbb{R}$ with $g:=-|f|$. Then $\operatorname{MAX}(g)=f^{-1}\{0\}=A$. This proves the claim. We now prove $\mathrm{MAX}_{X} \leq_{\mathrm{W}} \mathrm{Z}_{X}$. Given a continuous function $f: X \rightarrow \mathbb{R}$ with $\operatorname{MAX}(f)=A \neq \emptyset$, we can compute $g: X \rightarrow \mathbb{R}$ with $g:=f-\max f(X)$, since $X$ is computably compact. Now we obtain $g^{-1}\{0\}=\operatorname{MAX}(f)=A$. This proves the claim.

We now arrive at the following result that is compiled from different sources. It shows that the equivalence class of choice on Cantor space contains several problems whose non-computably was proved with the help of Weak Kőnig's Lemma or with the help of two c.e. sets that are computably inseparable. We point out that the sequential version of the Intermediate Valued Theorem formulated in Proposition 6 can be modeled by parallelization. For $f: \subseteq X \rightrightarrows Y$ we define its parallelization $\widehat{f}: \subseteq X^{\mathbb{N}} \rightrightarrows Y^{\mathbb{N}},\left(x_{n}\right)_{n} \mapsto \times_{n=0}^{\infty} f\left(x_{n}\right)$, which lifts $f$ to sequences. Parallelization is a closure operation in the Weihrauch lattice [14].
Theorem 28 (Choice on Cantor space). The following are all Weihrauch equivalent to each other and complete among non-deterministically computable functions with a binary sequence as advice [12]:

1. Choice on Cantor Space $\mathrm{C}_{2^{\mathrm{N}}}$.
2. Weak Kőnig's Lemma WKL [14,21].
3. The Theorem of the Maximum $\mathrm{MAX}_{[0,1]}$ (Theorem 27).
4. The Hahn-Banach Theorem (Gherardi and Marcone 2009) [21].
5. The parallelization $\widehat{\mathrm{VT}}$ of the Intermediate Value Theorem [13].
6. The Brouwer Fixed Point Theorem $\mathrm{BFT}_{n}$ for dimension $n \geq 2$ [18].

We note that [18] contains the proof for the Brouwer Fixed Point Theorem only for dimension $n \geq 3$ and the results for $n=2$ is due to Joseph Miller. It is easy to see that IVT $\equiv_{\mathrm{W}} \mathrm{BFT}_{1}[18]$. We mention that this result implies Propositions $6,9,11,14$ and 19 and constitutes a more general uniform classification. In some cases the proofs can easily be derived from known techniques and results, in other cases (for instance for the Hahn-Banach Theorem and the Brouwer Fixed Point Theorem) completely new techniques are required. One can prove that the equivalence classes appearing in Theorems 26 and 28 are incomparable [13]. The next equivalence class that we are going to discuss is an upper bound of both.

We first prove that the Monotone Convergence Theorem MCT is equivalent to the Fréchet-Riesz Representation Theorem $\mathrm{FRR}_{H}$.

Theorem 29. $\mathrm{FRR}_{H} \equiv_{\mathrm{W}} \mathrm{MCT}$ for every computable infinite-dimensional Hilbert space $H$.

Proof. Since every infinite-dimensional computable Hilbert space is computably isometrically isomorphic to $\ell_{2}$ by [20, Corollary 3.7], if suffices to consider $H=$ $\ell_{2}$. We first prove $\mathrm{FRR}_{\ell_{2}} \leq_{W}$ MCT. Given a functional $f: \ell_{2} \rightarrow \mathbb{R}$ we need to find a $y \in \ell_{2}$ such that $f_{y}=f$ and $\|y\|=\|f\|$. There is a computable sequence $\left(x_{n}\right)_{n}$ in $\ell_{2}$ such that $\overline{\left\{x_{n}: n \in \mathbb{N}\right\}}$ is dense in $\left\{x \in \ell_{2}:\|x\| \leq 1\right\}$. Hence $\|f\|=\sup _{\| x| | \leq 1}|f(x)|=\sup _{n \in \mathbb{N}}\left|f\left(x_{n}\right)\right|=\sup _{n \in \mathbb{N}} \max _{i \leq n}\left|f\left(x_{i}\right)\right|$ and hence we can compute $\|f\|$ with the help of MCT. Now given $f=f_{y}$ and $\|f\|$ we can easily compute $y$ be evaluating $f_{y}\left(e_{n}\right)$ on the unit vectors $e_{n}$. We still need to prove $\mathrm{MCT} \leq_{\mathrm{W}} \mathrm{FRR}_{\ell_{2}}$. By [8, Proposition 9.1] it suffices to show that we can utilize $\mathrm{FRR}_{\ell_{2}}$ to translate enumerations $g$ of sets $A \subseteq \mathbb{N}$ into their characteristic functions. Let us assume that $A=\{n: n+1 \in$ range $(g)\}$. Without loss of generality we can assume that no value different from zero appears twice in $(g(n))_{n}$. Using the idea of [20, Example 4.6] we choose $a_{k}:=2^{g(k)-1}$ if $g(k) \neq 0$ and $a_{k}:=0$ otherwise. Then $a=\left(a_{k}\right)_{k} \in \ell_{2}$ and we can compute $f \in \mathcal{C}\left(\ell_{2}\right)$ with $f(x):=\sum_{k=0}^{\infty} x_{k} a_{k}=\langle x, a\rangle$. Now, with the help of $\mathrm{FRR}_{\ell_{2}}$ we obtain a $y \in \ell_{2}$ with $\|y\|=\|f\|=\|a\|=\sqrt{\sum_{k=0}^{\infty}\left|a_{k}\right|^{2}}$. But using the number $\|y\|^{2}$ we can decide $A$, since its binary representation has in the even positions the characteristic function of $A$.

We note that $\mathrm{FRR}_{H}$ for finite-dimensional spaces $H$ is computable. Altogether we obtain the following result for this equivalence class.

Theorem 30 (The limit). The following are all Weihrauch equivalent to each other and complete for limit computable functions:

1. The limit map lim on Baire space [8].
2. The parallelization $\widehat{\mathrm{C}_{\mathbb{N}}}$ of choice on the natural numbers [13].
3. The parallelization $\widehat{\mathrm{BIM}}$ of Banach's Inverse Mapping Theorem [13].
4. The Monotone Convergence Theorem MCT [16].
5. The Fréchet-Riesz Representation Theorem FRR for $\ell_{2}$ (Theorem 29).
6. The Radon-Nikodym Theorem (Hoyrup, Rojas, Weihrauch 2012) [24].

This theorem implies Propositions 3, 23 and 25. Finally, we mention that we also have a concept of a jump $f^{\prime}: \subseteq X \rightrightarrows Y$ for every $f: \subseteq X \rightrightarrows Y$, which essentially replaces the input representation of $X$ in such a way that a name of $x \in X$ for $f^{\prime}$ is a sequence that converges to a name in the sense of $f$. This makes problems potentially more complicated since less input information is available. It allows us to phrase results as the following [16].

Theorem 31 (B., Gherardi and Marcone 2012). $\mathrm{WKL}^{\prime} \equiv_{W} B W T_{\mathbb{R}}$.
This result does not only imply Proposition 17, but also the following [16].
Corollary 32 (B., Gherardi and Marcone 2012). Every computable sequence $\left(x_{n}\right)_{n}$ in the unit cube $[0,1]^{n}$ has a cluster point $x \in[0,1]^{n}$ that is low relative to the halting problem.

Here $x$ is low relative to the halting problem if $x^{\prime} \leq_{\mathrm{T}} \emptyset^{\prime \prime}$ (some authors would only call this a partial relativization of lowness). In light of Proposition 17 this is one of the strongest positive properties that one can expect for a cluster point. These examples demonstrate that a classification of the Weihrauch degree of a theorem yields a large variety of computability properties of the theorem, uniform and non-uniform ones on the one hand, and positive and negative ones on the other hand.

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# The Brouwer Fixed Point Theorem Revisited 

Vasco Brattka ${ }^{1,2}$, Stéphane Le Roux ${ }^{3}$, Joseph S. Miller ${ }^{4}$, and Arno Pauly ${ }^{3(\boxtimes)}$<br>${ }^{1}$ Faculty of Computer Science, University of the Armed Forces Munich, Neubiberg, Germany<br>${ }^{2}$ Department of Mathematics and Applied Mathematics, University of Cape Town, Cape Town, South Africa<br>Vasco.Brattka@cca-net.de<br>${ }^{3}$ Département d'Informatique, Université libre de Bruxelles, Brussels, Belgium<br>Stephane.Le.Roux@ulb.ac.be, Arno.Pauly@cl.cam.ac.uk<br>${ }^{4}$ Department of Mathematics, University of Wisconsin-Madison, Madison, USA<br>jmiller@math.wisc.edu


#### Abstract

We revisit the investigation of the computational content of the Brouwer Fixed Point Theorem in [7], and answer the two open questions from that work. First, we show that the computational hardness is independent of the dimension, as long as it is greater than 1 (in [7] this was only established for dimension greater than 2 ). Second, we show that restricting the Brouwer Fixed Point Theorem to $L$-Lipschitz functions for any $L>1$ also does not change the computational strength, which together with prior results establishes a trichotomy for $L>1$, $L=1$ and $L<1$.


## 1 Introduction

In this paper we continue with the programme to classify the computational content of mathematical theorems in the Weihrauch lattice (see $[3-5,8,11,13$, $17,18])$. This lattice is induced by Weihrauch reducibility, which is a reducibility for partial multi-valued functions $f: \subseteq X \rightrightarrows Y$ on represented spaces $X, Y$. Intuitively, $f \leq_{\mathrm{W}} g$ reflects the fact that the function $f$ can be realized with a single application of the function $g$ as an oracle. Hence, if two functions are equivalent in the sense that they are mutually reducible to each other, then they are equivalent as computational resources, as far as computability is concerned.

Many theorems in mathematics are actually of the logical form

$$
(\forall x \in X)(\exists y \in Y) P(x, y)
$$

and such theorems can straightforwardly be represented by a multi-valued function $f: X \rightrightarrows Y$ with $f(x):=\{y \in Y: P(x, y)\}$ (sometimes partial $f$ are needed,

[^6]where the domain captures additional requirements that this input $x$ has to satisfy). In some sense the multi-valued function $f$ directly reflects the computational task of the theorem to find some suitable $y$ for any $x$. Hence, in a very natural way the classification of a theorem can be achieved via a classification of the corresponding multi-valued function that represents the theorem. In this paper we attempt to classify the Brouwer Fixed Point Theorem.

Theorem 1 (Brouwer Fixed Point Theorem 1911). Every continuous function $f:[0,1]^{n} \rightarrow[0,1]^{n}$ has a fixed point $x \in[0,1]^{n}$.

The fact that Brouwer's Fixed Point Theorem cannot be proved constructively has been confirmed in many different ways; most relevant for us is the counterexample in Russian constructive analysis by Orevkov [16], which was transferred into computable analysis by Baigger [1].

Constructions similar to those used for the above counterexamples have been utilized in order to prove that the Brouwer Fixed Point Theorem is equivalent to Weak Kőnig's Lemma in reverse mathematics $[21,22]$ and to analyze computability properties of fixable sets [14], but a careful analysis of these reductions reveals that none of them can be straightforwardly transferred into a uniform reduction in the sense that we are seeking here. The results cited above essentially characterize the complexity of fixed points themselves, whereas we want to characterize the complexity of finding the fixed point, given the function. This requires full uniformity.

In the Weihrauch lattice the Brouwer Fixed Point Theorem of dimension $n$ is represented by the multi-valued function $\mathrm{BFT}_{n}: \mathcal{C}\left([0,1]^{n},[0,1]^{n}\right) \rightrightarrows[0,1]^{n}$ that maps any continuous function $f:[0,1]^{n} \rightarrow[0,1]^{n}$ to the set of its fixed points $\mathrm{BFT}_{n}(f) \subseteq[0,1]^{n}$. The question now is where $\mathrm{BFT}_{n}$ is located in the Weihrauch lattice?

In order to approach this question, we introduce a choice principle $\mathrm{CC}_{n}$ that we call connected choice and which is just the closed choice operation restricted to connected subsets. That is, in the sense discussed above, $\mathrm{CC}_{n}$ is the multivalued function that represents the following mathematical statement: every non-empty connected closed set $A \subseteq[0,1]^{n}$ has a point $x \in A$. Since closed sets are represented by negative information (i.e. by an enumeration of open balls that exhaust the complement), the computational task of $\mathrm{CC}_{n}$ consists in finding a point in a closed set $A \subseteq[0,1]^{n}$ that is promised to be non-empty and connected and that is given by negative information.

One of our main results, presented in Sect.3, is that the Brouwer Fixed Point Theorem is equivalent to connected choice for each fixed dimension $n$, i.e. $\mathrm{BFT}_{n} \equiv{ }_{\mathrm{W}} \mathrm{CC}_{n}$. This result allows us to study the Brouwer Fixed Point Theorem in terms of the function $\mathrm{CC}_{n}$ that is easier to handle since it involves neither function spaces nor fixed points. This is also another instance of the observation that several important theorems are equivalent to certain choice principles (see [3]) and many important classes of computable functions can be calibrated in terms of choice (see [2]). For instance, closed choice on Cantor space $C_{\{0,1\}^{N}}$ and on the unit cube $\mathrm{C}_{[0,1]^{n}}$ are both easily seen to be equivalent to Weak

Kőnig's Lemma WKL, i.e. $\mathrm{WKL} \equiv{ }_{\mathrm{W}} \mathrm{C}_{\{0,1\}^{\mathrm{N}}} \equiv{ }_{\mathrm{W}} \mathrm{C}_{[0,1]^{n}}$ for any $n \geq 1$. Studying the Brouwer Fixed Point Theorem in the form of $\mathrm{CC}_{n}$ now amounts to comparing $\mathrm{C}_{[0,1]^{n}}$ with its restriction $\mathrm{CC}_{n}$.

Our second main result, given in Sect. 5, is that from dimension two onwards connected choice is equivalent to Weak Kőnig's Lemma, i.e. $\mathrm{CC}_{n} \equiv_{\mathrm{W}} \mathrm{C}_{[0,1]}$ for $n \geq 2$.

This refutes an earlier conjecture [7] by some of the authors that connected choice in dimension two be computationally simpler than connected choice in dimension three. We then also consider the restriction of Brouwer's Fixed Point theorem to Lipschitz functions in Sect.4. In the following Sect. 2 we start with a short summary of relevant definitions and results regarding the Weihrauch lattice.

This extended abstract does not contain any proofs. Sections 1, 2 and 3 are taken mostly from [7]. An extended version including the omitted proofs can be found as [6].

## 2 The Weihrauch Lattice

In this section we briefly recall some basic results and definitions regarding the Weihrauch lattice. The original definition of Weihrauch reducibility is due to Weihrauch and has been studied for many years (see [9,23-25]). Only recently it has been noticed that a certain variant of this reducibility yields a lattice that is very suitable for the classification of mathematical theorems (see [2$5,8,10,17,18]$ ). The basic reference for all notions from computable analysis is [26], alternatively see [19].

The Weihrauch lattice is a lattice of multi-valued functions on represented spaces. A representation $\delta$ of a set $X$ is just a surjective partial map $\delta: \subseteq \mathbb{N}^{\mathbb{N}} \rightarrow$ $X$. In this situation we call $(X, \delta)$ a represented space. In general we use the symbol " $\subseteq$ " in order to indicate that a function is potentially partial. Using represented spaces we can define the concept of a realizer. We denote the composition of two (multi-valued) functions $f$ and $g$ either by $f \circ g$ or by $f g$.

Definition 1 (Realizer). Let $f: \subseteq\left(X, \delta_{X}\right) \rightrightarrows\left(Y, \delta_{Y}\right)$ be a multi-valued function on represented spaces. A function $F: \subseteq \mathbb{N}^{\mathbb{N}} \rightarrow \mathbb{N}^{\mathbb{N}}$ is called a realizer of $f$, in symbols $F \vdash f$, if $\delta_{Y} F(p) \in f \delta_{X}(p)$ for all $p \in \operatorname{dom}\left(f \delta_{X}\right)$.

Realizers allow us to transfer the notions of computability and continuity and other notions available for Baire space to any represented space; a function between represented spaces will be called computable if it has a computable realizer, etc. Now we can define Weihrauch reducibility.

Definition 2 (Weihrauch reducibility). Let $f, g$ be multi-valued functions on represented spaces. Then $f$ is said to be Weihrauch reducible to $g$, in symbols $f \leq_{\mathrm{w}} g$, if there are computable functions $K, H: \subseteq \mathbb{N}^{\mathbb{N}} \rightarrow \mathbb{N}^{\mathbb{N}}$ such that $K\langle\mathrm{id}, G H\rangle \vdash f$ for all $G \vdash g$. Moreover, $f$ is said to be strongly Weihrauch reducible to $g$, in symbols $f \leq_{\mathrm{sW}} g$, if there are computable functions $K, H$ such that $K G H \vdash f$ for all $G \vdash g$.

Here $\langle$,$\rangle denotes some standard pairing on Baire space. We note that the$ relations $\leq_{W}, \leq_{\mathrm{sW}}$ and $\vdash$ implicitly refer to the underlying representations, which we mention explicitly only when necessary. It is known that these relations only depend on the underlying equivalence classes of representations, but not on the specific representatives (see Lemma 2.11 in [4]). We use $\equiv_{\mathrm{W}}$ and $\equiv_{\mathrm{sW}}$ to denote the respective equivalences regarding $\leq_{\mathrm{W}}$ and $\leq_{\mathrm{sW}}$, and by $<_{\mathrm{W}}$ and $<_{\text {sW }}$ we denote strict reducibility.

A particularly useful multi-valued function in the Weihrauch lattice is closed choice (see $[2-4,8]$ ) and it is known that many notions of computability can be calibrated using the right version of choice. We will focus on closed choice for computable metric spaces, which are separable metric spaces such that the distance function is computable on the given dense subset. We assume that computable metric spaces are represented via their Cauchy representation (see [26] for details).

By $\mathcal{A}_{-}(X)$ we denote the set of closed subsets of a metric space $X$, where the index "-" indicates that we work with negative information. This information is given by a representation $\psi_{-}: \mathbb{N}^{\mathbb{N}} \rightarrow \mathcal{A}_{-}(X)$, defined by $\psi_{-}(p):=$ $X \backslash \bigcup_{i=0}^{\infty} B_{p(i)}$, where $B_{n}$ is some standard enumeration of the open balls of $X$ with center in the dense subset and rational radius. The computable points in $\mathcal{A}_{-}(X)$ are called co-c.e. closed sets. We now define closed choice for the case of computable metric spaces.

Definition 3 (Closed Choice). Let $\mathbf{X}$ be a computable metric space. Then the closed choice operation $\mathrm{C}_{\mathbf{X}}: \subseteq \mathcal{A}_{-}(X) \rightrightarrows X$ of this space is defined by $\operatorname{dom}\left(\mathrm{C}_{X}\right):=\left\{A \in \mathcal{A}_{-}(X): A \neq \emptyset\right\}$ and $x \in \mathrm{C}_{\mathbf{X}}(A)$ iff $x \in A$.

Intuitively, $\mathrm{C}_{X}$ takes as input a non-empty closed set in negative representation (i.e. given by $\psi_{-}$) and it produces an arbitrary point of this set as output. For short we use the notation $\mathcal{A}_{n}:=\left\{A \in \mathcal{A}_{-}\left([0,1]^{n}\right): A \neq \emptyset\right\}$ for the space of non-empty closed subsets with representation $\psi_{-}$in the following.

## 3 Brouwer's Fixed Point Theorem and Connected Choice

In this section we want to show that the Brouwer Fixed Point Theorem is computably equivalent to connected choice for any fixed dimension. We first define these two operations. By $\mathcal{C}(X, Y)$ we denote the set of continuous functions $f: X \rightarrow Y$ and for short we write $\mathcal{C}_{n}:=\mathcal{C}\left([0,1]^{n},[0,1]^{n}\right)$.

Definition 4 (Brouwer Fixed Point Theorem). By $\mathrm{BFT}_{n}: \mathcal{C}_{n} \rightrightarrows[0,1]^{n}$ we denote the operation defined by $\mathrm{BFT}_{n}(f):=\left\{x \in[0,1]^{n}: f(x)=x\right\}$ for $n \in \mathbb{N}$.

We note that $\mathrm{BFT}_{n}$ is well-defined, i.e. $\mathrm{BFT}_{n}(f)$ is non-empty for all $f$, since by the Brouwer Fixed Point Theorem every $f \in \mathcal{C}_{n}$ admits a fixed point $x$, i.e. with $f(x)=x$. We now define connected choice.

Definition 5 (Connected choice). $B y C C_{n}: \subseteq \mathcal{A}_{n} \rightrightarrows[0,1]^{n}$ we denote the operation defined by $\mathrm{CC}_{n}(A):=A$ for all non-empty connected closed $A \subseteq[0,1]^{n}$ and $n \in \mathbb{N}$. We call $\mathrm{CC}_{n}$ connected choice (of dimension $n$ ).

Hence, connected choice is just the restriction of closed choice $\mathrm{C}_{[0,1]^{n}}$ to connected sets. We also use the following notation for the set of fixed points of a function $f \in \mathcal{C}_{n}$.

Definition 6 (Set of fixed points). By $\mathrm{Fix}_{n}: \mathcal{C}_{n} \rightarrow \mathcal{A}_{n}$ we denote the function with $\operatorname{Fix}_{n}(f):=\left\{x \in[0,1]^{n}: f(x)=x\right\}$.

It is easy to see that $\mathrm{Fix}_{n}$ is computable, since $\operatorname{Fix}_{n}(f):=(f-\mathrm{id})^{-1}\{0\}$ and it is well-known that closed sets in $\mathcal{A}_{n}$ can also be represented as zero sets of continuous functions (see [26]).

Definition 7 (Connectedness components). By $\operatorname{Con}_{n}: \mathcal{A}_{n} \rightrightarrows \mathcal{A}_{n}$ we denote the map with $\operatorname{Con}_{n}(A):=\{C: C$ is a connectedness component of $A\}$ for every $n \geq 1$.

Theorem 2 (Connectedness components). Con $_{n} \equiv_{\mathrm{sW}} \mathrm{WKL}$ for $n \geq 1$.
We note that the Brouwer Fixed Point Theorem can be decomposed to $\mathrm{BFT}_{n}=\mathrm{CC}_{n} \circ \mathrm{Con}_{n} \circ \mathrm{Fix}_{n}$.

The main result of this section will be that the Brouwer Fixed Point Theorem and connected choice are (strongly) equivalent for any fixed dimension $n$ (see Theorem 3 below).

The direction $\mathrm{CC}_{n} \leq_{s W} \mathrm{BFT}_{n}$ can be seen as a uniformization of an earlier construction of Baigger [1] that is in turn built on results of Orevkov [16]. This part of the construction was explained in some detail by Potgieter in [20].

For the other direction $\mathrm{BFT}_{n} \leq_{s \mathrm{w}} \mathrm{CC}_{n}$ of the reduction we uniformize ideas from the third author's PhD thesis [14]. A central technique is topological degree theory. For the uniform aspects of both directions, a representation of closed sets via trees of rational complexes is employed.

The first observation is that the map $\mathrm{Con}_{n} \circ \mathrm{Fix}_{n}$ is computable (which might be surprising in light of Theorem 2).

Proposition 1. $\operatorname{Con}_{n} \circ \operatorname{Fix}_{n}: \mathcal{C}_{n} \rightrightarrows \mathcal{A}_{n}$ is computable for all $n \in \mathbb{N}$.
Since $\mathrm{BFT}_{n} \supseteq \mathrm{CC}_{n} \circ \mathrm{Con}_{n} \circ \mathrm{Fix}_{n}$ we can directly conclude $\mathrm{BFT}_{n} \leq_{s W} \mathrm{CC}_{n}$ for all $n$. Together with $\mathrm{CC}_{n} \leq_{s W} \mathrm{BFT}_{n}$ we obtain the following theorem.

Theorem 3 (Brouwer Fixed Point Theorem). $\mathrm{BFT}_{n} \equiv_{\mathrm{sW}} \mathrm{CC}_{n}$ for all $n$.
It is easy to see that in general the Brouwer Fixed Point Theorem and connected choice are not independent of the dimension. In case of $n=0$ the space $[0,1]^{n}$ is the one-point space $\{0\}$ and hence $\mathrm{BFT}_{0} \equiv_{\text {sW }} \mathrm{CC}_{0}$ are both computable. In case of $n=1$ connected choice was already studied in [3] and it was proved that it is equivalent to the Intermediate Value Theorem IVT (see Definition 6.1 and Theorem 6.2 in [3]).

Corollary 1 (Intermediate Value Theorem). IVT $\equiv_{\mathrm{sW}} \mathrm{BFT}_{1} \equiv_{\mathrm{sW}} \mathrm{CC}_{1}$.
It is also easy to see that the Brouwer Fixed Point Theorem $\mathrm{BFT}_{2}$ in dimension two is more complicated than in dimension one. For instance, it is known that the Intermediate Value Theorem IVT always offers a computable function value for a computable input, whereas this is not the case for the Brouwer Fixed Point Theorem $\mathrm{BFT}_{2}$ by Baigger's counterexample [1]. We continue to discuss this topic in Sect. 5.

Here we point out that Proposition 1 implies that the fixed point set Fix ${ }_{n}(f)$ of every computable function $f:[0,1]^{n} \rightarrow[0,1]^{n}$ has a co-c.e. closed connectedness component. The converse direction is true, too, and in a uniform way: We denote by $(f, g): \subseteq X \rightrightarrows Y \times Z$ the juxtaposition of two functions $f: \subseteq X \rightrightarrows Y$ and $g: \subseteq X \rightrightarrows Z$, defined by $(f, g)(x)=(f(x), g(x))$.

Theorem 4 (Fixability). ( $\mathrm{Fix}_{n}, \mathrm{Con}_{n} \circ \mathrm{Fix}_{n}$ ) is computable and has a multivalued computable right inverse for all $n \in \mathbb{N}$.

Roughly speaking a closed set $A \in \mathcal{A}_{n}$ together with one of its connectedness components is as good as a continuous function $f \in \mathcal{C}_{n}$ with $A$ as set of fixed points. As a non-uniform corollary we obtain immediately Miller's original result.

Corollary 2 (Fixable sets, Miller 2002). A set $A \subseteq[0,1]^{n}$ is the set of fixed points of a computable function $f:[0,1]^{n} \rightarrow[0,1]^{n}$ if and only if it is non-empty and co-c.e. closed and contains a co-c.e. closed connectedness component.

## 4 The Lipschitz Trichotomy

It seems to be a natural question ${ }^{1}$ to what extent finding fixed points becomes easier if the class of functions to be considered is further restricted. In particular we will denote by $L-\mathrm{LBFT}_{n}$ the restriction of $\mathrm{BFT}_{n}$ to $L$-Lipschitz functions.

Proposition 2. For $L_{1}, L_{2}>1$ we find that $L_{1}-\mathrm{LBFT}_{n} \equiv_{\mathrm{W}} L_{2}-\mathrm{LBFT}_{n}$.
Proof. If $f$ is $L_{1}$-Lipschitz and $L_{2}>1$, then $\mathrm{id}+\frac{L_{2}-1}{L_{1}+1}(f-\mathrm{id})$ is $L_{2}$-Lipschitz and has the same fixed points as $f$.

With some additional constructions and a careful analysis, the proof of Theorem 3 can be adapted to yield:

Theorem 5. $2-\mathrm{LBFT}_{n} \equiv_{\mathrm{W}} \mathrm{BFT}_{n} \equiv_{\mathrm{W}} \mathrm{CC}_{n}$.
Being $L$-Lipschitz for $L<1$ implies the uniqueness of the fixed point, which in turn implies the computability of $L-\mathrm{LBFT}_{n}$ for $L<1$. The remaining $L=1$ case is also a special (since finite-dimensional) case of the Browder-Goehde-Kirk Fixed Point theorem. Its Weihrauch degree was studied by Neumann in [15], and shown to be equivalent to $\mathrm{XC}_{n}$ - closed choice for convex sets in $[0,1]^{n}$.

[^7]Theorem 6 (Le Roux and Pauly [12]).

$$
\mathrm{CC}_{1} \equiv_{\mathrm{W}} \mathrm{XC}_{1}<_{\mathrm{W}} \mathrm{XC}_{2}<_{\mathrm{W}} \mathrm{XC}_{3}<\mathrm{W} \ldots<{ }_{\mathrm{W}} \mathrm{C}_{[0,1]}
$$

Corollary 3 (Lipschitz dichotomy in dimension 1).

- $L-$ LBFT $_{1} \equiv_{\mathrm{W}}$ id, iff $L<1$
$-L-\mathrm{LBFT}_{1} \equiv_{\mathrm{W}} \mathrm{CC}_{1}$, iff $L \geq 1$
Corollary 4 (Lipschitz trichotomy). Let $n>1$.
$-L-$ LBFT $_{n} \equiv_{\mathrm{W}}$ id, iff $L<1$
$-L-\mathrm{LBFT}_{n} \equiv_{\mathrm{W}} \mathrm{XC}_{n}$, iff $L=1$
$-L-\mathrm{LBFT}_{n} \equiv_{\mathrm{W}} \mathrm{C}_{[0,1]}$, iff $L>1$


## 5 Classifying Connected Choice

In this section we want to discuss the degree of connected choice, in particular in relation to the dimension of the ambient space. We will consider three geometric constructions: The one employed in the original proof by Orevkov/Baigger - this construction is insufficient for the uniform aspects. Then a simple construction showing that connected choice is computably complete from dimension three onwards in the sense that it is strongly equivalent to Weak Kőnig's Lemma. Finally, a significantly more involved construction shows even connected choice in two dimensions to be computably complete, too.

A superficial reading of the results of Orevkov [16] and Baigger [1] can lead to the wrong conclusion that they actually provide a reduction of Weak Kőnig's Lemma to the Brouwer Fixed Point Theorem $\mathrm{BFT}_{n}$ of any dimension $n \geq 2$. However, this is only correct in a non-uniform way and the corresponding uniform result does not follow from the known constructions. The Orevkov-Baigger result is built on the following fact.

Proposition 3 (Mixed cube). The function $M: \subseteq \mathcal{A}_{-}[0,1] \rightarrow \mathcal{A}_{2}$ with $M(A)=(A \times[0,1]) \cup([0,1] \times A)$ is computable and maps non-empty closed sets $A \subseteq[0,1]$ to non-empty connected closed sets $M(A) \subseteq[0,1]^{2}$.

It follows straightforwardly from the definition that the pairs $(x, y) \in M(A)$ are such that one out of two components $x, y$ is actually in $A$. In order to express the uniform content of this fact, we introduce the concept of a fraction.

Definition 8 (Fractions). Let $f: \subseteq X \rightrightarrows Y$ be a multi-valued function and $0<n \leq m \in \mathbb{N}$. We define the fraction $\frac{n}{m} f: \subseteq X \rightrightarrows Y^{m}$ such that $\frac{n}{m} f(x)$ is the set of all $\left(y_{1}, \ldots, y_{m}\right) \in$ range $(f)^{m}$ with $\left|\left\{i: y_{i} \in f(x)\right\}\right| \geq n$ for all $x \in \operatorname{dom}\left(\frac{n}{m} f\right):=\operatorname{dom}(f)$.

The idea of a fraction $\frac{n}{m} f$ is that it provides $m$ potential answers for $f$, at least $n \leq m$ of which have to be correct. The uniform content of the OrevkovBaigger construction is then summarized in the following result.

Proposition 4 (Dimension two). $\frac{1}{2} \mathrm{C}_{[0,1]} \leq_{\mathrm{sW}} \mathrm{CC}_{2}$.
However, the following results shows that the uniform content of the preceding proposition is very weak, as it cannot even solve closed choice on the two-point space 2 (which is equivalent to LLPO):

Proposition 5. $C_{2} \not \leq W ~ \frac{1}{2} C_{[0,1]}$
That is, given a closed set $A \subseteq[0,1]$ we can utilize connected choice $\mathrm{CC}_{2}$ of dimension 2 in order to find a pair of points $(x, y)$ one of which is in $A$. This result directly implies the counterexample of Baigger [1] because the fact that there are non-empty co-c.e. closed sets $A \subseteq[0,1]$ without computable points immediately implies that $\frac{1}{2} \mathrm{C}_{[0,1]}$ is not non-uniformly computable (i.e. there are computable inputs without computable outputs) and hence $\mathrm{CC}_{2}$ is also not non-uniformly computable.

Corollary 5 (Orevkov 1963, Baigger 1985). There exists a computable function $f:[0,1]^{2} \rightarrow[0,1]^{2}$ that has no computable fixed point $x \in[0,1]^{2}$. There exists a non-empty connected co-c.e. closed subset $A \subseteq[0,1]^{2}$ without computable point.

Instead, we shall use a different construction to classify connected choice from three dimensions upwards:

Proposition 6 (Twisted cube). The function $T: \subseteq \mathcal{A}_{-}[0,1] \rightarrow \mathcal{A}_{3}$ with $T(A)=(A \times[0,1] \times\{0\}) \cup(A \times A \times[0,1]) \cup([0,1] \times A \times\{1\})$ is computable and maps non-empty closed sets $A \subseteq[0,1]$ to non-empty connected closed sets $T(A) \subseteq[0,1]^{3}$.

Here tuples $\left(x_{1}, x_{2}, x_{3}\right) \in T(A)$ have the property that at least one of the first two components provide a solution $x_{i} \in A$, but the third component provides the additional information which one surely does. If $x_{3}$ is close to 1 , then surely $x_{2} \in A$ and if $x_{3}$ is close to 0 , then surely $x_{1} \in A$. If $x_{3}$ is neither close to 0 nor 1 , then both $x_{1}, x_{2} \in A$. Hence, there is a computable function $H$ such that $\mathrm{C}_{[0,1]}=H \circ \mathrm{CC}_{3} \circ T$, which proves $\mathrm{C}_{[0,1]} \leq_{\mathrm{sW}} \mathrm{CC}_{3}$. Together with Theorem 3 we obtain the following conclusion.

Theorem 7 (Completeness of three dimensions). For $n \geq 3$ we obtain $\mathrm{CC}_{n} \equiv_{\mathrm{sW}} \mathrm{BFT}_{n} \equiv_{\mathrm{sW}} \mathrm{WKL} \equiv_{\mathrm{sW}} \mathrm{C}_{[0,1]}$.

We note that the reduction $\mathrm{CC}_{n} \leq_{\mathrm{sW}} \mathrm{C}_{[0,1]^{n}}$ holds for all $n \in \mathbb{N}$, since connected choice is just a restriction of closed choice and $\mathrm{C}_{[0,1]^{n}} \equiv_{\mathrm{sW}} \mathrm{C}_{[0,1]} \equiv_{\mathrm{sW}} \mathrm{WKL}$ is known for all $n \geq 1$ (see [2]).

Originally, three of the authors had conjectured in [7] that $\mathrm{CC}_{2}<{ }_{W} \mathrm{C}_{[0,1]}$. However, a more involved construction actually establishes that:


Fig. 1. The geometric pattern after the third round

Theorem 8 (Completeness of two dimensions). $\mathrm{CC}_{2} \equiv{ }_{W} \mathrm{C}_{[0,1]}$
The proof of Theorem 8 exhibits a reduction $\widehat{\mathrm{C}_{\mathbf{2}}} \leq{ }_{\mathrm{W}} \mathrm{CC}_{2}$ instead, using the equivalence $\widehat{\mathrm{C}_{2}} \equiv{ }_{\mathrm{W}} \mathrm{C}_{\{0,1\}^{\mathrm{N}}}$ from [4]. The geometric pattern constructed produces an infinitely long line which is then subdivided based on both the information obtained about the input to $\widehat{C_{2}}$, as well as the order in which this information is found. A glimpse of the construction might be gained from Fig. 1.

## 6 Two Versus Three Dimensions

A noticeable difference between the construction from the proof of Theorem 8 and Proposition 6 is that the latter yields even a path-connected set, whereas the former does not. Thus, path-connected choice is computably-complete from dimension three onwards, but might be simpler in dimension two.

While the status of path-connected choice in dimension two remains open, we can exhibit a related choice principle distinguishing two from three dimensions.

Definition 9. We say that $A \in \mathcal{A}_{2}$ has a straight cross, if there are $x, y \in[0,1]$, $\delta>0$ s.t $\forall \varepsilon \in(-\delta, \delta) \quad(x+\varepsilon, y) \in A \wedge(x, y+\varepsilon) \in A$. Let $\dagger \mathrm{C}_{[0,1]^{2}}$ be choice for sets having a straight cross.

Proposition 7. $\dagger \mathrm{C}_{[0,1]^{2}} \leq_{W} \frac{1}{2} \mathrm{C}_{[0,1]} \star \mathrm{C}_{\mathbb{N}}$.
Corollary 6. $\dagger \mathrm{C}_{[0,1]^{2}}<{ }_{\mathrm{W}} \mathrm{CC}_{2}$
Proof. Combine Proposition 7 with the Fractal Absorption Theorem from [12].
An analogous argument would not succeed in dimension 3 , as $\frac{2}{3} \mathrm{C}_{[0,1]} \equiv{ }_{W} \mathrm{C}_{[0,1]}$ by a majority-voting argument.

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# Secret Sharing Schemes with Algebraic Properties and Applications 

Ignacio Cascudo ${ }^{(\boxtimes)}$<br>Department of Mathematics, Aalborg University, Aalborg, Denmark<br>ignacio@math.aau.dk


#### Abstract

Secret sharing concerns the distribution of some secret information among a number of parties and is among the most well known tools in cryptography. Secret sharing schemes with certain additional algebraic properties, known as linearity and multiplicativity, have important applications in the area of secure multiparty computation and other areas such as zero knowledge proofs. Secret sharing also has a strong relationship with coding theory and motivates new problems in that field. I will survey several of the recent results in the area and some of their applications.


## 1 Introduction

Secret sharing, introduced in [26], is among the most well known tools in cryptography. Secret sharing schemes allow to distribute some secret information among a number of parties by sending some share (some other piece of information) to each of them, in such a way that a small number of shares give no information about the secret which was shared, but a large enough number of shares allow to reconstruct the secret entirely. Secret sharing schemes are useful as a stand-alone primitive (for distributed information storage), but also play an important role as a building block of many cryptographic protocols.

Before defining secret sharing formally, the following notation is fixed. A vector of random variables is a vector $\mathbf{X}=\left(X_{i}\right)_{i \in \mathcal{I}}$ such that the index-set $\mathcal{I}$ is finite and non-empty and the $X_{i}$ 's are random variables defined on the same finite probability space. If $\mathbf{X}$ is a vector of random variables with index-set $\mathcal{I}$ and $A \subset \mathcal{I}$ with $A \neq \emptyset$, then $\mathbf{X}_{A}$ denotes the vector of random variables $\left(X_{i}\right)_{i \in A}$. For each $i \in \mathcal{I}$, we denote by the caligraphic $\mathcal{X}_{i}$ the finite alphabet where $X_{i}$ takes its values. Let $H(X)$ denote the Shannon entropy of a random variable $X$.

Definition 1 (Secret Sharing Scheme). A secret sharing scheme is a vector of random variables $\mathbf{S}$ with index-set $\mathcal{I}=\{0, \ldots, n\}$ and such that:

- Uniformity of the secret: The set $\mathcal{S}_{0}$ satisfies $\left|\mathcal{S}_{0}\right|>1$ and the variable $S_{0}$ satisfies $H\left(S_{0}\right)=\log _{2}\left|\mathcal{S}_{0}\right|$, i.e., $S_{0}$ has a-priori the uniform distribution on $\mathcal{S}_{0}$.

A major part of this work was written while the author was working at the Department of Computer Science, Aarhus University, Denmark.

- Joint reconstruction: $H\left(S_{0} \mid \mathbf{S}_{\mathcal{I} \backslash\{0\}}\right)=0$., i.e., $S_{0}$ is completely determined by the rest of the variables in $\mathbf{S}$.

We abbreviate $\mathcal{I} \backslash\{0\}=\{1, \ldots, n\}$ by $\mathcal{I}^{*}$ and its cardinality is denoted by $n(\mathbf{S})$. The variable $S_{0}$ is the secret, while the remaining variables $S_{i}, i \in \mathcal{I} \backslash\{0\}$ are the shares.

As mentioned before, secret sharing schemes can be used to distribute secret information. If some party (a dealer) wants to distribute certain information $s \in \mathcal{S}_{0}$ among a set of $n(\mathbf{S})$ parties (that we assumed indexed by $\mathcal{I}^{*}$ ), this can be done in the following way: the dealer samples a vector $\mathbf{a}^{\prime}=\left(s, a_{1}, \ldots, a_{n}\right)$ according to the distribution of the conditioned variable $\mathbf{S} \mid S_{0}=s$ and then, for each $i \in \mathcal{I}^{*}$, he sends the value $a_{i} \in \mathcal{S}_{i}$, taken by the variable $S_{i}$, to the $i$-th party. The vector $\mathbf{a}=\left(a_{1}, \ldots, a_{n}\right)$ is then called a sharing of $s$.

Definition 2 (Reconstruction and privacy sets). Let $\mathbf{S}$ be a secret sharing scheme. Let $A \subset \mathcal{I}^{*}$ with $A \neq \emptyset$. Then:

- $A$ is a reconstructing set if $H\left(S_{0} \mid \mathbf{S}_{A}\right)=0$, i.e., the shares of the set $A$ jointly determine the secret with probability 1.
$\mathbf{S}$ has $r$-reconstruction if each subset of $\mathcal{I}^{*}$ of cardinality at least $r$ is a reconstructing set. The reconstruction threshold, denoted by $r(\mathbf{S})$, is the smallest $r$ such that $\mathbf{S}$ has $r$-reconstruction.
- On the other hand, $A$ is a privacy set if $H\left(S_{i} \mid \mathbf{S}_{A}\right)=H\left(S_{i}\right)$, i.e., the a posteriori uncertainty about the secret when given the shares for $A$, equals the a priori uncertainty about the secret (or equivalently, $S_{i}$ and $\mathbf{S}_{A}$ are independent). By definition, $\emptyset$ is a privacy set.
$\mathbf{S}$ has $t$-privacy if each subset of $\mathcal{I}^{*}$ of cardinality at most tis a privacy set. The privacy threshold, denoted by $t(\mathbf{S})$, is the largest $t$ such that $\mathbf{S}$ has $t$-privacy.
Obviously if a secret sharing scheme has $t$-privacy and $r$-reconstruction, it must hold that $r>t$. Typically, for applications it is interesting to have $t$-privacy for as large $t$ as possible, while having $r$-reconstruction for as small $r$ as possible.


## 2 Linear Secret Sharing

Linear secret sharing schemes are the most well known class of secret sharing schemes. They are described next.
Definition 3. Let $\mathbb{F}_{q}$ be a finite field. A linear secret sharing scheme (LSSS) over $\mathbb{F}_{q}$ is a secret sharing scheme $\Sigma$ where the secret and share spaces $\mathcal{S}_{j}$ are $\mathbb{F}_{q}$-vector spaces and $\mathbf{S}$ has the uniform distribution on a $\mathbb{F}_{q}$-linear subspace $V \leq X_{j \in \mathcal{I}} \mathcal{S}_{j}$.

From now on we focus on LSSSs where all the share spaces are the field $\mathbb{F}_{q}$ and the secret space is $\mathbb{F}_{q}^{k}$, for some $k \geq 1$. In that case, it is useful to describe such secret sharing schemes from the point of view of linear codes. Indeed, the support of the variable $\mathbf{S}$ is a linear code $C$ over $\mathbb{F}_{q}$ of length $k+n$, i.e., a linear subspace of $\mathbb{F}_{q}^{k+n}$ (we denote this by $C \leq \mathbb{F}_{q}^{k+n}$ ). Conversely, we have the following construction:

Definition 4. Let $1 \leq k<m$ be integers. Given a linear code $C \leq \mathbb{F}_{q}^{m}$, we define the vector of random variables $\mathbf{S}^{(k)}(C)=\left(S_{i}\right)_{i=0, \ldots, n} \quad$ (where $\left.n=m-k\right)$, whose distribution is given by selecting $\mathbf{c}$ uniformly at random in $C$ (we index its coordinates as $\left.\mathbf{c}=\left(c_{0,1}, c_{0,2}, \ldots, c_{0, k}, c_{1}, c_{2} \ldots, c_{n}\right)\right)$ and defining the secret $S_{0}$ to be $\left(c_{0,1}, c_{0,2}, \ldots, c_{0, k}\right)$ and the $i$-th share $S_{i}$ to be $c_{i}$ for $i=1, \ldots, n$.

Proposition 1. $\mathbf{S}^{(k)}(C)$ is a secret sharing scheme as long as $C$ satisfies two conditions, where $\mathbf{c}_{0}$ will denote the vector $\left(c_{0,1}, c_{0,2}, \ldots, c_{0, k}\right)$ :

1. for all $\mathbf{x} \in \mathbb{F}_{q}^{k}$, there is a word $\mathbf{c} \in C$ with $\mathbf{c}_{0}=\mathbf{x}$ and
2. there is no vector in $C$ of the form $\mathbf{c}=(\mathbf{x}, 0, \ldots, 0)$ where $\mathbf{x} \in \mathbb{F}_{q}^{k} \backslash\{\mathbf{0}\}$.

The advantage of this representation is that we can obtain valuable information about the reconstructing and privacy sets of $\mathbf{S}^{(k)}(C)$ from the supports of the words in $C$. For $A \subseteq \mathcal{I}^{*}, \mathbf{c} \in C$, let $\mathbf{c}_{A}$ denote the vector of coordinates $\left(c_{i}\right)_{i \in A}$. Then one can show the following results.

Proposition 2 (based on [11,25]). Let $A \subseteq \mathcal{I}^{*}$.

- Reconstruction: A is a reconstructing set in the scheme $\mathbf{S}^{(k)}(C)$ if and only if for all $\mathbf{c} \in C$ such that $\mathbf{c}_{A}=\mathbf{0}$, we have $\mathbf{c}_{0}=\mathbf{0}$.
- Privacy: $A$ is a privacy set in the scheme $\mathbf{S}^{(k)}(C)$ if and only if for all $\mathbf{s} \in \mathbb{F}_{q}^{k}$, there exists $\mathbf{c} \in C$ such that $\mathbf{c}_{0}=\mathbf{s}$ and $\mathbf{c}_{A}=\mathbf{0}$.

Another characterization makes use of the dual code $C^{\perp}$ of $C$. Remember that for a linear code $C \leq \mathbb{F}_{q}^{m}, C^{\perp}$ is the set of all $\mathbf{x} \in \mathbb{F}_{q}^{m}$ such that, for every $\mathbf{c} \in C$, the inner product of $\mathbf{x}$ and $\mathbf{c}$ is $0 . C^{\perp}$ is a linear code of the same length as $C$.

Proposition 3 (based on [11,25]). Let $A \subseteq \mathcal{I}^{*}$ and $B:=\mathcal{I}^{*} \backslash A$.

- Reconstruction: A is a reconstructing set in the scheme $\mathbf{S}^{(k)}(C)$ if and only if for all $\mathbf{s} \in \mathbb{F}_{q}^{k}$ there exists $\mathbf{c}^{*} \in C^{\perp}$ with $\mathbf{c}_{0}^{*}=\mathbf{s}$ and $\mathbf{c}_{B}^{*}=\mathbf{0}$.
- Privacy: $A$ is a privacy set in the scheme $\mathbf{S}^{(k)}(C)$ if and only if for all $\mathbf{c}^{*} \in C^{\perp}$ with $\mathbf{c}_{B}^{*}=\mathbf{0}$, we have $\mathbf{c}_{0}^{*}=\mathbf{0}$.

Using any of these two propositions we can observe that, in the case $k=1$, every set $A$ of shares is either a privacy or a reconstructing set. However, when $k>1$, there are sets which are neither privacy nor reconstructing sets, i.e., the knowledge of that set of shares gives some partial information about the secret, but does not determine it completely. In fact, it is always satisfied that

Proposition 4. If $A \subseteq B \subseteq \mathcal{I}^{*}$ are such that $A$ is a privacy set and $B$ is a reconstruction set of $\mathbf{S}^{(k)}(C)$, then $|B|-|A| \geq k$. Consequently, $r\left(\mathbf{S}^{(k)}(C)\right)-$ $t\left(\mathbf{S}^{(k)}(C)\right) \geq k$.

Furthermore, from the characterizations above we can bound the reconstruction and privacy thresholds of $\mathbf{S}^{(k)}(C)$ in terms of the minimum distances of the codes $C$ and $C^{\perp}$. Let $d(C)$ denote the minimum distance of $C$, i.e., the minimum Hamming distance between two distinct codewords in $C$, which, since $C$ is linear, is also the minimum Hamming weight of a nonzero codeword in $C$.

Proposition 5. The reconstruction and privacy thresholds of $\mathbf{S}^{(k)}(C)$ satisfy $r\left(\mathbf{S}^{(k)}(C)\right) \leq n-d(C)+k+1$ and $t\left(\mathbf{S}^{(k)}(C)\right) \geq d\left(C^{\perp}\right)-k-1$.

The most famous example of a secret sharing scheme is Shamir scheme, which is linear. Its original description was only for the case $k=1$, but here a version of it is described that admits secrets in $\mathbb{F}_{q}^{k}$ for larger $k$. Let $k, t, n$ satisfy $1 \leq$ $t+k-1 \leq n$ and $n+k \leq q$. Fix $\alpha_{1}, \ldots, \alpha_{n}, \beta_{1}, \ldots, \beta_{k}$ pairwise distinct elements in $\mathbb{F}_{q}$. Denote by $\mathbb{F}_{q}[X]_{\leq t+k-1}$ the set of univariate polynomials of degree less than $t+k-1$ and coefficients in $\mathbb{F}_{q}$. In order to share a secret $\mathbf{s} \in \mathbb{F}_{q}^{k}$ among $n$ parties with this scheme, a polynomial $f$ is chosen in $\mathbb{F}_{q}[X]_{\leq t+k-1}$ uniformly at random under the additional condition that $f\left(\beta_{j}\right)=s_{j}$ for $j=1, \ldots, k$. Then, the $i$-th share is defined as the evaluation $f\left(\alpha_{i}\right) \in \mathbb{F}_{q}$. This secret sharing scheme is hence constructed as $\mathbf{S}^{(k)}(C)$ from the Reed Solomon code

$$
C=\left\{\left(f\left(\beta_{1}\right), \ldots, f\left(\beta_{k}\right), f\left(\alpha_{1}\right), \ldots, f\left(\alpha_{n}\right)\right): f \in \mathbb{F}_{q}[X]_{\leq t+k-1}\right\}
$$

Shamir's secret sharing scheme has $t$-privacy and $(t+k)$-reconstruction, and it attains both the bounds given by Propositions 4 and 5 .

Linear secret sharing schemes have a property which is of key importance for applications, as it will be explained in Sect.5: suppose several secrets $\mathbf{s}^{(1)}, \mathbf{s}^{(2)}, \ldots, \mathbf{s}^{(\ell)}$ are shared with $\mathbf{S}$, where the sharing for $\mathbf{s}^{(j)}$ is $\left(\mathbf{a}_{1}^{(j)}, \ldots, \mathbf{a}_{n}^{(j)}\right)$. Then given $\lambda^{(1)}, \ldots, \lambda^{(\ell)} \in \mathbb{F}_{q}$, the linear combinations $\left(\sum_{j=1}^{\ell} \lambda^{(j)} \mathbf{a}_{i}^{(j)}\right)_{i=1, \ldots, n}$ constitute a sharing of $\sum_{j=1}^{\ell} \lambda^{(j)} \mathbf{S}^{(j)}$ in the same scheme $\mathbf{S}$.

## 3 Multiplicativity

For applications, it would be useful if a property similar to what we just mentioned held also for multiplications. Ideally, we would like that given sharings of $\mathbf{s}$ and $\mathbf{s}^{\prime}$, the local coordinatewise products of the shares constituted a sharing (in the same scheme) of the coordinatewise product $\mathbf{s} * \mathbf{s}^{\prime}$ of the secrets. Unfortunately, it is easy to see that LSSSs cannot satisfy this property unless in trivial cases. The notion of multiplicative linear secret sharing schemes, introduced in [12], is a relaxation of this property.

Definition 5. An ideal LSSS $\mathbf{S}^{(k)}(C)$ has $\hat{r}$-product reconstruction if for every set $B \subseteq \mathcal{I}$ of size at least $\hat{r}$, there exists some linear function $\rho_{B}: \mathbb{F}_{q}^{|B|} \rightarrow \mathbb{F}_{q}^{k}$ such that for every $\left(\mathbf{c}_{0}, c_{1}, \ldots, c_{n}\right),\left(\mathbf{d}_{0}, d_{1}, \ldots, d_{n}\right) \in C$, it holds that $\rho_{B}\left(\mathbf{c}_{B} * \mathbf{d}_{B}\right)=$ $\mathbf{c}_{0} * \mathbf{d}_{0}$.

This means that the product of two secrets can be reconstructed from the set of the products of the individual shares of the players in $B$.

Definition 6 ([12]). A multiplicative LSSS is a LSSS with n-product reconstruction.

A $t$-strongly multiplicative LSSS is a LSSS with $t$-privacy and $(n-t)$-product reconstruction.

In order to study the multiplicativity of a LSSS, it is useful to introduce the following notion.

Definition 7 ( $m$-th power of a linear code). Let $C \subseteq \mathbb{F}_{q}^{n}$ be a linear code over $\mathbb{F}_{q}, d>0$ an integer. Let

$$
C^{* m}:=\mathbb{F}_{q}\left\langle\left\{c^{(1)} * c^{(2)} \ldots * c^{(m)}:\left(c^{(1)}, c^{(2)}, \ldots, c^{(m)}\right) \in C^{m}\right\}\right\rangle
$$

Remark 1. By definition, $\mathbf{S}^{(k)}(C)$ has $\hat{r}$-product reconstruction if and only if $\mathbf{S}^{(k)}\left(C^{* 2}\right)$ has $\hat{r}$-reconstruction. Consequently, $\mathbf{S}^{(k)}(C)$ has $\hat{r}$-product reconstruction for $\hat{r}=n-d\left(C^{* 2}\right)+k+1$ and it is $t$-strongly multiplicative if $d\left(C^{\perp}\right), d\left(C^{* 2}\right) \geq t+k+1$.

Proposition 6. Shamir's scheme (with threshold t) has $(2 t+2 k-1)$-product reconstruction, as long as $2 t+2 k-1 \leq n$. Hence, Shamir's scheme is $t$-strongly multiplicative as long as $3 t+2 k-1 \leq n$.

This stems from the fact that if $C$ is a Reed-Solomon code given by evaluations of the polynomials in $\mathbb{F}_{q}[X]_{\leq t+k-1}$, then $C^{* 2}$ is another Reed-Solomon code given by the evaluations, in the same points, of the polynomials in $\mathbb{F}_{q}[X]_{\leq 2 t+2 k-2}$.

On the other hand, the following limitation is easy to argue in view of the characterization from Proposition 2.

Proposition 7. Any LSSS with t-privacy and $\hat{r}$-product reconstruction has $(\tilde{r}-t)$-reconstruction. Consequently it needs to satisfy $\hat{r} \geq 2 t+k$. As a consequence of this, any t-strongly multiplicative LSSS satisfies $3 t+k \leq n$.

In the case $k=1$, the bounds are tight: Shamir's scheme attains the maximum value of $t$ (with respect to the number of players $n$ ) for which there is a $t$-strongly multiplicative LSSS. As far as the author knows, the bound above is not known to be tight for $k>1$ :

Conjecture 1. Any LSSS with $t$-privacy and $\hat{r}$-product reconstruction satisfies $\hat{r} \geq 2 t+2 k-1$ and any $t$-strongly multiplicative LSSS satisfies $3 t+2 k-1 \leq n$.

## 4 Asymptotics

Shamir's scheme achieves several good properties from the point of view of the applications. It matches the bound for $r\left(\mathbf{S}^{(k)}(C)\right)-t\left(\mathbf{S}^{(k)}(C)\right)$ given in Proposition 4. In particular, for $k=1$, one has the property that any set of $t$ players has no information about the secret, while any set of $t+1$ can reconstruct it. Furthermore, possibly up to an additive factor $k$, Shamir's scheme also attains, for a given $n$, the largest possible $t$ for which there can be $t$-strong multiplication.

However, Shamir's scheme has the limitation that the number of players $n$ is upper bounded by the size of the field. Indeed, in the definition we have given
it is required that $k+n \leq q$ (one can slightly weaken this by adding, as an evaluation point, the coefficient of order $t+k-1$ of the evaluation polynomial, so that this condition is relaxed as $k+n \leq q-1$, and all the properties are preserved). ${ }^{1}$

This raises the question whether similar properties can be attained by LSSS $\mathbf{S}$ if we fix the size of the field $q$, but want the number of players $n(\mathbf{S})$ to be arbitrarily large. However, the following result sets a restriction in what we can hope to achieve.

Theorem 1. [6] For any $L S S S \mathbf{S}=\mathbf{S}^{(1)}(C)$ we have $r(\mathbf{S})-t(\mathbf{S}) \geq \frac{n(\mathbf{S})+2}{2 q-1}$. Consequently, if $\mathbf{S}$ has $\hat{r}$-product reconstruction, then $\hat{r}-2 t(\mathbf{S}) \geq \frac{n(\mathbf{S})+2}{2 q-1}$ and if $\mathbf{S}$ is $t$-strongly multiplicative, then $3 t \leq\left(1-\frac{1}{2 q-1}\right) n(\mathbf{S})$.

In fact, by applying somewhat more elaborate arguments, [6] shows a stronger bound for t-strongly multiplicative schemes, namely $3 t \leq\left(1-\frac{3 q-2}{3 q^{2}-3 q+1}\right) n$. Extensions for the case where the space of secrets is $\mathbb{F}_{q}^{k}$ are also shown in [6].

These results imply that, when $n$ is much larger than $q$, we will lose a factor $c(q) \cdot n$ in the maximum value of $t$ for which we can have $t$-strong multiplicativity with respect to what Shamir's scheme attains. The question then is if we can achieve $t=\Omega(n)$ (i.e. $t \geq c^{\prime}(q) n$ for some other constant $\left.c^{\prime}(q)>0\right)$ at all. This is answered by the following result.

Theorem 2 ([2,4,7,10]). For any finite field $\mathbb{F}_{q}$, there exists an infinite family of secret sharing schemes $\mathbf{S}^{\left(k_{n}\right)}\left(C_{n}\right)$, indexed by $n \in \mathbb{N}$, such that $\mathbf{S}^{\left(k_{n}\right)}\left(C_{n}\right)$ has $n$ players and is $t_{n}$-strongly multiplicative, where $k_{n}, t_{n}=\Omega(n)$ as $n \rightarrow \infty$.

The result makes use of algebraic geometric codes defined on asymptotically good families of algebraic function fields [19]. It was originally proved for all fields of large enough cardinality in [10]. This in particular included all fields $\mathbb{F}_{q}$ with $q \geq 49$ and square. Later, the result was extended to every finite field $\mathbb{F}_{q}$ in [2], by using certain dedicated concatenation of codes, where the outer code is defined to be a code from the family in [10], defined on an appropriate constant-degree extension field of $\mathbb{F}_{q}$ and the inner code is a Reed-Solomon code over $\mathbb{F}_{q}$. In [4], it was shown that, by choosing the parameters carefully, and applying more involved algebraic-geometric arguments which, in particular, include bounding the size of the 2-torsion subgroup of the divisor class group of the function fields involved, one can use the algebraic geometric construction in [10] (without the concatenation technique in [2]) directly to show the result for all fields $\mathbb{F}_{q}$ with $q \geq 16$, and also $q=8,9$. Furthermore, the actual values of $t_{n}$ promised by the construction are larger for smaller fields. The next result is an example of which values of $t_{n}, k_{n}$ (with respect to $n$ ) can be attained in some cases. For the full results see $[2,4,7,10]$.

[^8]Theorem 3. Let $\mathbb{F}_{q}$ be a finite field. There exists an infinite family of secret sharing schemes $\mathbf{S}^{\left(k_{n}\right)}\left(C_{n}\right)$, indexed by $n \in \mathbb{N}$ such that $\mathbf{S}^{\left(k_{n}\right)}\left(C_{n}\right)$ has $n$ players, $t_{n}$-privacy and $\hat{r}_{n}$-product reconstruction, where $k_{n}=\lfloor\kappa n\rfloor, t_{n}=\lfloor\tau n\rfloor, \hat{r}_{n}=$ $\lfloor\hat{\rho} n\rfloor$ in the following cases:

- $q$ square with $q \geq 49$, as long as $1 \geq \hat{\rho} \geq 2 \kappa+2 \tau+\frac{4}{\sqrt{q}-1}$.
- $q$ square, even, with $q \geq 16$, as long as $\hat{\rho} \leq 1$ and

$$
\hat{\rho}-\frac{h_{2}(\hat{\rho})}{\log q}-2 \tau-2 \frac{h_{2}(\tau)}{\log q} \geq\left(2+\frac{(\sqrt{q}+1) \log q+1}{(q-1) \log q}\right) \kappa+\frac{(\sqrt{q}+1) \log q+1}{(q-1) \log q} .
$$

$-q=2$, as long as $\hat{\rho} \leq 1$ and

$$
\hat{\rho}-\frac{h_{2}(\hat{\rho})}{4}-30 \tau-\frac{h_{2}(15 \tau)}{2} \geq\left(2+\frac{21}{60}\right) \kappa+\frac{21}{60},
$$

where in all cases $h_{2}(\cdot)$ denotes the binary entropy function.
A drawback of these constructions is the high complexity of constructing the generator matrices of algebraic geometric codes defined on the asymptotically good families of function fields from [19]. Unfortunately, so far there is no known way to attain the result of Theorem 2 using "elementary" techniques, for example randomized constructions or polynomial evaluations. This holds even if we drop the condition $t_{n}$-strongly multiplicative in Theorem 2 and we require instead $t_{n}$-privacy and $n$-product reconstruction, still with $t_{n}, k_{n}=\Omega(n)$. However, if we weaken our requirements further and set $k_{n}=1$, then there exists a much more elementary construction using random self-dual codes [11].

Nevertheless, randomized constructions do not seem to have good multiplicative properties in general. In fact, for random codes $C$, squaring seems to be a "destructive" operation, in the sense that the dimension of $C^{* 2}$ grows as the square of the dimension of $C$. Indeed, [3] showed that

Theorem 4. Let $\ell \leq N \leq \frac{\ell(\ell+1)}{2}$. Write $u=\frac{\ell(\ell+1)}{2}-N$. Let $C$ be a code chosen uniformly at random among all codes of length $N$ and dimension $\ell$. Then

$$
\operatorname{Pr}\left[C^{* 2}=\mathbb{F}^{N}\right]=1-2^{-\Theta(\ell)}-2^{-\Theta(u)} .
$$

This means that if, in particular $\ell=\Theta(N)$, then $C^{* 2}$ will be the full space $\mathbb{F}^{N}$ with overwhelming probability.

## 5 Applications

## Secure Computation

Secure computation is concerned with the following situation: $n$ parties, each holding some private input $x_{i}$, want to correctly compute $f\left(x_{1}, x_{2}, \ldots, x_{n}\right)$ for some agreed upon function $f$ in such a way that the intended output is the only
new information released. This guarantee should be fulfilled even if a small number $t$ of the players are corrupted and cheat. Corruption is passive if corrupted players follow the specified protocol (but compare their views of the protocol to try to gain additional information) and active if they can behave arbitrarily. For formal security definitions and more information about the topic, see [13].

For the case of a computationally unbounded adversary, it is known $[1,9]$ that secure computation of any function is possible as long as less than $n / 2$ players are passively corrupted, or less than $n / 3$ players are actively corrupted ${ }^{2}$. The results make use of the linearity and multiplicativity of Shamir's scheme. Later in [12], it was shown that, in fact, secure multiparty computation protocol secure against a passive (resp. active) adversary corrupting $t$ players can be constructed from any multiplicative LSSS with $t$-privacy (resp. $t$-strong multiplicative LSSS).

The idea of the protocol in the passive case is as follows: assume for the moment $f: \mathbb{F}^{n} \rightarrow \mathbb{F}$ is a linear function. Suppose each input is in $\mathbb{F}, \mathbf{S}:=\mathbf{S}^{(1)}(C)$ is a LSSS over $\mathbb{F}$ with $t$-privacy. At the beginning of the protocol, each player can secret share her input $x_{i}$ with $\mathbf{S}$. Then each player applies $f$ to her received shares. By linearity, the resulting set of values is a sharing of $f\left(x_{1}, \ldots, x_{n}\right)$ as mentioned at the end of Sect.2. So now players can broadcast the resulting shares, and reconstruct the output. Ideally, we would want to extend this idea to any function $f$. Since every function can be computed as an arithmetic circuit over a finite field (for example a Boolean circuit), it would be enough if a sharing of $s s^{\prime}$ can be created from the sharings of $s$ and $s^{\prime}$. However, we mentioned at the beginning of Sect. 3, if $s$ and $s^{\prime}$ are secret shared with a LSSS $\mathbf{S}:=\mathbf{S}^{(1)}(C)$, and each player multiplies her shares, even if the original scheme was multiplicative, the resulting set of values is not a sharing of $s s^{\prime}$ in the same scheme $\mathbf{S}^{(1)}(C)$, but in $\mathbf{S}^{(1)}\left(C^{* 2}\right)$.

A better alternative is obtained by allowing interaction among the players. Again let $\mathbf{S}^{(1)}(C)$ be multiplicative. By definition there is a linear function $\rho$ such that $s s^{\prime}=\rho\left(c_{1} c_{1}^{\prime}, \ldots, c_{n} c_{n}^{\prime}\right)$ where $c_{i}, c_{i}^{\prime}$ are the shares of $s$ and $s^{\prime}$ respectively. The following protocol allows to create a sharing $s s^{\prime}$ in the scheme $\mathbf{S}^{(1)}(C)$ : Each player $i$ shares the product $c_{i} c_{i}^{\prime}$ of his shares with $\mathbf{S}^{(1)}(C)$. Now, players have shares of $c_{1} c_{1}^{\prime}, c_{2} c_{2}, \ldots, c_{n} c_{n}^{\prime}$ and they can locally compute shares of $s s^{\prime}=$ $\rho\left(c_{1} c_{1}^{\prime}, \ldots, c_{n} c_{n}^{\prime}\right)$ since $\rho$ is linear. This can be shown to be secure against a passive adversary corrupting at most $t$ parties if the scheme has $t$-privacy and it is enough to argue that any function can be securely computed by $n$ players in this situation.

The case of an active adversary requires additional techniques. The problem is that corrupted players (which can deviate from the protocol) may choose to deal inconsistent shares to the honest players or to share wrong values. The solution in [12] involves using verifiable secret sharing, in which the dealer sends additional information to each player which can be used to verify that the shares have been well constructed. If the underlying LSSS is $t$-strongly multiplicative,

[^9]this leads to a secure protocol to compute any function in the presence of an adversary corrupting at most $t$ players.

## Two-Party Cryptography and Other Applications

In recent years, secure multi-party computation protocols have surprisingly been applied on cryptographic problems involving only two players. This is due to the MPC-in-the-head technique, which was introduced in [22] for the purpose of zero knowledge proofs with high communication efficiency, where a virtual multiparty computation protocol with a large number of parties is run by of the two players at some point of the protocol. The ideas from [22] have inspired other applications in areas such as multiparty computation with dishonest majority [17,24], OT combiners [20], OT from noisy channels [21], correlation extraction [23] and UC homomorphic commitment schemes $[8,16,18]$. In many of these applications, the best results in terms of efficiency are attained by setting up multiparty computation protocols with a very large number of players, while the size of the field should be small. This has been a motivation for the results on asymptotics of $t$-strongly multiplicative LSSS discussed in Sect. 4.

Finally, one can generalize the concept of multiplicative secret sharing schemes in a number of ways. A generalization is given by the notion of arithmetic codices introduced in [5]: this notion allows to address also other applications in two-party cryptography (zero knowledge proofs of algebraic relations [14]), algebraic complexity (see [7]) and more recently has been applied back in the realm of error correcting, more specifically in the problem of local decodability of Reed-Muller codes (see [15]).

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# Squeezing Feasibility 

Walter Dean ${ }^{(\boxtimes)}$<br>Department of Philosophy, University of Warwick, Coventry CV4 7AL, UK<br>W.H.Dean@warwick.ac.uk<br>http://go.warwick.ac.uk/whdean


#### Abstract

This note explores an often overlooked question about the characterization of the notion model of computation which was originally identified by Cobham [5]. A simple formulation is as follows: what primitive operations are allowable in the definition of a model such that its time and space complexity measures provide accurate gauges of practical computational difficulty? After exploring the significance of this question in the context of subsequent work on machine models and simulations, an adaptation of Kreisel's squeezing argument [17] for Church's Thesis involving Gandy machines [11] is sketched which potentially bears on this question.


## 1 Cobham's Problem

The goal of this note is to highlight a historically significant but often overlooked question about the characterization of what is commonly called a model of computation. A simple formulation is as follows: what primitive operations are allowable in the definition of such a model such that its time and space complexity measures provide accurate gauges of practical computational difficulty?

A version of this problem was posed in the paper "The intrinsic computational difficulty of functions" [5] in which Alan Cobham first defined the class FP of functions computable in polynomial time and conjectured that it coincides with the informal notion of a feasibly computable function:
[W]e find that it makes a definite difference what class of computational methods and devices we consider in our attempt to formalize the definition [of a feasibly computable function]. The problem is reminiscent of, and obviously closely related to, that of the formalization of the notion of effectiveness. But the emphasis is different in that the physical aspects of the computation process are here of predominant concern. The question of what may legitimately be considered to constitute a step of a computation is quite unlike that of what constitutes an effective operation . . . If we are to make fine distinctions, say between [TIME $(n)$ ] and $\left[\operatorname{TIME}\left(n^{2}\right)\right]$, then we must have an equally fine analysis of all phases of the computational process ... We must be prepared to argue that we haven't taken
too broad a class for $[\operatorname{TIME}(n)]$ and thus admitted to it functions not in actuality computable in a number of steps linearly bounded by the lengths of its arguments. ${ }^{1}$
[5], pp. 29-30
Cobham illustrated the significance of the choice of basic operations by posing the following question: is there a model-independent - or "absolute" - sense in which integer multiplication is intrinsically harder to compute than integer addition? ${ }^{2}$ A theory of computational complexity based on a model of computation which treats both addition and multiplication as operations which can be computed in a single primitive step is clearly incapable of providing a principled answer to this question. But since any model of computation must treat some operations as primitive, how are we to make a principled choice? Cobham suggested that not only was this one of the "fundamental problems" facing the development of such a theory but that it "may well call for considerable patience and discrimination" ([5] p. 30).

## 2 Machine Models and the Invariance Thesis

My present goal will not be to attempt to resolve Cobham's problem directly. For in hindsight it is clear that many of the issues he identified are likely related to fundamental unresolved questions about the separation of complexity classes. I merely wish to argue that some of the surrounding issues may be clarified by reflecting further on the developments which led to the adoption of the multitape Turing machine $\mathfrak{T}$ as the standard model employed in complexity theory and its use in formulating what is now often called the Cobham-Edmonds Thesis:

A function $f(x)$ is feasibly computable just in case it is computed by a machine $T \in \mathfrak{T}$ such that the running time complexity of $T$ is asymptotically bounded by $|x|^{k}$ for some fixed $k \in \mathbb{N}$ (i.e. such that time ${ }_{T}(|x|) \in O\left(|x|^{k}\right)$ where $|x|$ denotes the size of the input).

This proposal is often presented in textbooks as playing the same role for complexity theory which Church's Thesis plays for computability theory - i.e. that of providing a mathematically precise analysis of the pre-theoretical notion of feasibility (i.e. computability in practice) in the same way which Church's Thesis is
${ }^{1}$ Cobham characterized the class $\mathcal{L}$ of feasibly computable functions as those computable by an informal algorithm in "a number of steps ... bounded by polynomials in the lengths of the numbers involved". He proposed that $\mathcal{L}$ coincided with the functions definable using a restricted recursion scheme known as limited recursion on notation. He then conjectured that the class of functions definable in this manner coincided with FP under its contemporary definition - i.e. the class of functions computable in polynomial time by a deterministic Turing machine. This was later demonstrated by Rose [20].
2 "Is it harder to multiply than to add? ... I grant I have put [this] question loosely; nevertheless, I think the answer ought to be yes" [5], p. 24. By the mid-1960s it was already possible to marshall a good deal of evidence in favor of this hypothesis e.g., [7].
generally taken to provide an analysis of the pre-theoretical notion of effectivity (i.e. computability in principle). This in turn suggests that the complexity class FP enjoys a status similar to that of the class of Turing computable functions as an "intrinsic" or "absolute" standard of computational difficulty. But although it remains controversial whether this is indeed the case, it is not the identification of feasibility with polynomial time computability to which I wish to draw attention. Rather, it is the rationale for using $\mathfrak{T}$ in the definition of "polynomial time" and "polynomial space" provided that these quantities are intended to serve as useful metrics of computational difficulty.

An early precedent for this was provided by Hartmanis and Stearns [13] who demonstrated the first hierarchy theorems showing that more functions become computable relative to a sufficiently large increase in the number of allowable steps consumed or tape cells accessed during the course of a Turing machine computation. Soon after this, simulation results were obtained linking the time or space bounds for $\mathfrak{T}$ to a variety of other tape-based models which might have initially seemed more or less powerful. For instance if $f(x)$ can be computed in time $O(t(n))$ using a multi-tape machine from $\mathfrak{T}$, then it can be computed in time $O\left(t(n)^{2}\right)$ using a single-tape machine. And if $f(x)$ can be computed in time $O(t(n))$ using a machine with a 2-dimensional tape then it can be computed in time $O\left(t(n)^{3 / 2}\right)$ by a machine from $\mathfrak{T}$ using multiple one-dimensional tapes.

Similar results were also obtained for the basic RAM model $\mathfrak{R}$ introduced by Cook and Reckhow [8]. Recall that this model consists of a finite program, one or more accumulator registers, an instruction counter, and an infinite collection of memory registers $R_{0}, R_{1}, \ldots$ which may contain integers of unbounded size. The operations performable by $\mathfrak{R}$ include reading inputs and printing outputs, conditional jump operations based on comparisons between the accumulator registers or other registers, instructions for transferring values between registers and the accumulator (inclusive of indirect addressing), and the following arithmetical operations: addition and subtraction. In calculating the running time complexity of a RAM machine, it is often assumed that these operations can be carried out in unit time - i.e. as a single unmediated step.

Although this latter assumption is evidently false for tape-based models, a basic simulation result about the relative efficiency of $\mathfrak{R}$ and $\mathfrak{T}$ is as follows: if $f(x)$ can be computed in time $O(t(n))$ using a machine from $\Re$, then it can be computed in time $O\left(t(n)^{3}\right)$ using a machine from $\mathfrak{T}$. Similar results apply to some (although as we will see not all) register-based models which are obtained by modifying $\mathfrak{R}$ in various straightforward ways - e.g. allowing that values stored in registers may be treated as binary numerals on which it is possible to apply bit-wise Boolean operations in parallel. ${ }^{3}$

The significance of these results is borne out by the observation that $\mathfrak{R}$ is sufficiently flexible to allow for implementations of many algorithms which arise in mathematical practice - say in number theory, linear algebra or graph theory - which are not only reasonably direct, but also have running time proportional

[^10](often by scalar factors) to their informal counterparts. ${ }^{4}$ This provides inductive confirmation that measuring computational complexity relative to $\mathfrak{R}$ provides an accurate gauge of the practical difficulty of solving computational problems of the sort which we confront in practice.

It has become conventional to call a model of computation with this property (such as $\mathfrak{R}$ ) reasonable. On the basis of simulation results of the sort mentioned above, van Emde Boas [27] has proposed that the notion of "reasonability" should be extended to other models as follows:
Invariance Thesis: Suppose that $\mathfrak{N}_{0}$ is a reasonable model of computation. Then another model $\mathfrak{N}_{1}$ is also reasonable just in case every machine in $\mathfrak{N}_{0}$ can be simulated by a machine $\mathfrak{N}_{1}$ with polynomially bounded overhead in time and a constant-factor overhead in space and conversely with respect to $\mathfrak{N}_{1}$ and $\mathfrak{N}_{0} .{ }^{5}$

It thus follows from the identification of $\mathfrak{R}$ as a reasonable model coupled with the simulation results mentioned above that the Invariance Thesis predicts that $\mathfrak{T}$ and its variants (e.g. single-tape and multi-dimensional tape machines) are also reasonable models. Some complications remain in the formulation of a version of the Invariance Thesis which is sufficiently general to simultaneously cover all of the apparently reasonable register models which have been proposed. ${ }^{6}$ But as we have seen, the original class of tape-based models which satisfy the Invariance Thesis with respect to $\mathfrak{T}$ is quite robust. This demonstrates that we would not change the extensional characterization of the class of polynomial time computable functions if we had elected to measure complexity with respect to single tape Turing machines, machines with two dimensional tapes, machines with stacks or counters instead of conventional tapes, etc.

Basing the definition of $\mathbf{F P}$ on $\mathfrak{R}$ would also yield an extensionally equivalent characterization. However, this depends more sensitively on the details of the register-based model considered. Consider, for instance, the so-called MRAM model $\mathfrak{M}$ obtained by supplementing the instructions available in $\mathfrak{R}$ with unit time multiplication. It follows by results of Hartmanis and Simon [12] and Bertoni et al. [2] that what can be computed in polynomial time relative to $\mathfrak{M}$ coincides with what can be computed in polynomial space relative to $\mathfrak{R}$ (and hence also $\mathfrak{T}$ ). ${ }^{7}$ To put this point more precisely, let $\mathbf{C}_{\mathfrak{N}}$ denote the class of languages decidable using the time or space resource $\mathbf{C}$ relative to the model $\mathfrak{N}$.

[^11]Then the results just described establish that the class $\mathbf{P}_{\mathfrak{M}}$ of problems decidable in polynomial time relative to the MRAM model contains the class $\mathbf{P S P A C E}_{\mathfrak{T}}$ of problems decidable in polynomial space by a multi-tape Turing machine.

Although it is presently unknown whether $\mathbf{P}_{\mathfrak{T}} \neq \mathbf{P S P A C E} \mathbf{E}_{\mathfrak{T}}$, it is widely believed that the former class is a proper subset of the latter. Relative to this conjecture, $\mathfrak{M}$ is not a reasonable model of computation. ${ }^{8}$ And despite the fact that $\mathfrak{M}$ is obtained from $\mathfrak{R}$ by a seemingly innocent extension, there is good evidence for such a classification. For it is not hard to see that polynomial time and polynomial space coincide for $\mathfrak{M}$ - i.e. $\mathbf{P}_{\mathfrak{M}}=\mathbf{P S P A C E} \boldsymbol{M}_{\mathfrak{M}}$. This has traditionally been taken to be the touchstone of models which allow for parallelism involving exponential growth in the number of processing units - e.g. the $k$-PRAM model $\mathfrak{P}$ of Savitch and Stimson [21] which in a single step allows a RAM-like device to create $k \geq 2$ copies of itself (inclusive of all its registers). It is easy to see, however, that there exist concrete instances of problems which we cannot currently decide in practice but which are in $\mathbf{P}_{\mathfrak{P}}$ (and hence also in $\mathbf{P}_{\mathfrak{M}}$ ). van Emde Boas [27] refers to such models as inhabiting what he refers to as the second machine class. Of such models he writes that "it is not clear at all that they can be considered to be reasonable; it seems that the marvelous speed-ups obtained by the parallel models of the second machine class require severe violations of basic laws of nature." (p. 14).

The foregoing observations illustrate that not only is the choice of basic operations central to developing an account of intrinsic computational difficulty, but that Cobham was prescient in using the contrast in the apparent difficulty of addition and multiplication to illustrate one of the fundamental challenges of providing such an account. For although it is easy to see that there is an $O(n)$ algorithm for computing integer addition relative to $\mathfrak{T}$ (where $n$ is the maximum of the lengths of the inputs), it is currently unknown whether this is the case for multiplication with respect to both $\mathfrak{T}$ and $\mathfrak{R} .{ }^{9}$ In choosing a benchmark for defining complexity classes, there are, of course, clear practical reasons to move from the single-tape Turing machine model (which does not allow for a linear time implementation of addition) to the multi-tape model $\mathfrak{T}$, and again from $\mathfrak{T}$ to a register-based model such as $\mathfrak{R}$. But the apparent sensitivity of "polynomial time" with respect to the inclusion or exclusion of arithmetical operations from this class suggests that we have yet to realize Cobham's hope of finding an intrinsic rationale for using this model (together with the Invariance Thesis) as a means of characterizing feasible computability.

[^12]
## 3 Arguing for the Cobham-Edmonds Thesis

Drawing on the analogy with Church's Thesis, there are at least three ways one might go about arguing for the Cobham-Edmonds Thesis. First one might attempt to show inductively that FP accurately circumscribes the class of functions which we can compute in practice. Second, one might attempt to argue that this class of functions is "intrinsic" in virtue of being definable by extensionally coincident but independently motivated definitions. And third, one might attempt to argue directly that at least one of these definitions serves as a correct analysis of the pre-theoretical notion of a function computable in practice.

These strategies have been explored in as much detail in the complexity theoretic setting as have the analogous arguments in favor of Church's Thesis. Nonetheless, it is evident that they all face complications which their computability-theoretic counterparts do not have to confront. There are, for instance, well known objections even to the claim that FP extensionally coincides with feasible computability both on the basis that the former class admits functions computable only by Turing machines whose running time complexity involves "infeasible" constant and scalar factors and also because of the potential vagueness or interest relativity in our background notion of feasibility itself. ${ }^{10}$

With respect to the second argument, recall that there are several well-known alternative characterizations of effective computability in terms of (e.g.) general recursiveness or unrestricted grammars which yield extensionally equivalent definitions of the class of Turing computable functions. We now also know of alternative characterizations of $\mathbf{F P}$ - e.g. in terms of Cobham's own characterization in terms of limited recursion on notation, the provably recursive functions of certain systems of bounded arithmetic, or logical definability as studied within descriptive complexity theory.

There is, however, at least one salient difference between these cases. For in the original instances, the alternative characterization of the Turing computable functions are arrived at by generalizing definitions to achieve a maximal characterization of what can potentially be computed relative to a certain class of formalisms which meet only very general finiteness requirements - e.g. by eliminating restrictions on the allowable forms of recursion schemes or production rules. This, however, is not the case for the sort of "machine independent" characterizations of FP just alluded to. In particular, by changing obvious parameters in the relevant definitions we can also arrive at characterizations of well-known complexity classes which are both narrower and broader than FP. ${ }^{11}$ Thus while

[^13]this class is robust in the sense of admitting independent definitions, the alternative definitions which are presently available lack the sort of canonicity which is traditionally thought to accompany the various characterizations of Turing computability whose coincidence is often cited in favor of Church's Thesis.

## 4 The Squeezing Argument

Let us now consider the status of the third potential sort of argument for the Cobham-Edmonds Thesis - i.e. that of attempting to show that one of the available characterizations of FP provides a conceptual analysis of the pre-theoretical notion of a feasibly computable function. Presuming that mathematical practice affords us a sufficiently robust understanding of this notion, nothing appears to preclude the possibility of such an argument. But if we proceed in this manner, it would seem that we run directly into Cobham's problem - i.e. if we lack a prior mathematical characterization of feasibility, how are we to make a principled decision as to what basic operations may legitimately be included in a model of computation on which to base the analysis?

This might at first appear to leave us at an impasse. But I now wish to suggest that it may be possible to modify an indirect argument which has sometimes been put forward in favor of Church's Thesis to argue for the Cobham-Edmonds Thesis. The method in question is a variant of the so-called squeezing argument due to Kreisel. Kreisel [17] originally used this method to argue that the pretheoretically identified class InfVal of statements which are valid in virtue of their logical form extensionally coincides with the class FormVal of formal validities of first-order logic. In order to reach this conclusion he argued that InfVal could be "squeezed" between FormVal and the class Prov of formulas provable in a given proof system which is sound and complete for first-order logic.

First note that since we regard the rules of (say) the sequent calculus as individually sound in the informal sense, a simple induction yields that any statement derivable using these rules from no premises is informally valid. Since provability is hence a sufficient condition for informal validity, we have that
(1) Prov $\subseteq I n f V A L$

On the other hand, note that even if we do not start out with an entirely settled notion of what it means for a statement to be valid in virtue of its logical form, we presumably acknowledge that no statement which possesses a traditional Tarskian countermodel (i.e. a model with a set-sized domain) can be valid in this sense. Since Tarskian validity is thus a necessary condition for informal validity, we thus also have that
(2) Inf VAL $\subseteq$ FormVal

Note, however, that it is a consequence of the Completeness Theorem for firstorder logic that FormVal $\subseteq$ Prov. And from this it follows from (1) and (2) that $\operatorname{Inf}$ Val $=$ FormVal as desired.

On at least two occasions $[18,19]$ Kreisel hinted that the squeezing argument might be adapted to Church's Thesis. The evident strategy here would
be to attempt to "squeeze" the pre-theoretically identified class EffComp of effectively computable functions between the class of functions $\mathfrak{N}_{0}$-Comp and $\mathfrak{N}_{1}$-Comp computable relative to models of computation $\mathfrak{N}_{0}$ and $\mathfrak{N}_{1}$. If it could be established that membership in $\mathfrak{N}_{0}$-Comp was a sufficient condition for effectivity and that membership in $\mathfrak{N}_{1}$-Comp was a necessary condition, then we could similarly conclude that EffComp $=\mathfrak{N}_{0}$-Comp provided that we could demonstrate that $\mathfrak{N}_{0}$-Comp $=\mathfrak{N}_{1}$-Comp.

Although Kreisel was not explicit about the forms which such models might take, Turing [25] may plausibly be regarded as providing an argument that there is good justification for taking $\mathfrak{N}_{0}=\mathfrak{T}_{1}$ (i.e. the single tape Turing machine). For in the case of each defining property of this model - e.g. the finiteness of the number of distinguishable symbols and "states of mind", the discreteness and one-dimensionality of the tape, the fact that only a single cell can be scanned at a given step, etc. - he gives an argument that a computational method satisfying these properties jointly is effective. And for this reason Turing computability is traditionally regarded as sufficient for effective computability.

But as Turing himself admits (p. 249), it is less clear that these specific conditions are also necessary properties of an effective method. To take his own example, for instance, it is evident that some effective methods (e.g. carry addition and multiplication) take advantage of a two-dimensional computing medium. And we now also typically allow that certain forms of parallel processing - e.g. of the sort employed in conventional parallel algorithms for sorting and matrix multiplication - are also be effective. It is thus not immediately clear how to choose a model $\mathfrak{N}_{1}$ which simultaneously meets the conditions that $\mathfrak{T}_{1}$-COMP $=$ $\mathfrak{N}_{1}$-Comp and is also such that if a function failed to be $\mathfrak{N}_{1}$-computable, then we could be reasonably assured that was not effective computability.

A potential example of such a model was provided by Gandy [11]. His model $\mathfrak{G}$ formalizes a notion of computation on hereditarily finite sets which which is intended to generalize Turing's requirements as far as possible, compatible with the assumption that the described computations can be carried "mechanistically" i.e. by a finite, discrete, deterministic device which can be physically constructed. ${ }^{12}$ The constraints on such devices take the form of what Sieg [24] calls boundedness and locality conditions - i.e. i) there is a finite bound on the number of distinct symbol configurations which can be recognized in a single step and also on the number of distinct internal states; ii) only observed symbol configurations within a bounded distance of the previously observed configurations may be modified in a single step. These conditions are sufficiently liberal to allow for direct formalizations of algorithms which arise in mathematical practice in something like the manner of $\mathfrak{R}$, thus providing some evidence that EfFComp $\subseteq \mathfrak{G}$-Comp. But despite

[^14]the abstractness of the Gandy-Sieg conditions, it is still possible to show that the $\mathfrak{T}$-Comp $=\mathfrak{G}$-Comp - i.e. the class of functions computable by Gandy machines is coincident with the class of Turing computable functions.

If it is accepted that the Gandy-Sieg conditions provide a reasonable upper bound on the class of symbolic operations which can be performed in a single step, then the result just cited can be invoked to complete the squeezing argument for Church's Thesis in a manner which avoids the need to provide a prior characterization of what might be meant by an effective primitive operation-i.e. one of the sort illustrated but not fully characterized by Turing's analysis. This feature of the argument might also inspire hope that it could also be adapted to provide an argument for the Cobham-Edmonds Thesis which circumvents the need to directly confront Cobham's problem - i.e. that of giving a definition of a feasible primitive operation prior to an analysis of feasibly computable.

The relevant adaptation would attempt to squeeze the informally identified class of feasibly computable functions FeasComp between the classes of $\mathbf{F P}_{\mathfrak{Q}_{0}}$ and $\mathbf{F} \mathbf{P}_{\mathfrak{Q}_{1}}$ which are computable in polynomial time relative to models $\mathfrak{Q}_{0}$ and $\mathfrak{Q}_{1}$. In this case, we wish $\mathfrak{Q}_{0}$ to be what might be called a minimally reasonable model - i.e. one which satisfies the Invariance Thesis and also is such that the inclusion of $f(x)$ in $\mathbf{F} \mathbf{P}_{\mathfrak{Q}_{0}}$ is sufficient to ensure that $f(x)$ is intuitively feasible. On the other hand, we would desire to show that $\mathfrak{Q}_{1}$ is maximally reasonable - i.e. provides maximally direct implementations of feasible algorithms subject to the requirement of satisfying the Invariance Thesis.

Although we must ultimately demonstrate that $\mathbf{F P}_{\mathfrak{Q}_{0}}=\mathbf{F P}_{\mathfrak{Q}_{1}}$, nothing prohibits us from choosing $\mathfrak{Q}_{0}=\mathfrak{T}$ or even $\mathfrak{T}_{1}$. On the basis of the foregoing, one might additionally hope that it is also admissible to choose $\mathfrak{Q}_{1}=\mathfrak{G}$. However Gandy's model was explicitly designed to allow for the sort parallel evolution exhibited by cellular automata such as Conway's Game of Life. It is thus not immediately clear whether $\mathfrak{G}$ is a reasonable model in the sense discussed above or even whether it fails to belongs to the second machine class. ${ }^{13}$

Per van Emde Boas's prior admonishment, it is hence unclear whether $\mathfrak{G}$ may be directly employed in a squeezing argument for feasibility in the envisioned manner. As such, we have yet to find an argument which puts the CobhamEdmonds Thesis on the same footing as Church's Thesis. We are, however, left with a potentially promising question to explore in this regard - i.e. is it possible to find natural modifications to the Gandy-Sieg boundness and locality conditions which yield a model of computation which is still Turing complete but which may also be shown to satisfy the Invariance Thesis with respect to the reference model $\mathfrak{R}$ ?

[^15]
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# Recent Advances in Non-perfect Secret Sharing Schemes 

Oriol Farràs ${ }^{(\boxtimes)}$<br>Universitat Rovira i Virgili, Tarragona, Catalonia, Spain<br>oriol.farras@urv.cat


#### Abstract

A secret sharing scheme is non-perfect if some subsets of players that cannot recover the secret have partial information about it. This paper is a survey of the recent advances in non-perfect secret sharing schemes. We provide an overview of the techniques for constructing efficient non-perfect secret sharing schemes, bounds on the efficiency of these schemes, and results on the characterization of the ideal ones. We put special emphasis on the connections between non-perfect secret sharing schemes and polymatroids, matroids, information theory, and coding theory.


Keywords: Secret sharing • Non-perfect secret sharing • Ideal secret sharing scheme • Information ratio • Polymatroid • Matroid

## 1 Introduction

A secret sharing scheme is a cryptographic primitive that is used to protect a secret value by dividing it into shares. Secret sharing is used to prevent both the disclosure and the loss of the secret value. In the typical scenario, each share is sent privately to a different player. Then a subset of players is qualified if their shares determine the secret value, and forbidden if their shares do not contain any information on the secret value. In this work we just consider schemes that are information-theoretically secure, and so their security does not rely on any computational assumption. A secret sharing scheme is perfect if every set of players is qualified or forbidden.

Blakley [3] and Shamir [31] presented the first secret sharing schemes. These schemes are perfect, and are called threshold secret sharing schemes because the qualified subsets are those whose size is greater than a certain threshold. In a perfect secret sharing scheme, the length in bits of each share must be greater than or equal to the length of the secret. The schemes attaining this bound, as the ones of Blakley and Shamir [3,31], are called ideal.

In order to circumvent this restriction on the length of the shares, Blakley and Meadows [4] relaxed the privacy requirements of the threshold schemes.
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They presented the first non-perfect schemes, which were called ramp secret sharing schemes. The structure of a ramp scheme is described by two thresholds $t$ and $r$ with $t<r$ : Every subset with at most $t$ players is forbidden, every subset with at least $r$ players is qualified, and the other subsets have partial information about the secret value. The length of every share is $1 /(r-t)$ times the length of the secret, which is also optimal [28]. The threshold and the ramp schemes in $[3,4,31]$ are linear, that is, they can be described in terms of linear maps over a finite field. Because of their efficiency and their homomorphic properties, linear secret sharing schemes play a fundamental role in several areas of cryptography such as secure multiparty computation and distributed cryptography (see [1]).

Blakley and Meadows showed that if we allow some subsets to obtain partial information about the secret value, we can then construct schemes with shorter shares. Their construction motivated the search for efficient non-perfect secret sharing schemes, the exploration of their limitations, and their use in cryptographic applications. The connections between information theory, matroid theory and secret sharing found by Fujishige [20,21], Brickell and Davenport [5], and Csirmaz [13], are powerful tools in the study of perfect secret sharing schemes. The extension of these connections to the non-perfect case was initiated twenty years ago [25, 27,30], and has been improved in recent works [15-18, 32,33]. Many other works such as $[7,8,10,11,14,22,23]$ presented efficient non-perfect secret sharing schemes with interesting cryptographic applications.

This paper is a survey of the recent advances in non-perfect secret sharing schemes. Section 2 is dedicated to the definition of secret sharing and the main efficiency measures. The main constructions of non-perfect schemes are presented in Sect. 3, putting special emphasis on the linear ones. We present the connection between secret sharing schemes and polymatroids and the resulting lower bounds in Sect.4. Finally, in Sect. 5, we present the results on ideal secret sharing schemes and their connection with matroids.

## 2 Secret Sharing Schemes

In this survey we use an information-theoretic definition of secret sharing that was introduced in $[15,16]$. This definition is very general and covers both perfect and non-perfect secret sharing schemes. The secret and the shares of the scheme are defined as discrete random variables, and the information known about the secret is measured using the Shannon entropy function. Before presenting the definition, we need to introduce some notation.

Given a discrete random vector $S=\left(S_{i}\right)_{i \in Q}$ and a set $X \subseteq Q$, the Shannon entropy of the random variable $S_{X}=\left(S_{i}\right)_{i \in X}$ is denoted by $H\left(S_{X}\right)$. In addition, for such random variables, one can consider the mutual information $I\left(S_{X}: S_{Y}\right)=$ $H\left(S_{X}\right)+H\left(S_{Y}\right)-H\left(S_{X \cup Y}\right)$. The reader is referred to [12] for a textbook on information theory, and to [1,29] for a survey on secret sharing.

Definition 1. Let $Q$ be a finite set of players, let $p_{0} \in Q$ be a distinguished player, which is called dealer, and take $P=Q \backslash\left\{p_{0}\right\}$. A secret sharing scheme
on the set $P$ is a discrete random vector $\Sigma=\left(S_{i}\right)_{i \in Q}$ such that $H\left(S_{p_{0}}\right)>0$ and $H\left(S_{p_{0}} \mid S_{P}\right)=0$.

In this definition, the random variable $S_{p_{0}}$ corresponds to the secret value, while the random variables $\left(S_{i}\right)_{i \in P}$ correspond to the shares of the secret value of the players in $P$. We use the access function of the scheme $[15,16]$ to describe the amount of information that the shares provide about the secret value.

Definition 2. An access function on a set $P$ is a monotone increasing function $\Phi: \mathcal{P}(P) \rightarrow[0,1]$ with $\Phi(\emptyset)=0$ and $\Phi(P)=1$.

Definition 3. The access function $\Phi$ of a secret sharing scheme $\Sigma=\left(S_{i}\right)_{i \in Q}$ is the map $\Phi: \mathcal{P}(P) \rightarrow[0,1]$ defined by

$$
\Phi(X)=\frac{I\left(S_{p_{0}}: S_{X}\right)}{H\left(S_{p_{0}}\right)}
$$

If $\Phi(X)=1$, then $I\left(S_{p_{0}}: S_{X}\right)=H\left(S_{p_{0}}\right)$, which implies that the secret value is determined by the shares of the players in $X$. If $\Phi(X)=0$, the random variables $S_{p_{0}}$ and $S_{X}$ are independent and the shares of the players in $X$ therefore do not provide any information about the secret. For this reason, a set $X \subseteq P$ is called forbidden for $\Phi$ if $\Phi(X)=0$, and it is called qualified for $\Phi$ if $\Phi(X)=1$. If every subset is forbidden or qualified, we say that the scheme is perfect. Accordingly, an access function is said to be perfect if it only takes the values 0 and 1 . The gap of an access function $\Phi$ is defined as the minimum distance between a qualified and a forbidden subset, that is, $\operatorname{gap}(\Phi)=\min \{|Y \backslash X|: \Phi(X)=0, \Phi(Y)=1\}$. We say that an access function is rational if it takes values only in $\mathbb{Q}$.

As a measure of the efficiency of a scheme, we consider the ratio of the maximum length in bits of the shares to the length of the secret value. We use the Shannon entropy as an approximation of the shortest binary codification.

Definition 4. The information ratio $\sigma(\Sigma)$ of a secret sharing scheme $\Sigma=$ $\left(S_{i}\right)_{i \in Q}$ is defined by

$$
\sigma(\Sigma)=\frac{\max _{i \in P} H\left(S_{i}\right)}{H\left(S_{p_{0}}\right)}
$$

There exists a secret sharing scheme for every access function [15]. However, the general constructions provide schemes that are not efficient for most of the access functions. Namely, for most of the access functions, the information ratios of the resulting schemes are exponential in the number of players. This is also the situation in perfect secret sharing schemes [2]. The need of more efficient schemes motivated the search for lower bounds on the information ratio. The optimal information ratio $\sigma(\Phi)$ of an access function $\Phi$ is the infimum of the information ratios of the secret sharing schemes realizing $\Phi$. We say that a secret sharing scheme for $\Phi$ is optimal if its information ratio attains $\sigma(\Phi)$.

## 3 Constructions

In this section we present the main non-perfect secret sharing constructions. We present the family of linear secret sharing schemes, the constructions for uniform access functions, and a general method to construct a secret sharing scheme for any access function.

### 3.1 Linear Secret Sharing Schemes

Definition 5. Let $\mathbb{K}$ be a finite field and let $\ell$ be a positive integer. In a ( $\mathbb{K}, \ell$ )linear secret sharing scheme, the random variables $\left(S_{i}\right)_{i \in Q}$ are given by surjective $\mathbb{K}$-linear maps $S_{i}: E \rightarrow E_{i}$, where the uniform probability distribution is taken on $E$ and $\operatorname{dim} E_{p_{0}}=\ell$.

We present the linear secret sharing scheme of Blakley and Meadows [4] in Example 1. In the original paper, the scheme was presented using polynomial notation.

Example 1. Let $P=\{1, \ldots, n\}$, and let $t, r$ be integers with $0 \leq t<r \leq$ $n$. Define $g=r-t$. Let $\mathbb{K}$ be a finite field with $|\mathbb{K}| \geq n+g$, and let $y_{1}, \ldots, y_{g}, x_{1}, \ldots, x_{n}$ be different elements in $\mathbb{K}$. Define $\mathbf{x}_{i}=\left(1, x_{i}, \ldots, x_{i}^{r-1}\right) \in$ $\mathbb{K}^{r}$ for $i \in P$ and $\mathbf{y}_{i}=\left(1, y_{i}, \ldots, y_{i}^{r-1}\right) \in \mathbb{K}^{r}$ for $i=1, \ldots, g$. Consider the secret sharing scheme $\Sigma=\left(S_{i}\right)_{i \in Q}$ determined by the linear mappings
$-S_{p_{0}}: \mathbb{K}^{r} \rightarrow \mathbb{K}^{g}: \mathbf{u} \mapsto\left(\mathbf{u} \cdot \mathbf{y}_{1}, \ldots, \mathbf{u} \cdot \mathbf{y}_{g}\right)$, and
$-S_{i}: \mathbb{K}^{r} \rightarrow \mathbb{K}: \mathbf{u} \mapsto \mathbf{u} \cdot \mathbf{x}_{i}$ for $i \in P$,
and take the uniform probability distribution on $\mathbb{K}^{r}$.
In a $\mathbb{K}$-linear secret sharing scheme $\left(S_{i}\right)_{i \in Q}$, the random variable $S_{X}$ is uniform on its support for every $X \subseteq Q$. Because of that, $H\left(S_{X}\right)=\log |\mathbb{K}| \cdot \operatorname{rank} S_{X}$, and hence $I\left(S_{p_{0}}: S_{X}\right)=\log |\mathbb{K}| \cdot\left(\operatorname{rank} S_{p_{0}}+\operatorname{rank} S_{X}-\operatorname{rank} S_{X \cup\left\{p_{0}\right\}}\right)$. Therefore, in order to determine the access function of a linear scheme, we just need to compute the ranks of the linear mappings. Taking into account the properties of the Vandermonde matrix, the computation of the access function $\Phi$ of the scheme $\Sigma$ in Example 1 is straightforward. It can be described as follows.
$-\Phi(X)=0$ if $|X| \leq t$,
$-\Phi(X)=(|X|-t) / g$ if $t<|X|<r$, and
$-\Phi(X)=1$ if $|X| \geq r$.
This access function is called the $(t, r)$-ramp access function, and the schemes with access functions of this kind are called ramp secret sharing schemes. The scheme is $(\mathbb{K}, g)$-linear, it has information ratio $1 / g$, and it is optimal [28]. This scheme is an extension of the threshold secret sharing scheme of Shamir [31], which corresponds to the case $g=1$. A scheme has $t$-privacy if the subsets of size at most $t$ are forbidden, and it has $r$-reconstruction if the subsets of size greater than or equal to $r$ are qualified. The $(t, r)$-ramp secret sharing schemes
have $t$-privacy, $r$-reconstruction, and the subsets of size greater than $t$ and smaller than $r$ obtain a certain amount of information about the secret, which increases at a constant rate $1 / g$.

Linear secret sharing schemes have homomorphic properties that are useful for many cryptographic applications. If we share two secrets $s_{1}$ and $s_{2}$ among a set of players using a linear scheme $\Sigma$, each player can obtain a share of $s_{1}+s_{2}$ by just computing the sum of the two shares he holds. Namely, each of these sums is a share of $s_{1}+s_{2}$ by $\Sigma$. Moreover, the sharing and reconstruction operations of linear secret sharing schemes are efficient [16]. The reconstruction of the secret value by a set of players can be obtained by solving the linear equations defined by their shares.

We can describe the shares and the secret of the scheme in Example 1 as evaluations of a polynomial of degree at most $r-1$ at different points. Hence, the secret and the shares are elements of a codeword of a Reed-Solomon code. This connection can be generalized, and we can obtain linear secret sharing schemes from any linear code. Roughly speaking, given a linear code, we can define linear secret sharing schemes in which the shares and the secret are subsets of coordinates of the codewords (see $[7,16]$ for more details). For many cryptographic applications, the requirements on the secret sharing schemes are $t$-privacy and $r$-reconstruction for certain values of $t$ and $r$, and there are no restrictions for the subsets of size between $t$ and $r$. In this case, the construction of schemes from linear codes is especially relevant, because $r$ and $t$ can be estimated from the minimum distance of the code and from the minimum distance of its dual, respectively [7].

The scheme in Example 1 has the restriction that $|\mathbb{K}|$ must be greater than or equal to $n+g$. It may be a drawback if we want to share a short secret among a large set of players. Several works $[7,8,22,23]$ study this problem and present non-perfect linear schemes that are defined over smaller fields that can substitute the ramp schemes. They are constructed from algebraic geometry codes and random codes, and have multiplicative properties, such as the one in Example 1. Multiplicative non-perfect schemes are building blocks of several secure multiparty computation protocols $[7,11,19]$. Linear codes have also been used to construct linear schemes with more efficient sharing and secret reconstruction [10, 14].

### 3.2 Uniform Access Functions

Ramp access functions belong to the family of uniform access functions. We say that an access function is uniform if its values only depend on the size of the subset. Yoshida et al. [33] found optimal linear secret sharing schemes for every rational uniform access function. Farràs et al. [15] provided an alternative proof using different techniques, as well as constructions for the non-rational case. The constructions in both works are based on the following observation. Let $\Sigma_{1}$ be a $\left(\mathbb{K}, \ell_{1}\right)$-linear scheme and let $\Sigma_{2}$ be a $\left(\mathbb{K}, \ell_{2}\right)$-linear scheme for some $\ell_{1}$ and $\ell_{2}$. Let $\Phi_{1}$ and $\Phi_{2}$ be the respective access functions. If we take a secret $s \in \mathbb{K}^{\ell_{1}+\ell_{2}}$ and we share the first $\ell_{1}$ coordinates with $\Sigma_{1}$ and the remaining $\ell_{2}$
coordinates with $\Sigma_{2}$, we obtain a $\left(\mathbb{K}, \ell_{1}+\ell_{2}\right)$-linear secret sharing scheme whose access function is a convex combination of $\Phi_{1}$ and $\Phi_{2}$, and its information ratio is also a convex combination of $\sigma\left(\Sigma_{1}\right)$ and $\sigma\left(\Sigma_{2}\right)$. By combining ramp secret sharing schemes in this way, it is possible to construct optimal schemes for every uniform access function $[15,33]$. A drawback of these optimal schemes is that the secret can be very large for certain access functions. Namely, the schemes are $(\mathbb{K}, \ell)$-linear for a reasonably large finite field $\mathbb{K}$ and attain the optimal information ratio, but $\ell$ can be very large. Indeed, as in the perfect case, the secret sharing schemes with optimal information ratio are not necessarily the most efficient ones. Constructions of non-optimal linear secret sharing schemes with shorter shares for every rational uniform access functions were presented in $[15,32]$.

This problem leads to the search for bounds on the length the shares instead of bounds on the information ratio. For ramp access functions, constructing linear secret sharing schemes over finite fields of size smaller than $n+r-t-1$ is an open problem that is related to the maximum distance separable codes conjecture. Cascudo, Cramer and Xing [6] studied the family of secret sharing schemes with $t$-privacy and $r$-reconstruction. They found general lower bounds on the length of the shares that only depend on $t, r$ and $n$; and specific lower bounds for ( $\mathbb{K}, \ell$ )-linear secret sharing schemes with information ratio $1 / \ell$.

### 3.3 General Constructions

The access function of a linear secret sharing scheme is a rational function. Conversely, every rational access function admits a linear secret sharing scheme.

Theorem 1 [15]. Every access function admits a secret sharing scheme. Moreover, every rational access function admits a $(\mathbb{K}, \ell)$-linear secret sharing schemes for every finite field $\mathbb{K}$ and large enough $\ell$.

Next, we present a sketch the proof of this theorem, which is constructive. For a rational access function $\Phi$, the secret of the scheme is set to be in $\mathbb{K}^{\ell}$, where $\ell$ is determined by $\Phi$. The smaller the minimum of the non-zero values of $|\Phi(X \cup\{y\})-\Phi(X)|$ for $X \subseteq P$ and $y \notin P \backslash X$ is, the larger $\ell$ is. For every subset of players $X \subseteq P$, we define a secret sharing scheme $\Sigma_{X}$ that shares among $X$ a part of the secret that is proportional to $\Phi(X)$. Then, for every $X$ in a certain family of subsets $\mathcal{A} \subseteq \mathcal{P}(P)$, the secret is shared independently by means of $\Sigma_{X}$. Access functions with non-rational values require constructions that are not linear. The general construction in [15] follows the same idea, but the schemes $\Sigma_{X}$ are substituted by schemes in which the distribution of the secret, which is chosen in a special way, may be not uniform.

Ishai et al. [24] studied the representation of rational access functions for a fixed secret size. By using Markov chain techniques, they presented non-linear constructions for every uniform rational access function, and another construction for every rational access function. For every perfect access function $\Phi$, Benaloh and Leichter [2] presented a way to transform any monotone Boolean
formula for $\Phi$ to a secret sharing scheme for $\Phi$. This construction has not been extended to the non-perfect case yet.

## 4 Lower Bounds on the Information Ratio

Despite the limitations described above, the information ratio is the most common parameter for the study of the efficiency of secret sharing schemes. Many results on the information ratio are based on information inequalities, including Shannon inequalities, rank inequalities, and Non-Shannon inequalities. Csirmaz [13] found a connection between perfect secret sharing schemes and polymatroids that translates the search for the optimal information ratio into combinatorial and linear programming problems. The description of the structure of non-perfect secret sharing schemes by means of their access function allowed to extend these techniques to the non-perfect case [15].

Definition 6. A polymatroid is a pair $\mathcal{S}=(Q, f)$ formed by a finite set $Q$, the ground set, and a rank function $f: \mathcal{P}(Q) \rightarrow \mathbb{R}$ satisfying the following properties.
$-f(\emptyset)=0$.

- $f$ is monotone increasing: if $X \subseteq Y \subseteq Q$, then $f(X) \leq f(Y)$.
- $f$ is submodular: $f(X \cup Y)+f(X \cap Y) \leq f(X)+f(Y)$ for every $X, Y \subseteq Q$.

If $f$ is integer-valued and $f(X) \leq|X|$ for every $X \subseteq Q$, then $\mathcal{S}$ is called a matroid.

Fujishige $[20,21]$ showed that for every random vector $\left(S_{i}\right)_{i \in Q}$, the map $h: \mathcal{P}(Q) \rightarrow \mathbb{R}$ defined by $h(X)=H\left(S_{X}\right)$ is the rank function of a polymatroid with ground set $Q$. Since secret sharing schemes are random vectors, every secret sharing scheme defines a polymatroid. Conversely, it is not known which polymatroids are defined by secret sharing schemes. From certain polymatroids we can define an access function as follows.

Definition 7. Let $\Phi$ be an access function on $P$ and let $\mathcal{S}=(Q, f)$ be a polymatroid. Then $\mathcal{S}$ is a $\Phi$-polymatroid if

$$
\Phi(X)=\frac{f\left(p_{0}\right)+f(X)-f\left(\left\{p_{0}\right\} \cup X\right)}{f\left(p_{0}\right)}
$$

for every $X \subseteq P$.
The polymatroid defined by a secret sharing scheme $\Sigma$ with access function $\Phi$ is a $\Phi$-polymatroid. For perfect access functions $\Phi$, Martí-Farré and Padró [26] found a lower bound on $\sigma(\Phi)$ that is called $\kappa(\Phi)$ and that is based on the connection between secret sharing schemes and polymatroids. The parameter $\kappa(\Phi)$ was later generalized to the non-perfect case in [15] as follows.
Definition 8. For every access function $\Phi$ on $P$, we define

$$
\kappa(\Phi)=\inf \left\{\sigma_{0}(\mathcal{S}): \mathcal{S} \text { is a } \Phi \text {-polymatroid }\right\}
$$

where $\sigma_{0}(\mathcal{S})=\max _{x \in P} f(x) / f\left(p_{0}\right)$.

Note that for every secret sharing scheme $\Sigma$, the polymatroid $\mathcal{S}$ associated to $\Sigma$ satisfies $\sigma_{0}(\mathcal{S})=\sigma(\Sigma)$. Hence, $\kappa(\Phi) \leq \sigma(\Phi)$ for every access function $\Phi$. In the next lemma, we present a basic lower bound on $\kappa[15,27,30]$, which is attained by the scheme in Example 1.
Lemma 1. For every access function $\Phi, \kappa(\Phi) \geq 1 / \operatorname{gap}(\Phi)$.
The search for general lower bounds and the search for ideal access functions (see Sect. 5) induced the study of the family of the access functions with constant increment. In the next theorem we present the best known lower bound on the information ratio of secret sharing schemes [16], which is an extension of a result of Csirmaz [13, Theorem 3.2].
Definition 9. An access function $\Phi$ on $P$ has constant increment $\mu$ if $\Phi(X \cup$ $\{y\})-\Phi(X) \in\{0, \mu\}$ for every $X \subset P$ and $y \in P$. In this situation, $\mu=1 / k$ for some positive integer $k$.
Theorem 2. For every positive integer $k$ and for infinitely many positive integers $n$, there exists an access function $\Phi_{n}$ with constant increment $1 / k$ on a set of size $n+k-1$ satisfying $\kappa\left(\Phi_{n}\right) \geq n /(2 k \log n)$.

For every uniform access function $\Phi, \kappa(\Phi)=\sigma(\Phi)[9,15]$. However, in general, $\kappa$ is not a tight bound on $\sigma$. Indeed, $\kappa(\Phi) \leq n$ for every access function $\Phi$ [16]. Up to now, all the lower bounds on the information ratio of non-perfect secret sharing schemes are a consequence of the Shannon information inequalities.

## 5 Ideal Secret Sharing Schemes

A perfect secret sharing scheme $\Sigma$ is ideal if $\sigma(\Sigma)=1$. In this case, we say that its access function is also ideal. Brickell and Davenport [5] proved that every ideal perfect secret sharing scheme determines a matroid. Because of this result, matroids became a crucial tool for the study and for the construction of efficient secret sharing schemes. The connection between non-perfect secret sharing schemes and matroids, and the notion of ideality in the non-perfect case have been studied in $[16-18,25,30]$. The following definition of ideal secret sharing scheme was introduced in [16], and it is equivalent to the definitions in $[25,30]$.
Definition 10. A secret sharing scheme $\Sigma=\left(S_{i}\right)_{i \in Q}$ is ideal if its access function has constant increment $1 / k$ for some $k>0$ and $H\left(S_{i}\right)=H\left(S_{p_{0}}\right) / k$ for every $i \in P$. In this case, its access function is called ideal as well.

If $\Phi$ is an ideal access function, then $\Phi$ has constant increment $1 / \operatorname{gap}(\Phi)$. Next, we extend the notion of matroid port.

Definition 11. Let $P, P_{0}$ be a pair of nonempty, disjoint sets and $\mathcal{M}=(P \cup$ $\left.P_{0}, r\right)$ a matroid such that $r\left(P_{0}\right)=\left|P_{0}\right|=$ and $r(P)=r\left(P \cup P_{0}\right)$. The port of the matroid $\mathcal{M}$ at the set $P_{0}$ is the access function $\Phi$ on $P$ defined by

$$
\Phi(X)=\frac{r\left(P_{0}\right)+r(X)-r\left(P_{0} \cup X\right)}{r\left(P_{0}\right)}
$$

Now we can state the extension of the Brickell and Davenport Theorem [5] to the non-perfect case [18].

Theorem 3. The access function of every ideal secret sharing scheme is a generalized matroid port.

If $\phi$ is a port of a matroid $\mathcal{M}=\left(P \cup P_{0}, r\right)$ at $P_{0}$, and $\mathcal{M}$ is $\mathbb{K}$-linearly representable for a finite field $\mathbb{K}$, then $\phi$ admits an ideal linear secret sharing scheme (see [16]). However, as in the perfect case, not all matroid ports admit ideal secret sharing schemes. The characterization of the matroid ports that are ideal access functions is an open problem. If an access function $\Phi$ has constant increment $k$ and it is not a matroid port at a set of size $k$, then $\kappa(\Phi) \geq 3 /(2 k)$ [17].

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# A Computational Approach to the Borwein-Ditor Theorem 

Aleksander Galicki and André Nies ${ }^{(\boxtimes)}$<br>Department of Computer Science, University of Auckland, Auckland, New Zealand andre@cs.auckland.ac.nz


#### Abstract

Borwein and Ditor (Canadian Math. Bulletin 21 (4), 497$498,1978)$ proved the following. Let $\mathcal{A} \subset \mathbb{R}$ be a measurable set of positive measure and let $\left\langle r_{m}\right\rangle_{m \in \omega}$ be a null sequence of real numbers. For almost all $z \in \mathcal{A}$, there is $m$ such that $z+r_{m} \in \mathcal{A}$.

In this note we mainly consider the case that $\mathcal{A}$ is $\Pi_{1}^{0}$ and the null sequence $\left\langle r_{m}\right\rangle_{m \in \omega}$ is computable. We show that in this case every Oberwolfach random real $z \in \mathcal{A}$ satisfies the conclusion of the theorem. We extend the result to finitely many null sequences. The conclusion is now that for almost every $z \in \mathcal{A}$, the same $m$ works for each null sequence.

We indicate how this result could separate Oberwolfach randomness from density randomness.


## 1 Introduction

Our paper is based on the following result, which extends a previous weaker result by Kestelman [11].

Theorem 1 (D. Borwein and S. Z. Ditor [4], Theorem 1(i)). Suppose $\mathcal{A} \subset \mathbb{R}$ is a measurable set of positive measure and $\left\langle r_{m}\right\rangle_{m \in \omega}$ is a sequence of real numbers converging to 0 . For almost all $z \in \mathcal{A}$, there is an $m$ such that $z+r_{m} \in \mathcal{A}$.

Since one can consider the tails of a given null sequence of reals, for almost every $z \in \mathcal{A}$ there are in fact infinitely many $m$ such that $z+r_{m} \in \mathcal{A}$. (This is the form in which they actually stated the result). We thank Ostaszewski for pointing out the Borwein-Ditor theorem to Nies during his visit at the London School of Ecomonics in June 2015. Ostaszewski's 2007 book with Bingham provides some background related to topology [3].

Nies' colloquium at LSE was about the study of effective versions of "almost everywhere" theorems via algorithmic randomness. The goal for that direction of study is to pin down the level of algorithmic randomness needed for a point $x$ so that the conclusion of a particular effective version of the theorem holds. For instance, Pathak et al. [15] study effective versions of the Lebesgue differentiation theorem, Brattka et al. [5] look at the a.e. differentiability of nondecreasing functions, Galicki and Turetsky [10] study the a.e. differentiability of Lipschitz functions on $\mathbb{R}^{n}$, and Miyabe et al. [13] consider the Lebesgue density theorem, recalled in Theorem 4 below.

Unless stated otherwise, we will consider effectively closed (i.e. $\Pi_{1}^{0}$ ) sets $\mathcal{A} \subseteq$ $\mathbb{R}$. Without also imposing an effectiveness condition on the null sequences, the points $z$ for which the Borwein-Ditor property holds for all $\Pi_{1}^{0}$ sets are precisely the 1-generics. Recall that a real is 1-generic if it is not on the boundary of any $\Sigma_{1}^{0}$ set.

Proposition 2. $z \in \mathbb{R}$ is 1-generic $\Longleftrightarrow$ for every $\Pi_{1}^{0}$ set $\mathcal{A} \subset \mathbb{R}$ containing $z$ and for every null sequence of real numbers $\left\langle r_{m}\right\rangle_{m \in \omega}, z \in \mathcal{A}+r_{m}$ for some $m$.

Proof. $(\Rightarrow)$ Suppose $z \notin \mathcal{A}+r_{m}$ for all $m$. Then $z$ belongs to the boundary of the complement of $\mathcal{A}, B=\mathbb{R} \backslash \mathcal{A}$, a $\Sigma_{1}^{0}$ class.
$(\Leftarrow)$ Suppose $z \in \mathbb{R}$ is not 1-generic. Then it belongs to the boundary of some $\Sigma_{1}^{0}$ set $\mathcal{B} \subset \mathbb{R}$. Let $\left\langle z_{m}\right\rangle_{m \in \omega}$ be a sequence of points in $\mathcal{B}$ converging to $z$. Define $r_{m}=z-z_{m}$ for all $m$. Then $\mathcal{A}=\mathbb{R} \backslash \mathcal{B}$ is a $\Pi_{1}^{0}$ class, $z \in \mathcal{A},\left\langle r_{m}\right\rangle_{m \in \omega}$ is a sequence converging to 0 , and $z \notin \mathcal{A}+r_{m}$ for all $m$.

## 2 Comparison of Lebesgue Density and the Borwein-Ditor Property

The definitions below follow [2]. Let $\lambda$ denote Lebesgue measure on $\mathbb{R}$.
Definition 3. We define the lower Lebesgue density of a set $\mathcal{C} \subseteq \mathbb{R}$ at a point $z$ to be the quantity

$$
\varrho(\mathcal{C} \mid z):=\liminf _{\gamma, \delta \rightarrow 0^{+}} \frac{\lambda([z-\gamma, z+\delta] \cap \mathcal{C})}{\gamma+\delta} .
$$

Note that $0 \leq \underline{\varrho}(\mathcal{C} \mid z) \leq 1$.
Theorem 4 (Lebesgue [12]). Let $\mathcal{C} \subseteq \mathbb{R}$ be a measurable set. Then $\underline{\varrho}(\mathcal{C} \mid z)=1$ for almost every $z \in \mathcal{C}$.

The Borwein-Ditor theorem is analogous to the Lebesgue density theorem. Both results say that for almost every point in a measurable class there are, in a specific sense, many arbitrarily close other points in the class.

An open set $\mathcal{C}$ clearly has lower Lebesgue density 1 at each of its members. Thus, the simplest non-trivial case is when $\mathcal{C}$ is closed. We say that a real $z \in[0,1]$ is a density-one point if $\varrho(\mathcal{C} \mid z)=1$ for every effectively closed class $\mathcal{C}$ containing $z$. Similar to the implication $(\Rightarrow)$ of Proposition 2, every 1-generic is a densityone point. So being a density-one point is by itself not a randomness notion, and neither is the Borwein-Ditor property for effectively closed sets. In both cases, to remedy this one has to add as an additional condition that the real is Martin-Löf random.

Definition 5. Let $z \in \mathbb{R}$ be $M L$-random. We say that $z$ is density random if $z$ is also a density-one point.

Definition 6. Let $z$ be ML-random. We say that $z$ is Borwein-Ditor (BD) random if for each $\Pi_{1}^{0}$ set $\mathcal{A} \subseteq \mathbb{R}$ with $z \in \mathcal{A}$, and each computable null sequence of reals $\left\langle r_{m}\right\rangle_{m \in \omega}$, there is an $m$ such that $z+r_{m} \in \mathcal{A}$.

Neither of the two randomness notions is equivalent to ML-randomness, because the least element of a non-empty effectively closed set of ML-randoms is neither density random nor BD-random. Density randomness is in fact known to be stronger than difference randomness (i.e. ML-randomness together with Turing incompleteness) by Bienvenu et al. [2] together with Day and Miller [6]. Much less is known at present about the placement of BD-randomness within the established notions.

## 3 Oberwolfach Randomness Implies BD Randomness

Ah! the ancient pond
as a frog takes the plunge
sound of water
(Matsuo Basho)
To simplify notation, we identify the unit interval with Cantor space ${ }^{\omega} 2$ in what follows, ignoring dyadic rationals. For a string $\sigma$, as usual by $[\sigma]$ we denote the corresponding basic dyadic interval; for example [101] denotes the interval [5/8, 3/4].

Bienvenu et al. [1] introduced Oberwolfach (OW) randomness, and also gave the following equivalent definition. A left-c.e. bounded test is a descending sequence $\left\langle\mathcal{V}_{n}\right\rangle$ of uniformly $\Sigma_{1}^{0}$ classes in Cantor space such that for some nondecreasing computable sequence of rationals $\left\langle\beta_{s}\right\rangle$ with $\beta=\sup _{s} \beta_{s}<\infty$, we have $\lambda\left(\mathcal{V}_{n}\right) \leq \beta-\beta_{n}$ for all $n$. $Z$ is $O W$-random iff $Z$ passes each such test in the sense that $Z \notin \bigcap_{n} \mathcal{V}_{n}$.

OW-randomness implies density randomness [1]. The converse implication remains unknown. The question is intriguing. All the equivalent characterizations of OW-randomness are within the, by now, almost classical framework of computability and randomness $[7,14]$. For instance, if $Z$ is ML-random, $Z$ is OW-random iff it does not compute every $K$-trivial set [1]. On the other hand, the seemingly very close notion of density randomness is defined analytically, and mainly has analytical characterizations such as via differentiability of interval-c.e. functions in [13, Theorem 4.2].

Using Theorem 1 it is easy to check that weak 2-randomness implies BDrandomness. We show that the much weaker notion of OW-randomness already implies BD-randomness.

While density and BD randomness are analogous, it seems unlikely that density implies BD . This provides evidence that OW-randomness is strictly stronger than density randomness.

Theorem 7. Let $Z$ be Oberwolfach random. Then $Z$ is $B D$ random.
Proof. Suppose we are given a $\Pi_{1}^{0}$ class $\mathcal{P} \subseteq{ }^{\omega} 2$ with $Z \in \mathcal{P}$, and a computable null sequence of reals $\left\langle r_{m}\right\rangle_{m \in \omega}$. We may assume that $r_{m} \leq 2^{-m}$. Let $\left\langle\sigma_{m}\right\rangle_{m \in \omega}$ be a computable prefix-free sequence of strings such that $\mathcal{S}={ }^{\omega} 2 \backslash \mathcal{P}=\left[\left\{\sigma_{m}: m \in\right.\right.$ $\mathbb{N}\}]^{\prec}$, and let $\mathcal{S}_{m}=\left[\sigma_{0}, \ldots, \sigma_{m-1}\right]^{\prec}$, the class of all bit sequences extending one of the $\sigma_{i}$. Let $q(m)=1+\max \left(m, \max _{i<m}\left|\sigma_{m}\right|\right)$. Now define a left-c.e. bounded test by

$$
G_{m}=\bigcap_{i \leq q(m)}\left(\mathcal{S}+r_{m}\right) \backslash \mathcal{S}_{m}
$$

Clearly this is a descending sequence of uniformly $\Sigma_{1}^{0}$ sets. Let

$$
\beta=\lambda \mathcal{S} \text { and } \beta_{m}=\lambda \mathcal{S}_{m}-m 2^{-m}
$$

so that $\beta=\sup _{m} \beta_{m}$.
Claim. $\lambda G_{m} \leq \beta-\beta_{m}$.
We actually show this bound for $\mathcal{S}+r_{q(m)} \backslash \mathcal{S}_{m}$ instead of $G_{m}$. Since $A \backslash C \subseteq$ $(A \backslash B) \cup(B \backslash C)$ for sets $A, B, C$, and by the translation invariance of $\lambda$,

$$
\lambda\left(\mathcal{S}+r_{q(m)} \backslash \mathcal{S}_{m}\right) \leq \lambda\left(\mathcal{S} \backslash \mathcal{S}_{m}\right)+\lambda\left(\mathcal{S}_{m}+r_{q(m)} \backslash \mathcal{S}_{m}\right)
$$

Recall that $r_{k} \leq 2^{k}$. Hence by definition of $q(m)$, for each $i<m$ we have

$$
\lambda\left(\left[\sigma_{m}\right]+r_{q(m)} \backslash\left[\sigma_{m}\right]\right) \leq 2^{-q(m)}
$$

Therefore $\lambda\left(\mathcal{S}_{m}+r_{q(m)} \backslash \mathcal{S}_{m}\right) \leq m 2^{-q(m)} \leq m 2^{-m}$ as required for the claim. If $Z+r_{n} \notin \mathcal{P}$ for each $n$ then $Z \in \bigcap_{m} G_{m}$, so $Z$ is not OW-random.

We note that this proof works in much greater generality for an abelian group $(S,+)$ that is also a computable probability space $(S, \mu)$ with a translation invariant measure, such that $\lim _{r \rightarrow 0} \mu((A+r) \triangle A)=0$ effectively for every basic open set $A$. For instance, the general theorem also applies to Cantor space with the usual ultrametric and the group structure of the 2-adic integers $\left(\mathbb{Z}_{2},+\right)$.

Finally, similar to [4] we extend the foregoing theorem to the case of finitely many null sequences, and show that for an OW-random $Z$, one position works for all of them. This is in the spirit of multiple recurrence in ergodic theory, initiated by Furstenberg and others in the 1970s [9].

Theorem 8. Let $Z$ be Oberwolfach random. For each $\Pi_{1}^{0}$ class $\mathcal{P} \subseteq \omega^{\omega}$ with $Z \in \mathcal{P}$, and $k$ many computable null sequences of reals $\left\langle r_{m, v}\right\rangle_{m \in \omega}, 0 \leq v<k$, there is $m$ such that $Z+r_{m, v} \in \mathcal{P}$ for each $v<k$.

Proof. We may assume that $r_{m, v} \leq 2^{-m}$ for each $v$. Let $\left\langle\sigma_{m}\right\rangle_{m \in \omega}, \mathcal{S}_{m}, q(m)$ and $\beta_{m}$ be defined as above. Let

$$
G_{m}=\bigcap_{i \leq q(m)} \bigcup_{v<k}\left(\mathcal{S}+r_{i, v}\right) \backslash \mathcal{S}_{m}
$$

Claim. $\lambda G_{m} \leq k\left(\beta-\beta_{m}\right)$.
We actually show this bound for $\bigcup_{v<k}\left(\mathcal{S}+r_{q(m), v}\right) \backslash \mathcal{S}_{m}$ instead of $G_{m}$. By the translation invariance of $\lambda$,

$$
\lambda \bigcup_{v<k}\left(\mathcal{S}+r_{q(m), v}\right) \backslash \mathcal{S}_{m} \leq k \lambda\left(\mathcal{S} \backslash \mathcal{S}_{m}\right)+\lambda\left(\bigcup_{v<k} \mathcal{S}_{m}+r_{q(m), v} \backslash \mathcal{S}_{m}\right)
$$

By the definition of $q(m)$, we have

$$
\lambda\left(\bigcup_{v<k} \mathcal{S}_{m}+r_{q(m), v} \backslash \mathcal{S}_{m}\right) \leq k m 2^{-q(m)} \leq k m 2^{-m}
$$

which establishes the claim.
If for each $n$ there is $v<k$ such that $Z+r_{n, v} \notin \mathcal{P}$, then $Z \in \bigcap_{m} G_{m}$, so $Z$ is not OW -random.

## 4 Open Questions

Due to the novelty of the concept of BD-randomness, a number of natural questions remain; they are not necessarily hard. Our first two questions have been tried by a number of researchers; the third has not been considered in any detail so far.

Question 9. Does density randomness imply BD randomness?
Question 10. Does BD-randomness imply difference randomness?
Question 11. Does lowness for BD-randomness coincide with lowness for MLrandomness?

Miyabe et al. [13, Theorem 2.6] show that lowness for density randomness coincides with lowness for ML-randomness. The containment " $\subseteq$ " is immediate from the result of Downey et al. [8] that Low(W2R, MLR) $=\operatorname{Low}($ MLR $) ~(w h e r e ~ W 2 R ~ i s ~$ the class of weakly 2 -randoms). This proof works the same for BD-randomness. Thus, low for BD-random implies low for ML-random. However, the proof of the converse containment in the case of density randomness cannot be adapted in any obvious way, because we now have to consider a null sequence computable in the oracle.

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# Semantic Security and Key-Privacy with Random Split of St-Gen Codes 

Danilo Gligoroski ${ }^{1(\boxtimes)}$ and Simona Samardjiska ${ }^{2}$<br>${ }^{1}$ Department of Telematics, NTNU, The Norwegian University of Science and Technology, Trondheim, Norway<br>danilog@item.ntnu.no<br>${ }^{2}$ Faculty of Computer Science and Engineering, UKIM, Skopje, Macedonia simona.samardjiska@finki.ukim.mk


#### Abstract

Recently we have defined Staircase-Generator codes (St-Gen codes) and their variant with a random split of the generator matrix of the codes. One unique property of these codes is that they work with arbitrary error sets. In this paper we analyze the semantic security against chosen plaintext attack (IND-CPA) and key-privacy i.e. indistinguishability of public keys under chosen plaintext attack (IK-CPA) of the encryption scheme with random split of St-Gen codes. In a similar manner as it was done by Nojima et al. and later by Yamakawa et al. we show that padding the plaintext with a random bit-string provides INDCPA and IK-CPA in the standard model. The difference with McEliece scheme is that with our scheme the length of the padded random string is significantly shorter.


Keywords: Public key cryptography • Code based cryptosystems • Semantic security • Key-privacy

## 1 Introduction

The idea about semantic security against chosen-plaintext attack (i.e., indistinguishability against chosen-plaintext attack (IND-CPA)) for a public-key cryptosystem (PKC) was initially presented by Goldwasser and Micali in [4]. By replacing the deterministic encryption with probabilistic one, they showed the existence of public key schemes where the ciphertext does not leak any useful information about the plaintext (except its length). Later, in the work of Belare et al. [2] the semantic security against chosen-plaintext attack was systematized in a broader security perspective in relation with other security notions in publickey encryption schemes. Then, in 2001 we got a definition for yet another security notion: key-privacy or anonymity in public-key schemes. It was introduced by Bellare et al. in [1]. In a nutshell key-privacy asks that an adversary receiving a ciphertext is not able to determine which specific public-key, out of a set of known public keys was used to produce that ciphertext. Under the assumption that the Decision Diffie-Hellman problem is hard, they have showed that

El Gamal scheme provides anonymity i.e. key-privacy under chosen-plaintext attack (CPA) and that the Cramer-Shoup scheme provides stronger security, i.e., it provides anonymity under chosen-ciphertext attack (CCA). They have also showed that neither the classical RSA scheme nor the RSA-OAEP does not provide key-privacy. All these schemes do not belong to the so-called family of "post-quantum" crypto schemes since they are vulnerable to attacks with quantum computers.

The McEliece public key scheme [6] was published in 1978 and is based on the theory of error-correcting codes and the NP-hardness of the problem of decoding random linear codes. It is considered as a post-quantum scheme. However, the original scheme does not provide neither CPA nor CCA security, it provides even less a key-privacy. A conversion of McEliece scheme that offers CCA security was proposed by Kobara and Imai in [5] in the random oracles model. The weaker security of CPA, but in the so-called standard model where there is no reference to the random oracles, for a modified McEliece scheme was proposed by Nojima et al. in [8] and later based on that work Yamakawa et al. in [12] showed that Nojima's modification provides also a key-privacy.

Recently we proposed an encryption and signature variant of the McEliece scheme based on Staircase-Generator matrix, a list decoding algorithm, and generalized error sets in [3]. Soon after its initial eprint publication, a distinguisher that distinguishes its public key from random matrices was proposed [11], and recently a very similar distinguishing strategy and an Information Set Decoding (ISD) attack was presented as a full and practical key recovery attack by Moody and Perlner [7]. While the public key schemes based on Staircase-Generator matrices and a list decoding strategy have succumbed to the distinguishing attacks, there are some useful applications for the technique of matrix-embedding in steganography where the Staircase-Generator matrices are not public, but private [9]. In such cases, the matrix-embedding technique with Staircase-Generator matrices almost achieves the information theoretical bound with codelengths that are the smallest known in the literature.

In order to thwart distinguishing and ISD attacks of [7,11] to the encryption scheme defined in [3] we proposed to split and replace the public generator matrix into $s$ randomly generated matrices [10]. With the splitting we made distinguishing attacks improbable to mount, i.e., the probability of the attacker obtaining conditions under which a distinguisher or an ISD attack can be mounted, is close to zero.

## 2 Definition of Staircase-Generator Codes and Random Split of Their Generator Matrix

Throughout the paper, we will denote by $\mathcal{C} \subseteq \mathbb{F}_{2}^{n}$ a binary $(n, k)$ code of length $n$ and dimension $k$. We will denote the generator matrix of the code by $G$, and $w t(\mathbf{x})$ will denote the Hamming weight of the word $\mathbf{x}$. We recall some of the basic definitions and properties for St-Gen codes from [3,9] and the types of errors used.

Definition 1. Let $k_{i}, n_{i} \in \mathbb{N}$, and let $k=k_{1}+k_{2}+\cdots+k_{w}$ and $n=k+n_{1}+n_{2}+$ $\cdots+n_{w}$. Further, let $B_{i}$ be a random binary matrix of dimension $\sum_{j=1}^{i} k_{j} \times n_{i}$. A linear binary $(n, k)$ code $\mathcal{C}$ with the following generator matrix in standard form:

is called Staircase-Generator code (St-Gen code).
Definition 2. Let $\ell$ be a positive integer and let $p_{d} \in \mathbb{F}_{2}\left[x_{1}, x_{2}, \ldots, x_{\ell}\right]$ be a multivariate polynomial of degree $\geqslant 2$. We say that $E_{\ell}$ is an error set if it is the solution set of $p_{d}$, i.e., $E_{\ell}=\left\{\mathbf{e} \in \mathbb{F}_{2}^{\ell} \mid p_{d}(\mathbf{e})=0\right\}$. We will refer to $p_{d}$ as the defining polynomial.

We define the density of the error set $E_{\ell}$ to be $D\left(E_{\ell}\right)=\left|E_{\ell}\right|^{1 / \ell}$. We will refer to the integer $\ell>0$ as the granulation of $E_{\ell}$.

In [3] it was proven that if two error sets $E_{\ell_{1}} \subseteq \mathbb{F}_{2}^{\ell_{1}}, E_{\ell_{2}} \subseteq \mathbb{F}_{2}^{\ell_{2}}$, have the same density $\rho$, then $D\left(E_{\ell_{1}} \times E_{\ell_{2}}\right)=\rho$.

Example 1. 1. Let $E_{2}=\left\{\mathbf{x} \in \mathbb{F}_{2}^{2} \mid w t(\mathbf{x})<2\right\}=\{(0,0),(0,1),(1,0)\}$. Then the error set can be described using the defining polynomial $p_{d}=x_{1} x_{2}$, and for the density of the error set we have $D\left(E_{2}\right)=\left|E_{2}\right|^{1 / 2}=3^{1 / 2}$.
2. Let $E_{4}=\left\{\mathbf{x} \in \mathbb{F}_{2}^{4} \mid 2 \leq w t(\mathbf{x}) \leq 3\right\}$. Then, the defining polynomial for $E_{4}$ is $p_{d}=1+x_{1} x_{2}+x_{1} x_{3}+x_{1} x_{4}+x_{2} x_{3}+x_{2} x_{4}+x_{3} x_{4}$ and the density is $D\left(E_{4}\right)=D\left(E_{4}^{m}\right)=\left(\sum_{i=2}^{3}\binom{4}{i}\right)^{1 / 4}=10^{1 / 4}$ for any positive integer $m$.

The decoding of St-Gen codes relies on the technique of list decoding. In list decoding, the decoder is allowed to output a list of possible messages one of which is correct. List decoding can handle a greater number of errors than that allowed by unique decoding. In order for the decoding to be efficient, the size of the resulting list has to be polynomial in the code length. The following Proposition from [3] determines the parameters of a St-Gen code that provide an efficient decoding.

Proposition 1 [3]. Let $\mathcal{C}$ be any binary $(n, k)$ code and $E \subset \mathbb{F}_{2}^{n}$ be an error set of density $\rho$. Let $\mathbf{w}$ be any word of length $n, W_{E}=\{\mathbf{w}+\mathbf{e} \mid \mathbf{e} \in E\}$ and let $\mathcal{C}_{W_{E}}$ denote the set of codewords in $W_{E}$. Suppose there exists a codeword $\mathbf{c} \in W_{E}$. Then the expected number of codewords in $W_{E} \backslash\{\mathbf{c}\}$ is approximately $\rho^{n} 2^{k-n}$ for large enough $n$ and $k$.

Let $E_{\ell}$ be an error set with density $\rho$ where $\ell$ divides $n$ and $m=n / \ell$. We recall Algorithm 1 from [3], that is an efficient algorithm for decoding a code $\mathcal{C}$, that corrects errors from the set $E_{\ell}^{m}$.

```
Algorithm 1. Decoding
Input: Vector \(\mathbf{y} \in \mathbb{F}_{2}^{n}\), and generator matrix \(G\) of the form (1).
Output: A list \(L_{w} \subset \mathbb{F}_{2}^{k}\) of valid decodings of \(\mathbf{y}\).
Procedure:
Let \(K_{i}=k_{1}+\cdots+k_{i}\). Represent \(\mathbf{x} \in \mathbb{F}_{2}^{k}\) as \(\mathbf{x}=\mathbf{x}_{1}\left\|\mathbf{x}_{2}\right\| \cdots \| \mathbf{x}_{w}\) where each \(\mathbf{x}_{i}\) has length
\(k_{i}\). Similarly, represent \(\mathbf{y} \in \mathbb{F}_{2}^{n}\), as \(\mathbf{y}=\mathbf{y}_{0}\left\|\mathbf{y}_{1}\right\| \mathbf{y}_{2}\|\cdots\| \mathbf{y}_{w}\), where each \(\mathbf{y}_{i}\) has length \(n_{i}\) and
\(\left|\mathbf{y}_{0}\right|=k\). We further identify \(\mathbf{y}_{0}\) with \(\mathbf{y}_{0}=\mathbf{y}_{0}[1]\left\|\mathbf{y}_{0}[2]\right\| \cdots \| \mathbf{y}_{0}[w]\), where each \(\mathbf{y}_{0}[i]\) is of length
\(k_{i}\).
During decoding, we will maintain lists \(L_{1}, L_{2}, \ldots, L_{w}\) of possible decoding candidates of length \(K_{i}\).
```

Step 0: Set a temporary list $T_{0}=L_{0}$ to contain all possible decodings of the first $k_{1}$ coordinates of y :

$$
T_{0} \leftarrow\left\{\mathbf{x}^{\prime}=\mathbf{y}_{0}[1]+\mathbf{e} \mid \mathbf{e} \in E^{k_{1} / \ell}\right\}
$$

Step $1 \leq i \leq w$ : Perform list-decoding to recover a list of valid decodings:
For each candidate $\mathbf{x}^{\prime} \in T_{i-1} \subset \mathbb{F}_{2}^{K_{i}}$, add to $L_{i}$ all the candidates for which $\mathbf{x}^{\prime} B_{i}+\mathbf{y}_{i} \in E^{n_{i} / \ell}$ :

$$
\begin{equation*}
L_{i} \leftarrow\left\{\mathbf{x}^{\prime} \in T_{i-1} \mid \mathbf{x}^{\prime} B_{i}+\mathbf{y}_{i} \in E^{n_{i} / \ell}\right\} \tag{2}
\end{equation*}
$$

If $i<w$ then create the temporary list $T_{i}$ of candidates of length $K_{i+1}$ from $L_{i}$ :

$$
\begin{equation*}
T_{i} \leftarrow\left\{\mathbf{x}^{\prime} \|\left(\mathbf{y}_{0}[i+1]+\mathbf{e}\right) \mid \mathbf{x}^{\prime} \in L_{i}, \mathbf{e} \in E^{k_{i+1} / \ell}\right\} . \tag{3}
\end{equation*}
$$

Return: $L_{w}$.

### 2.1 Random Split of the Staircase-Generator Matrix

A key parameter of the public-key encryption scheme where we split the staircase-generator matrix is the number of splits $s$, the number of summands the generator matrix of the code is split in. This parameter further determines the nature of the error used during encryption. We have the following:

Definition 3. Let $E_{\ell} \subset \mathbb{F}_{2}^{\ell}$ be an error set of granulation $\ell$ and let $s$ denote the number of splits. The s-tuple ErrorSplit $=\left(e_{1}, \ldots, e_{s}\right)$, where $e_{i} \in \mathbb{F}_{2}^{\ell}, i \in$ $\{1, \ldots, s\}$ is called A Valid Error Split for $E_{\ell}$ if the sum of its elements permuted with any permutation $\sigma_{i} \in \mathcal{S}_{\ell}$ is an element of $E_{\ell}$ i.e. it holds that $e=$ $\sum_{i=1}^{s} \sigma_{i}\left(e_{i}\right) \in E_{\ell}$. The set of all valid error splits is denoted as ValidErrorSplits and its size with $V$, i.e., $V=|V a l i d E r r o r S p l i t s|$.

Example 2. Let $\ell=4, E_{\ell}=\{(0,0,0,0),(0,0,1,1),(0,1,0,1),(0,1,1,0)$, $(1,0,1,1),(1,1,0,0),(1,1,0,1),(1,1,1,0),(1,1,1,1)\}$ and $s=4$. The 4 -tuple ErrorSplit $=((1,0,0,0),(1,1,1,1),(1,1,1,1),(1,1,0,1))$ is a valid error split for $E_{\ell}$ because the sum of all its elements permuted by any of all possible $4!=24$ permutations always gives an element in $E_{\ell}$.

A formal description of the scheme is given through the next four algorithms for key generation, error set generation, encryption and decryption. Note that Algorithm 3 is run only once at the time of the initialization of the system with parameters $\ell, E_{\ell}, s$. Even more, in practice, this set can be pre-calculated and publicly available.

```
Algorithm 2. Key Generation
Parameters: Let \(\ell \mid n, m=n / \ell\) and \(E \subset \mathbb{F}_{2}^{\ell}\) be an error set of granulation \(\ell\) and density \(\rho\). Let \(s\) be
the number of splits.
```


## Key generation:

The following matrices make up the private key:

- A generator matrix $G$ of a binary ( $n, k$ ) code of the form (1).
- An invertible matrix $S \in \mathbb{F}_{2}^{k \times k}$.
- An array of permutation matrices $P_{1}, P_{2}, \ldots, P_{s}$ created as follows:

1. Select a permutation $\pi$ on $\{1,2, \ldots, m\}$, and let $P \in \mathbb{F}_{2}^{n \times n}$ be the permutation matrix induced by $\pi$, so that for any $\mathbf{y}=\mathbf{y}_{1}\left\|\mathbf{y}_{2}\right\| \ldots \| \mathbf{y}_{m} \in\left(\mathbb{F}_{2}^{\ell}\right)^{m}$ :

$$
\begin{equation*}
\mathbf{y} P=\mathbf{y}_{\pi(1)}\left\|\mathbf{y}_{\pi(2)}\right\| \ldots \| \mathbf{y}_{\pi(m)}, \tag{4}
\end{equation*}
$$

i.e., $P$ only permutes the $m$ substrings of $\mathbf{y}$ of length $\ell$.
2. For $i:=1$ to $s$ :

Select randomly $m$ permutations $\sigma_{j}^{i} \in \mathcal{S}_{\ell}, j \in\{1, \ldots, m\}$.
Let $P_{i}$ be defined by

$$
\mathbf{y} P_{i}=\sigma_{1}^{i}\left(\mathbf{y}_{\pi(1)}\right)\left\|\sigma_{2}^{i}\left(\mathbf{y}_{\pi(2)}\right)\right\| \ldots \| \sigma_{m}^{i}\left(\mathbf{y}_{\pi(m)}\right)
$$

where $\sigma_{j}^{i}(\mathbf{x})=\sigma_{j}^{i}\left(x_{1}, x_{2}, \ldots, x_{\ell}\right)$.
The public key is formed as follows:
Generate uniformly at random $s-1$ matrices $G_{1}, \ldots, G_{s-1}$ of size $k \times n$ over $\mathbb{F}_{2}$.
Set $G_{s}=G+G_{1}+\cdots+G_{s-1}$.
For all $i \in\{1,2, \ldots, s\}$, set $G_{\text {pub }}^{i}=S G_{i} P_{i}$.
Public key: $G_{\text {pub }}^{1}, \ldots, G_{\text {pub }}^{s}$.
Private key: $S, G$ and $P_{1}, P_{2}, \ldots, P_{s}$.

```
Algorithm 3. Valid Error Splits ( \(\ell, E_{\ell}, s\) )
Input: Granulation \(\ell\), error set \(E_{\ell}\), number of splits \(s\).
Output: A set ValidErrorSplits of all possible valid error splits.
    Set ValidErrorSplits \(\leftarrow \emptyset\)
    for all \(\left(e_{1}, \ldots, e_{s}\right) \in\left(\mathbb{F}_{2}^{\ell}\right)^{s}\) do
        if \(\sum_{i=1}^{s} \sigma_{i}\left(e_{i}\right) \in E_{\ell}, \forall\left(\sigma_{1}, \ldots, \sigma_{s}\right) \in\left(\mathcal{S}_{\ell}\right)^{s}\) then
            Add \(\left(e_{1}, \ldots, e_{s}\right)\) to ValidErrorSplits.
        end if
    end for
    Return ValidErrorSplits.
```

```
Algorithm 4. Encryption (m, \(G_{\text {pub }}^{1}, \ldots, G_{\text {pub }}^{s}\), ValidErrorSplits)
Input: Message to be encrypted \(\mathbf{m}\) the public key \(G_{\text {pub }}^{1}, \ldots, G_{\text {pub }}^{s}\) and a set ValidErrorSplits of
all possible valid error splits.
Output: A ciphertext \(\mathbf{c}=\left(\mathbf{c}_{1}, \ldots, \mathbf{c}_{s}\right)\).
    1: Set \(\mathbf{c}_{i}=\mathbf{m} G_{\text {pub }}^{i}+\mathbf{e}_{i}, i=1, \ldots, s\), where \(\mathbf{e}_{i}=\left(e_{1, i}, \ldots, e_{\frac{n}{l}, i}\right)\) and where \(\left(e_{j, 1}, \ldots, e_{j, s}\right)\),
    \(j=1, \ldots, \frac{n}{l}\) are randomly drawn from ValidErrorSplits.
    : Return \(\mathbf{c}=\left(\mathbf{c}_{1}, \ldots, \mathbf{c}_{s}\right)\)
```

```
Algorithm 5. Decryption (c, \(S, G, P_{1}, P_{2}, \ldots, P_{s}\) )
Input: Ciphertext c, matrix \(S\), the generator matrix \(G\) and the permutation matrices \(P_{1}, P_{2}, \ldots, P_{s}\).
Output: A decrypted message m.
    1: Set \(\mathbf{c}_{i}^{\prime}=\mathbf{c}_{i} P_{i}^{-1}\)
    2: Set \(\mathbf{c}^{\prime}=\sum_{i=1}^{s} \mathbf{c}_{i}^{\prime}\)
    3: Set \(\mathbf{m}^{\prime}\) as the output of Algorithm 1 (List decoding of \(\mathbf{c}^{\prime}\) with generator matrix \(G\) ).
    4: Set \(\mathbf{m}=\mathbf{m}^{\prime} S^{-1}\)
    5: Return \(\mathbf{m}\)
```

The initial proposal for encryption scheme that uses St-Gen codes [3], is vulnerable to an Information Set Decoding (ISD) attack [7,11].

In a very recent analysis by Moody and Perlner [7] a modification of Stern's algorithm was provided, dedicated to cryptanalysis of the scheme in [3]. We refer the reader to [7] for details, and here we mention that complexity of the attack is in general given by $I S D_{S t}=P r_{S t}^{-1} \cdot C o s t_{S t}$ where $P r_{S t}^{-1}$ is the probability of success, and Cost $_{S t}$ the cost of finding the low weight codeword.

In [10] we gave a detailed security analysis how and why the random split of the generator matrix prevents from ISD attacks. We also give several practical parameter sets. Based on all that we give the following conjecture:

Conjecture 1. The public key $G_{\text {pub }}^{1}, \ldots, G_{\text {pub }}^{s}$ produced by Algorithm 1 is indistinguishable from a set of $s$ random $[n, k]$ codes and inverting the encryption with $G_{\text {pub }}^{1}, \ldots, G_{\text {pub }}^{s}$ without the knowledge of the private key $S, G$ and $P_{1}, P_{2}, \ldots, P_{s}$ is infeasible in polynomial time.

## 3 Achieving IND-CPA and IK-CPA by Padding the Plaintext with a Random Bit-String

In this section we apply the ideas described by Nojima et al. in [8] and by Yamakawa et al. in [12] for the McEliece scheme, to show that our encryption scheme with random split of the generator matrix can achieve semantic security against chosen plaintext attack (IND-CPA) and key-privacy, i.e., indistinguishability of public keys under chosen plaintext attack (IK-CPA).

Firs we give a definition of indistinguishability of encrypted data against the chosen plaintext attack (CPA) as it is given in [2].

Definition 4 (IND-CPA [2]). Let a PKE scheme be the following tuple of polynomial-time algorithms: $P K E=(G e n, E n c$, Dec $)$.

1. On input of security parameter $\kappa$, key generation algorithm Gen $\left(1^{\kappa}\right)$ outputs the set of private-key and public-key, (pk, sk).
2. Given ( $p k, s k$ ), a polynomial-time adversary $\mathcal{A}$ chooses two equal-length plaintexts $m_{0}, m_{1},\left(m_{0} \neq m_{1}\right)$, and sends them to the encryption oracle.
3. Encryption oracle (algorithm) randomly fips coin $b \in\{0,1\}$, to encrypt $\operatorname{Enc}\left(p k, m_{b}\right)=c$.
4. Given target ciphertext $c$, adversary $\mathcal{A}$ outputs $b^{\prime} \in\{0,1\}$, where the advantage of success probability over random guess is defined as follows:

$$
\begin{equation*}
\mathbf{A d v}_{\mathcal{A}}^{i n d-c p a}(\kappa)=\operatorname{Pr}\left[b^{\prime}=0 \mid b=0\right]+\operatorname{Pr}\left[b^{\prime}=1 \mid b=1\right] \tag{5}
\end{equation*}
$$

If $\mathbf{A d v} \mathbf{\mathcal { A }}^{i n d-c p a}(\kappa)$ is negligible, then, we say underlying PKE is IND-CPA secure. Here "negligible" means that for any constant const, there exists $k_{0} \in$ $\mathbb{N}$, s.t. for any $\kappa>k_{0}, \mathbf{A d v}$ is less than $\left(\frac{1}{\kappa}\right)^{\text {const }}$.

As we said, in order to achieve IND-CPA in a standard model, Nojima et al. [8] proposed a random prepadding for messages in McEliece scheme, i.e., instead of encrypting messages $m$ to encrypt messages $[r \mid m$ ] where $r$ is a random prepadding. However, in order to achieve more than $2^{80}$ security or more than $2^{128}$ security, $r$ should have a significant length. More concretely for the McEliece code $(2048,1289,69)$ to achieve security of $2^{85}$, out of 1289 bits the random prepadding $r$ has to have 1161 bits and only 128 bits are left for $m$. For the other code $(4096,2560,128)$, to achieve security of $2^{131}$, out of 2560 bits the random prepadding $r$ has to have 2048 bits and only 512 bits are left for $m$.

In what follows we investigate the IND-CPA security of the approach of encrypting messages in a form $\mathbf{m}=[r \mid m]$ for our scheme with a random split of St-Gen matrix (we abbreviate the name of that system as RRS-St-Gen Randomized Random Split of St-Gen - in the mathematical formulas that refer to that system). By having this form for $\mathbf{m}$ the encryption gets the following form: The ciphertext is the $s$-tuple $\mathbf{c}=\left(\mathbf{c}_{1}, \ldots, \mathbf{c}_{s}\right)$ where

$$
\begin{equation*}
\mathbf{c}_{i}=\mathbf{m} G_{\mathrm{pub}}^{i}+\mathbf{e}_{i}=\left(r G_{\mathrm{pub}_{1}}^{i}+\mathbf{e}_{i}\right)+m G_{\mathrm{pub}_{2}}^{i}, \tag{6}
\end{equation*}
$$

where $G_{\text {pub }}^{i}=\left[\frac{G_{\text {pub }_{1}}^{i}}{G_{\text {pub }_{2}}^{i}}\right], \mathbf{e}_{i}=\left(e_{1, i}, \ldots, e_{\frac{n}{l}, i}\right)$ and where $\left(e_{j, 1}, \ldots, e_{j, s}\right), j=$ $1, \ldots, \frac{n}{l}$ are randomly drawn from ValidErrorSplits.

Theorem 1. Let RRS-St-Gen is a randomized public key encryption scheme as defined in Algorithms 1, 2, 3, 4 and 5 with the following parameters: n, $k, l, s, E_{l}$, ValidErrorSplits, $V=\mid$ ValidErrorSplits $\mid$, that encrypts messages $\mathbf{m}=[r \mid m]$ where $r$ is a random prepadding. If the s-tuple of matrices $\left(G_{p u b}^{1}, \ldots, G_{p u b}^{s}\right)$ is indistinguishable from random, and the inversion of the encryption with $G_{p u b}^{1}, \ldots, G_{\text {pub }}^{s}$ without the knowledge of the private key $S, G$ and $P_{1}, P_{2}, \ldots, P_{s}$ is infeasible in polynomial time, then the advantage of an adversary given by relation (5) is:

$$
\begin{align*}
& \mathbf{A d v}_{\mathcal{A}-R R S-S t-G e n}^{i n d-c p a}(\kappa)=2\left(\left(1-\frac{1}{2^{|r|}}\right)\left(\frac{V}{2^{s \cdot l}}\right)^{\frac{n}{l}}\left(1-\left(\frac{V}{2^{s \cdot l}}\right)^{\frac{n}{l}}\right)+\right. \\
& \left.\quad+\frac{1}{2^{|r|}}\left(1-\left(\frac{V}{2^{s \cdot l}}\right)^{\frac{n}{l}}\right)\right) \tag{7}
\end{align*}
$$

Proof. First note that the assumption about the infeasibility of the inversion of the encryption is a crucial one. If the adversary is capable to invert the encryption, it will simply obtain the whole random prepadding and will guess the value of $b^{\prime}$ with probability 1 . The necessity of the assumption for the indistinguishability from random matrices is due to the attacks that reveal the secret key of the scheme and then is also connected with the inversion of the encryption. We discuss additionally these two assumptions at the end of this proof.

We will compute the probability $\operatorname{Pr}\left[b^{\prime}=0 \mid b=0\right]$ and due to the symmetry, the probability for $\operatorname{Pr}\left[b^{\prime}=1 \mid b=1\right]$ has the same value. First let us
note that every $\mathbf{c}_{i}$ have $\frac{n}{l}$ substrings of length $l$ bits. For a concrete value of $\mathbf{m}=[r \mid m]$ the encryption procedure randomly picks $\frac{n}{l}$ elements from set of $s$ tuples ValidErrorSplits. Note also that the set ValidErrorSplits is a subset of all possible $l$-bit $s$-tuples which number is $2^{s \cdot l}$. In case when $m_{0}$ was encrypted there are two possible and disjunctive events that can lead the adversary to make the right guess $b^{\prime}=0$. Those two events are the following:

- Event ${ }_{1}:$ Event $_{1,1} \cap$ Event $_{1,2} \cap$ Event $_{1,3}$

1. Event $t_{1,1}$ : The adversary made a wrong guess about the prepadded random value $r \operatorname{Pr}\left(\right.$ Event $\left._{1,1}\right)=1-\frac{1}{2^{|r|}}$;
2. Event $_{1,2}$ : The adversary computes $\mathbf{c}_{i}^{\left(m_{0}\right)}=r G_{\text {pub }_{1}}^{i}+m_{0} G_{\text {pub }_{2}}^{i}$ and for all $\frac{n}{l}$ chunks in all $\mathbf{c}_{i}$, there exist a valid error split in ValidErrorSplits as a connection between $\mathbf{c}_{i}^{\left(m_{0}\right)}$ and $\mathbf{c}_{i} \cdot \operatorname{Pr}\left(\right.$ Event $\left._{1,2}\right)=\left(\frac{V}{2^{s \cdot i}}\right)^{\frac{n}{l}}$;
3. Event $t_{1,3}$ : The adversary computes $\mathbf{c}_{i}^{\left(m_{1}\right)}=r G_{\mathrm{pub}_{1}}^{i}+m_{1} G_{\mathrm{pub}_{2}}^{i}$, and there is at least one chunk for which there is no valid error split in ValidErrorSplits. $\operatorname{Pr}\left(\right.$ Event $\left._{1,3}\right)=\left(1-\left(\frac{V}{2^{s \cdot 1}}\right)^{\frac{n}{l}}\right)$.

- Event ${ }_{2}$ : Event ${ }_{2,1} \cap$ Event $_{2,2}$

1. Event ${ }_{2,1}$ : The adversary made a correct guess about the prepadded random value $r . \operatorname{Pr}\left(\right.$ Event $\left._{2,1}\right)=\frac{1}{2^{1 r r}}$;
 there is at least one chunk for which there is no valid error split in ValidErrorSplits. $\operatorname{Pr}\left(\right.$ Event $\left._{2,2}\right)=\left(1-\left(\frac{V}{2^{s \cdot l}}\right)^{\frac{n}{l}}\right)$.

Composing probabilities for all events we get:

$$
\begin{aligned}
\operatorname{Pr}\left[b^{\prime}=0 \mid b=0\right]= & \operatorname{Pr}\left(\text { Event }_{1}\right)+\operatorname{Pr}\left(\text { Event }_{2}\right) \\
= & \operatorname{Pr}\left(\text { Event }_{1,1}\right) \operatorname{Pr}\left(\text { Event }_{1,1}\right) \operatorname{Pr}\left(\text { Event }_{1,1}\right)+ \\
& +\operatorname{Pr}\left(\text { Event }_{2,1}\right) \operatorname{Pr}\left(\text { Event }_{2,2}\right) \\
= & \left(1-\frac{1}{2^{|r|}}\right)\left(\frac{V}{2^{s \cdot l}}\right)^{\frac{n}{l}}\left(1-\left(\frac{V}{2^{s \cdot l}}\right)^{\frac{n}{l}}\right)+\frac{1}{2^{|r|}}\left(1-\left(\frac{V}{2^{s \cdot l}}\right)^{\frac{n}{l}}\right)
\end{aligned}
$$

Since the case $\operatorname{Pr}\left[b^{\prime}=1 \mid b=1\right]$ is symmetrical, the total $\mathbf{A d v}_{\mathcal{A}-R R S-S t-G e n}^{i n d-c p a}(\kappa)$ is:

$$
\begin{aligned}
& \mathbf{A d v}_{\mathcal{A}-R R S-S t-G e n}^{i n d-c p a}(\kappa)=2\left(\left(1-\frac{1}{2^{|r|}}\right)\left(\frac{V}{2^{s \cdot l}}\right)^{\frac{n}{l}}\left(1-\left(\frac{V}{2^{s \cdot l}}\right)^{\frac{n}{l}}\right)+\right. \\
& \left.\quad+\frac{1}{2^{|r|}}\left(1-\left(\frac{V}{2^{s \cdot l}}\right)^{\frac{n}{l}}\right)\right)
\end{aligned}
$$

A direct conclusion from the security analysis in [10] is that if the length of the message that is encrypted with the scheme is bigger than 256 bits, than the claims in the Conjecture 1 about the indistinguishability of the public key from random matrices and the infeasibility of the inversion of the encryption are
plausible. So, for any concrete instantiation of the scheme, the security levels that are achieved with the adversary advantage defined in the relation (3) have to give lengths of the random prepadding to be more than 256 bits.

By the final part of the proof of Theorem 1 we obtain that our scheme achieves the IND-CPA security level of $2^{80}$ for $|r|=264$ and the security level of $2^{128}$ for $|r|=424$. So these values being higher than 256 are in accordance with the security analysis in [10] and the Conjecture 1. Moreover, they are still significantly lower than the lengths of the prepadded random part in the modified McEliece scheme.

For the key-privacy issue we use the same approach as Yamakawa et al. have in [12].

Definition 5 (IK-CPA [1]). Let a PKE scheme be the following tuple of polynomial-time algorithms: PKE = (Gen, Enc, Dec). The security of keyprivacy is defined as follows.

1. On input of security parameter $\kappa$, key generation algorithm $\operatorname{Gen}\left(1^{\kappa}\right)$ outputs two independent sets of key pairs $\left(p k_{0}, s k_{0}\right),\left(p k_{1}, s k_{1}\right)$.
2. Given $\left(p k_{0}\right),\left(p k_{1}\right)$, a polynomial-time adversary $\mathcal{A}$ chooses a plaintext $m$ and sends them to the encryption oracle.
3. Encryption oracle randomly flips coin $b \in\{0,1\}$, to output $E n c_{p k_{b}}(m)=c$.
4. Given target ciphertext $c$, adversary $\mathcal{A}$ outputs $b^{\prime} \in\{0,1\}$, where the advantage of success probability over random guess is defined as follows:

$$
\begin{equation*}
\mathbf{A d v}_{\mathcal{A}}^{i k-c p a}(\kappa)=\operatorname{Pr}\left[b^{\prime}=0 \mid b=0\right]+\operatorname{Pr}\left[b^{\prime}=1 \mid b=1\right] \tag{8}
\end{equation*}
$$

If $\mathbf{A d v}{ }_{\mathcal{A}}^{i k-c p a}(\kappa)$ is negligible, then, we say underlying PKE is IK-CPA secure.
While the modeling of IK-CPA is not the same as IND-CPA, the value about $\mathbf{A d v}_{\mathcal{A}}^{i k-c p a}(\kappa)$ is the same as for IND-CPA case. Thus we have the following theorem that we give without a proof:

Theorem 2. Let RRS-St-Gen is a randomized public key encryption scheme as defined in Algorithms 1, 2, 3, 4 and 5 with the following parameters: $n$, $k, l, s, E_{l}$, ValidErrorSplits and $V=\mid$ ValidErrorSplits $\mid$, encrypting messages $\mathbf{m}=[r \mid m]$ where $r$ is a random prepadding. If the s-tuple of matrices $\left(G_{p u b}^{1}, \ldots, G_{p u b}^{s}\right)$ is indistinguishable from random, and the inversion of the encryption with $G_{p u b}^{1}, \ldots, G_{\text {pub }}^{s}$ without the knowledge of the private key $S, G$ and $P_{1}, P_{2}, \ldots, P_{s}$ is infeasible in polynomial time, then the advantage of an adversary given by relation (8) is:

$$
\begin{equation*}
\operatorname{Adv}_{\mathcal{A}-R R S-S t-G e n}^{i k-c p a}(\kappa)=\mathbf{A d} \mathbf{v}_{\mathcal{A}-R R S-S t-G e n}^{i n d-c p a}(\kappa) \tag{9}
\end{equation*}
$$

## 4 Conclusions

We have presented a public key encryption scheme based on St-Gen codes and its variant where we split and replace the public generator matrix into $s$ randomly
generated matrices. The split strategy is used to thwarts the ISD attacks on the encryption scheme. Then, we showed that randomized version of the encryption scheme offers semantic security against chosen plaintext attack (IND-CPA) and offers key-privacy, i.e., offers indistinguishability of public keys under chosen plaintext attack (IK-CPA) in the standard model. The difference with McEliece scheme is that with our scheme the length of the prepadded random string is significantly shorter. It remains as a next goal to investigate the modification of the scheme for achieving the stronger securities of chosen-ciphertext attacks (CCA and CCA2) both with and without the random oracle model.

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# The Typical Constructible Object 

Mathieu Hoyrup ${ }^{(\boxtimes)}$<br>LORIA, Inria Nancy Grand Est, Villers-lès-Nancy, France<br>mathieu.hoyrup@inria.fr

## 1 Introduction

Baire Category is an important concept in mathematical analysis. It gives a notion of large set, hence a way of identifying the properties of typical objects. One of the most important applications of Baire Category is to provide a way of proving the existence of objects with specified properties without having to give an explicit construction, showing at the same time that these properties are prevalent. For instance it has been extensively used in mathematical analysis to better understand and separate classes of real functions such as analytic and smooth functions (see [9] for a wide range of applications of the Baire Category Theorem in analysis).

This note is about Baire Category in constructive or computable analysis. This subject has been studied in many different directions, for instance in reverse mathematics [3], constructive mathematics [2] or computable analysis [1]. In these fields, one is often interested in studying the properties of "constructible" objects, to separate classes of constructible objects, and to identify the effectiveness of mathematical proofs.

Here we are interested in a particular question: how to apply Baire Category inside classes of constructible objects? Such classes are very small in the sense of Baire Category, as they are countable, so strictly speaking Baire Category cannot be applied inside them. However one can adapt Baire Category and save part of it in these small worlds, and this note illustrates this. We will focus on 4 particular classes of constructible objects, depicted in Fig. 1:

- The class $C$ of computable subsets of $\mathbb{N}$.
- The class CE of c.e. subsets of $\mathbb{N}$.
- The class LCE of left-c.e. real numbers in $[0,1]$.
- The class $\Delta_{2}^{0}$ of subsets of $\mathbb{N}$ that are computable relative to the halting problem.

The goal is, for each class of constructible objects, to find an analog of Baire Category to define a notion of large set, of typical or generic object. Why is it interesting? It helps understanding which parts of classical mathematics are still available in a constructive setting. From a more practical perspective, it gives a way to prove the existence of constructible objects with specified properties, avoiding as in the classical setting explicit constructions by using a simpler argument.


Fig. 1. Four classes of constructible subsets of $\mathbb{N}$

For each class $\mathscr{C}$ of objects, we want to investigate the following question:

## What does the typical object of $\mathscr{C}$ look like?

The way to do this is to define a notion of small subclass of $\mathscr{C}$. Such a notion should satisfy the following conditions:

Axiom 1. Every singleton $\{A\}$ with $A \in \mathscr{C}$ is small in $\mathscr{C}$,
Axiom 2. The class $\mathscr{C}$ is not small in itself,
Axiom 3. Effective countable unions of small sets are small, for some notion of effectiveness,
Axiom 4. In the hierarchy depicted in Fig. 1, a subclass is small in the superclass. In other words, a typical object of the superclass does not belong to the subclass.

A measure-theoretical approach to this problem has been developed by many authors using resource-bounded measure and dimension theory, in particular resource-bounded martingales (see [11,12] for instance). These methods were mainly applied to complexity classes, but not to classes of enumerable objects such as c.e. sets or left-c.e. reals: whether such a development can be done is an interesting problem.

Here we adopt the topological approach of Baire Category. For each class it is done by defining the analog of a nowhere dense subclass by means of effectivizatimon. The meager subclasses are then generated by the nowhere dense classes: they are the subsets of effective countable unions of nowhere dense classes. More precisely in an effective version,

1. Nowhere dense sets are effective, i.e., describable in some way by programs,
2. Countable unions $\bigcup_{i} A_{i}$ are effective, i.e., there is a single program that given $i$ as input, describes $A_{i}$ in the way specified in 1 .

Hence to obtain a notion of meager subclass one simply has to define what is an effective nowhere dense set. Axiom 3 will be automatically satisfied.

### 1.1 Notations and Background

The set of finite binary strings is denoted by $2^{*}$. The Cantor space $2^{\mathbb{N}}$ is the space of infinite binary sequences, also identified to subsets of $\mathbb{N}$ or reals numbers in $[0,1]$. If $A$ is an infinite binary sequence then $A \upharpoonright n$ is the prefix of $A$ of length $n$. For each finite binary string $u \in 2^{*}$, the cylinder $[u] \subseteq 2^{\mathbb{N}}$ is the class of infinite binary extensions of $u$. The Cantor space is endowed with the topology generated by the cylinders: the open classes $\mathcal{U} \subseteq 2^{\mathbb{N}}$ are the unions of cylinders. A subclass of $2^{\mathbb{N}}$ is nowhere dense if it is disjoint from a dense open class. A subclass is meager if it is a union of nowhere dense classes.

## 2 Computable Objects

Let $C$ be the class of computable subsets of $\mathbb{N}$.
An ordinary nowhere dense set is a set that is disjoint from a dense open set. A natural effective version is then to require the dense open set to be effective, i.e., expressible as a union $\bigcup_{u \in A}[u]$ where $A \subseteq 2^{*}$ is a computably enumerable set.

We then say that a class is meager $\boldsymbol{i n} \mathrm{C}$ if it is contained in an effective union of complements of dense open sets. A class is co-meager in C if it contains an effective intersection of dense open sets.

This notion of nowhere dense set makes the Baire Category theorem computable:

Theorem 2.1 (Baire Category Theorem in C). Every class that is comeager in C is dense in C , i.e., contains computable elements in every cylinder.

In other words, the class $C$ is not meager in itself and Axiom 2 is satisfied.
Observe that Axiom 1 is also satisfied, i.e., a singleton is meager in C : if $A \subseteq \mathbb{N}$ is computable then $2^{\mathbb{N}} \backslash\{A\}$ is a dense effective open class. Every subclass that can be effectively listed is also meager in C , for instance the class P of polynomialtime decidable problems.

Example. Consider a map $T$ from $[0,1]$ to $[0,1]$ that is computable. Think of $T$ as a dynamical system: if $x \in[0,1]$ is the state of the system at time $t$ then $T(x)$ is the state at time $t+1$. An initial state $x_{0}$ induces a trajectory defined by $x_{t+1}=$ $T\left(x_{t}\right)$. What do typical trajectories look like?

Now imagine that one simulates $T$ on a computer, computing the trajectory starting from some state $x_{0}$. Of course the computer can only manipulate computable real numbers, so only the computable part of the dynamical system can be observed on the computer, namely the map $T:[0,1] \cap \mathrm{C} \rightarrow[0,1] \cap \mathrm{C}$ (here $C$ is the set of computable real numbers). What do typical trajectories of the restricted system look like? Are they representative of the original system over $[0,1]$ ?

If the system has a dense trajectory then one can show that all the typical trajectories are dense (i.e., the set of initial states inducing a dense trajectory is
co-meager), and the computable Baire Category theorem directly implies that the typical trajectories of the restricted system are also dense, in particular there exists a computable dense trajectory.

## 3 Computable Relative to the Halting Problem

Let $\Delta_{2}^{0}$ be the class of subsets of $\mathbb{N}$ that are computable relative to the halting problem.

One can relativize the Computable Baire Category theorem to any oracle, in particular to the halting problem, in a straightforward way. It gives for free a notion of nowhere dense class in $\Delta_{2}^{0}$ : it is a class that is disjoint from a dense open class that is effective relative to the halting problem. Effective unions can be equivalently taken relative to the halting problem or not (the two notions are equivalent by the relativized $s-m-n$ theorem).

Again, Axioms 1 and 2 are satisfied, i.e., the class $\Delta_{2}^{0}$ is not meager in itself and every singleton is meager in $\Delta_{2}^{0}$. Moreover Axiom 4 is also satisfied as the subclass LCE is meager in $\Delta_{2}^{0}$ : the halting problem can effectively list LCE, which is then an effective union of singletons, hence is meager in $\Delta_{2}^{0}$.

We now present a particular class that is co-meager in $\Delta_{2}^{0}$ that has received a lot of attention.

### 3.1 1-Generic Sets

In Sect. 2 we mentioned that an ordinary nowhere dense class is a class that is disjoint from a dense open set, which naturally induces the first effective version. A nowhere dense class can equivalently be defined as a subset of the boundary of an arbitrary (i.e., not necessarily dense) open set, which gives another possible effective version: a subset of the boundary of an effective open set.

It gives a strictly weaker notion of nowhere dense class: in general the boundary of an effective open class is not disjoint from a dense effective open class. However, it is always disjoint from a dense open class that is effective relative to the halting problem. As a result, it is meager in $\Delta_{2}^{0}$. As the effective open classes can be effectively enumerated, taking the union of their boundaries gives a class that is meager in $\Delta_{2}^{0}$. Its complement is known as the class of 1-generic subsets of $\mathbb{N}$, and is co-meager in $\Delta_{2}^{0}$. In particular it is non-empty and even dense. It was introduced by Jockush [8] in order to simplify constructions in recursion theory, in the same way as Baire Category simplifies proofs of existence results in mathematical analysis.

While the class of 1 -generic sets is just one particular class that is comeager $\Delta_{2}^{0}$, it happens that it captures many interesting co-meager classes in $\Delta_{2}^{0}$, in the sense that it contains them. In other words, many typical properties are already satisfied by the 1 -generic sets. Indeed being 1-generic is a kind of universal property as it is about every effective open set.

For instance, we mentioned that the subclass LCE $\subsetneq \Delta_{2}^{0}$ is meager in $\Delta_{2}^{0}$, i.e., typical $\Delta_{2}^{0}$ sets are not left-c.e. Actually, no 1-generic set is left-c.e. Indeed,
if $x \in[0,1]$ is left-c.e. then the interval $[0, x)$ is an effective open set that is dense along $x$, i.e., contains $x$ in its boundary.

In the same way as Baire Cateory provides a simple way of proving existence results without giving explicit constructions, 1-genericity can be easily used to prove existence of $\Delta_{2}^{0}$-sets with prescribed properties, and is an alternative to explicit constructions.

A famous example is Kleene-Post's theorem, on the way to the solution to Post's problem:

Theorem 3.1 (Kleene and Post [10]). There exist two Turing-incomparable $\Delta_{2}^{0}$-sets, i.e. two $\Delta_{2}^{0}$-sets that are not computable relative to each other.

Actually for a typical $\Delta_{2}^{0}$-set $A$, its two halves $A_{0}=\{n \in \mathbb{N}: 2 n \in A\}$ and $A_{1}=\{n \in \mathbb{N}: 2 n+1 \in A\}$ are Turing-incomparable $\Delta_{2}^{0}$-sets. Moreover,

Theorem 3.2 (Jockush [8]). Every 1-generic set has Turing-incomparable halves.

Proof. This can be proved very easily: given a Turing machine $M$, let $\mathcal{U}=\{A$ : $\exists n, M^{A_{0}}(n)=0$ but $\left.n \in A_{1}\right\} . \mathcal{U}$ is an effective open class and if $A_{1}$ is Turing reducible to $A_{0}$ via $M$ then $A$ belongs to the boundary of $U$ : adding an arbitrary large element to $A_{1}$ makes $A$ fall in $\mathcal{U}$ (possible when $A_{1}$ is co-infinite). Hence if $A$ is 1-generic, $A$ does not belong to the boundary of $\mathcal{U}$ and as $A_{1}$ is easily co-infinite, $M$ does not compute $A_{1}$ relative to $A_{0}$.

Hence instead of constructing a $\Delta_{2}^{0}$-set with the specific property, one simply has to check that the property is co-meager in $\Delta_{2}^{0}$, or even that it is captured by 1-genericity. The relativized computable Baire Category theorem gives the existence result for free.

## 4 Enumerable Objects

As we saw, the case of computable sets is a straightforward effectivization of the ordinary, non-effective setting, from which the case of $\Delta_{2}^{0}$-sets is a straightforward relativization.

We now turn our attention to intermediate classes of objects. What is a small class inside the class CE of c.e. subsets of $\mathbb{N}$ ? What is a typical c.e. set? The same questions can be asked for left-c.e. sets, $\Pi_{1}^{0}$-classes, etc.

### 4.1 C.e. Sets

First observe that as $C \subseteq C E$, we can declare every class that is meager in $C$ to be meager in CE. Axiom 1 is satisfied, i.e., every singleton is meager in CE: (i) if $A \subseteq \mathbb{N}$ is finite then it is computable so as before $\{A\}$ is meager as its complement is a dense effective open set, (ii) if $A$ is infinite then the class $\mathcal{U}=$ $\{B \subseteq \mathbb{N}: A \nsubseteq B\}$ is a dense effective open set that does not contain $A$. However

Axiom 4 is not satisfied as $C$ is not meager in $C E$, so we have to add new meager classes in CE.

The idea is to weaken the definition of 1-genericity. Let us recall that a set $A \subseteq \mathbb{N}$ is 1-generic if it does not belong to the boundary of any effective open class $\mathcal{U} \subseteq 2^{\mathbb{N}} ; A$ belongs to the boundary of $\mathcal{U}$, or $\mathcal{U}$ is dense along $A$, means that $\mathcal{U}$ contains sets arbitrarily close to $A$, i.e., in every cylinder $[A \upharpoonright n]$.

Definition 4.1. We say that $A$ belongs to the down-closure of $\mathcal{U}$, or that $\mathcal{U}$ is dense above $\boldsymbol{A}$, if $\mathcal{U}$ contains sets $B \supseteq A$ arbitrarily close to $A$, i.e., in every cylinder $[A \upharpoonright n]$. We say that $A$ belongs to the down-boundary of $\mathcal{U}$ if $A$ belongs to the down-closure of $\mathcal{U}$ but not to $\mathcal{U}$.

Observe that the down-boundary of $\mathcal{U}$ is contained in the boundary of $\mathcal{U}$, so if $\mathcal{U}$ is an effective open set then its down-boundary is meager in $\Delta_{2}^{0}$. We declare that its down-boundary is already meager in CE. It gives a notion of genericity in the class CE.

Definition 4.2. $A$ set $A$ is generic from above if it belongs to every effective open class that is dense above $A$. In other words, $A$ is generic from above if it does not belong to the down-boundary of any effective open class.

This notion is equivalent to the notion of $p$-genericity introduced by Ingrassia [7]. We then prove a Baire Category theorem in CE.

Theorem 4.3 (Baire Category in CE [5, 7]). Let $\mathcal{U}_{n} \subseteq 2^{\mathbb{N}}$ be dense uniformly effective open classes. The class of c.e. sets in $\bigcap_{n} \mathcal{U}_{n}$ that are generic from above is dense in $2^{\mathbb{N}}$.

As a result, the class CE is not meager in itself and Axiom 2 is satisfied. Moreover Axiom 4 is satisfied, i.e., the subclass C is meager in CE: (i) the class of co-finite sets is already meager in C as it can be effectively listed, (ii) if $A \subseteq \mathbb{N}$ is co-infinite and computable then it is not generic from above, as the effective open set $2^{\mathbb{N}} \backslash\{A\}$ is dense above $A$.

Genericity from above is a weakening of 1-genericity, that is sometimes sufficient for our purpose. For instance the simple argument showing that the two halves of a 1-generic set are Turing-incomparable (Theorem 3.2) immediately applies to co-infinite sets that are generic from above. As a result, Theorem 4.3 implies the solution to Post problem invented by Friedberg and Muchnik.

Theorem 4.4 (Friedberg-Muchnik [4,13]). There exist two Turingincomparable c.e. sets.

Additionally, we can say that having Turing-incomparable halves is a typical property of c.e. sets, and every c.e. set that is generic from above has this property.

Obverse that it does not give an alternative proof of Friedberg-Muchnik's theorem, as Theorem 4.3 is showed using the priority method with finite injury, the method invented by Friedberg and Muchnik to prove their result. However,
many constructions using a simple form of the priority method are captured by the Baire Category theorem in CE. For instance, every co-infinite set $A$ that is generic from above is not autoreducible: there is no Turing functional that for each $n \in \mathbb{N}$, decides $n \in A$ given $A \backslash\{n\}$ as oracle. Not all finite injury arguments are captured by Theorem 4.3: for instance Ingrassia [7] proved the existence of a $p$-generic (i.e., co-infinite and generic from above) c.e. set that is Turing equivalent to the halting set, hence is not a solution to Post's problem.

### 4.2 Left-C.e. Reals

We can adapt the previous definitions in a straightforward way, replacing the inclusion ordering over $2^{\mathbb{N}}$ by the lexicographic ordering. Identifying subsets of $\mathbb{N}$ with real numbers in $[0,1]$, we get the natural ordering of real numbers.

Definition 4.5. We say that $x \in[0,1]$ belongs to the left-closure of $\mathcal{U} \subseteq[0,1]$, or that $\mathcal{U}$ is dense on the right of $x$, if $\mathcal{U}$ contains reals $y \geq x$ arbitrarily close to $x$, i.e., in every interval $[x, x+\epsilon$ ). We say that $x$ belongs to the left-boundary of $\mathcal{U}$ if $x$ belongs to the left-closure of $\mathcal{U}$ but not to $\mathcal{U}$.

For instance, if $0<a<1$ then the left-boundary of $(a, 1]$ is $\{a\}$ while the leftboundary of $[0, a)$ is empty. In particular, while every left-c.e. real $a>0$ belongs to the boundary of the effective open set $[0, a)$, preventing it from being 1-generic, it does not belong to its left-boundary.

We then declare the left-boundary of an effective open class to be meager in LCE, and we get a notion of genericity.

Definition 4.6. $A$ real $x$ is generic from the right if it belongs to every effective open class that is dense on the right of $x$. In other words, $x$ is generic from the right if $x$ does not belong to the left-boundary of any effective open class.

Again we have a Baire Category theorem in LCE.
Theorem 4.7 (Baire Category in LCE). The class of left-c.e. reals that are generic from the right is non-empty and dense.

Observe that we do not need to intersect with an effective family of dense open sets as in Theorem 4.3, as every dense set is also dense on the right of every $x<1$. It means that every real that is generic from the right is weakly1 -generic, i.e., belongs to every dense effective open set. The construction of a weakly-1-generic left-c.e. real presented in [14] actually builds a real that is generic on the right, hence proves Theorem 4.7.

Again the theorem implies that Axiom 2 is satisfied, i.e., that LCE is not meager in itself. Axiom 4 is satisfied as CE is meager in LCE: we know from the previous section that every c.e. set it outside some dense effective open set.

However this time Axiom 1 is not satisfied! By definition a left-c.e. real that is generic on the right does not belong to the left-boundary of any effective open class. We need to add more nowhere dense classes. However we were unable to identify a natural way of doing it.

### 4.3 Other Classes of Objects

In [5] we introduce a way of defining a notion of genericity for more general classes of enumerable objects and prove a Baire Category theorem for such classes. For instance, one can define what is a generic $\Pi_{1}^{0}$ subset of $2^{\mathbb{N}}$. We apply this method to obtain a result in computable analysis related to the non-computability of the ergodic decomposition theorem. In [6] we give other applications of this method showing that many complicated constructions in recursion theory can be more easily obtained by choosing the suitable topology on the space of objects, and using the corresponding Baire Category theorem on that space.

## 5 Conclusion

For many classes of constructible objects it is possible to define a notion of meager subclass and a corresponding notion of genericity. A typical element of the class is then generic and automatically satisfies many interesting properties. The relevance of these notions can be measured in two ways that oppose to each other:

- A Baire Category theorem should hold, i.e., the class should not be meager in itself. Said differently, the notion of genericity should not be too strong.
- The notion of genericity should be strong enough to capture many useful interesting properties.

When these conditions are satisfied, existence results become easy to derive.
In this note, the Baire Category theorem for classes of enumerable objects is proved using the simplest form of priority method with finite injury. A future direction would be to define weaker notions of meager subclass, or stronger notions of genericity, capturing more advanced methods from recursion theory such as the priority method with infinite injury.

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# Computability in Symbolic Dynamics 

Emmanuel Jeandel ${ }^{1,2,3(\boxtimes)}$<br>1 Université de Lorraine, LORIA, UMR 7503, 54506 Vandoeuvre-lès-Nancy, France<br>emmanuel.jeandel@loria.fr<br>${ }^{2}$ CNRS, LORIA, UMR 7503, 54506 Vandoeuvre-lès-Nancy, France<br>${ }^{3}$ Inria, 54600 Villers-lès-Nancy, France


#### Abstract

We give an overview of the interplay between computability and symbolic dynamics.


A multidimensional shift of finite type (SFT) is a set of colorings of $\mathbb{Z}^{d}$ given by local rules. SFTs are one of the most fundamental objects in symbolic dynamics [LM95], and are well understood when $d=1$ where they can be studied using finite automata theory. The situation becomes drastically different in dimension 2, where they are sometimes called tilings of the (discrete) plane, as almost any natural question about them becomes undecidable [Ber64, Rob71].

The uncomputability of many properties has for a long time being seen as a hurdle in the study of multidimensional symbolic dynamics. Douglas Lind [Lin04] has in particular described multidimensional SFTs as "The Swamp of Undecidability. It's a place you don't want to go".

In recent years, the position has changed, as many results have proven that it is actually possible to understand quite well many properties of multidimensional dynamical systems, as long as one accepts that the answer might involve computability theory.

We present here a few examples of this phenomenon. The focus of the first few sections is on one-dimensional symbolic dynamical systems given by computable constraints, and we show how these constraints translate into computability obstructions on their dynamics. In the last part, we explain how these results may be translated to multidimensional dynamical systems given by finite means, using strong embedding theorems.

## 1 Definitions

We start with a few relevant definitions.
Let $A$ be a finite alphabet. We denote by $A^{\star}$ the set of finite words over the alphabet $A$, and by $A^{\mathbb{Z}}$ the set of biinfinite words over the alphabet $A$. The empty word will be denoted by $\epsilon$. Given a finite word $w=w_{0} w_{1} \ldots w_{n-1}$ (whose length $n$ is denoted $|w|)$ and a biinfinite word $u=\ldots u_{-1} u_{0} u_{1} \ldots$, we say that $w$ appears in $u$ (or that $u$ contains $w$ ) if there exists some position $k$ s.t. $u_{i+k}=w_{i}$ for all $0 \leq i<n$. We will also use the notion " $w$ appears in $u$ " for $u$ a finite word, with a similar definition.
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Given a set $\mathcal{F} \subseteq \mathcal{A}^{\star}$ of words, the subshift defined by $\mathcal{F}$ is the set of all biinfinite words where no word of $\mathcal{F}$ appears. We usually denote by $X_{\mathcal{F}}$ the subshift defined by $\mathcal{F}$.

Example 1. Let $A=\{0,1\}$. Let $\mathcal{F}=\{00\}$. Then $X_{\mathcal{F}}$ is the set of biinfinite words that do not contain two consecutive symbols 0 . For $\mathcal{F}=\{0,01,11\}, X_{\mathcal{F}}$ is evidently the empty set. For an alphabet $A$, let $\mathcal{F}=\left\{u u\left|u \in A^{\star},|u| \geq 1\right\}\right.$. Then $X_{\mathcal{F}}$ is the set of biinfinite words that do not contain any square.

Definition 1. $A$ subset $X$ of $A^{\mathbb{Z}}$ is a subshift if there exists $\mathcal{F}$ s.t. $X=X_{\mathcal{F}}$. $\mathcal{F}$ will be called a set of forbidden patterns for $X$. If $\mathcal{F}$ can be chosen finite, $X$ is called a subshift of finite type (SFT for short).

If $\mathcal{F}$ is recursively enumerable, $X$ is called an effectively closed subshift.
Example 2. The first two previous examples are obviously subshifts of finite type by definition. If $A$ is a two letter alphabet, the set of biinfinite words that do not contain any square is a subshift of finite type. Indeed, in this case, $X=\emptyset=X_{\{\epsilon\}}$. If $A$ has more than two letters, this subshift is nonempty and it is easy to see it is not of finite type. In any case, all these subshifts are effectively closed.

The set $S$ of all words over the alphabet $A=\{0,1\}$ with exactly one symbol 1 is not a subshift. Indeed, suppose there is $\mathcal{F}$ s.t. $S=X_{\mathcal{F}}$. Then $\mathcal{F}$ cannot contain any word consisting only of the symbol 0 , therefore the biinfinite word containing only 0 is in $X_{\mathcal{F}}$, a contradiction.

As made evident by the previous examples, the same subshift can be given by different sets of forbidden patterns. In set theoretical terms, there is a largest set: If $X$ is a subshift, then the set $\mathcal{B}(X)$ of all words that do not appear in any word of $X$ is a set of forbidden patterns for $X$, that is $X=X_{\mathcal{B}(X)}$ and it is clearly maximal.

In terms of computability, it is however the minimal possible description of $X$, in the following sense.

Definition 2. (Enumeration-Reducibility [FR59]). Let $S \subseteq A^{\star}$ and $S^{\prime} \subseteq$ $A^{\star}$ be two sets of finite words.

We say that $S$ is enumeration-reducible to $S^{\prime}$, in symbols $S \leq_{e} S^{\prime}$, if there is a computable procedure that can enumerate $S$ given any enumeration of $S^{\prime}$.

Formally, there exists a partial computable function $f$ that associates to any pair $(u, n) \in A^{\star} \times \mathbb{N}$ a finite subset of $A^{\star}$ s.t. $u \in S \Longleftrightarrow \exists n, f(u, n) \subseteq S^{\prime}$.

Enumeration reducibility gives rise naturally to a notion of enumerationequivalence $\equiv_{e}$ whose classes are usually called enumeration degrees.

Proposition 1. Let $X=X_{\mathcal{F}}$ be a subshift. Then $\mathcal{B}(X) \leq_{e} \mathcal{F}$.
In terms of enumeration reducibility, $\mathcal{B}(X)$ is therefore the smallest possible description of $X$.

The key to understand this proposition is the compactness property: $u \in$ $\mathcal{B}(X)$ iff there exists a size $k>|u|$ s.t. all elements of $A^{\star}$ of size $k$ that contain
$u$ also contain some element of $\mathcal{F}$. This gives a way to enumerate all finite set $F$ of words that "force" $u$ to be forbidden.

In particular, if $X$ is an effectively closed subshift, its set of forbidden words $\mathcal{B}(X)$ is recursively enumerable. In general, the language $\mathcal{B}(X)$ can be arbitrarily complex. For example given a subset $S$ of $\mathbb{N}$, it is easy to see that the subshift defined by the set of forbidden words $\left\{10^{n} 1 \mid n \in S\right\}$ has the same enumeration degree as $S$.

Dynamical systems. Symbolic dynamics is well equipped to study general dynamical systems. Let $f: A^{\mathbb{N}} \rightarrow A^{\mathbb{N}}$ be a continuous map. The itinerary of $f$ from point $x$ is the infinite word of $A^{\mathbb{N}}$ defined by $\operatorname{It}(f)(x)_{n}=a$ if the first symbol of $f^{n}(x)$ is $a$. If $f$ is bijective, it is more natural to consider biinfinite trajectories, defined similarly for $n \in \mathbb{Z}$.

Then it is easy to see that the set $I t(f)$ of all itineraries of $f$ is a subshift. Moreover computability properties of $f$ translate into computability properties of $I t(f)$, see [Das08, CDK08] for more details. Interesting examples appear when $f$ is taken to be a map of the interval [Moo91], a cellular automaton, or a Turing machine [Kur97].

## 2 Computability of Subshifts

In this section we investigate computability properties of subshifts, and in particular of points inside a (nonempty) subshift. Typical properties we are interested in is whether a subshift contains a computable point and more generally on the structure of the Turing degrees of subshifts.

Effective subshifts are examples of $\Pi_{1}^{0}$ classes [CR98], a recursion-theoretic concept that appear everywhere in mathematics. $\Pi_{1}^{0}$ classes (of sets) can be defined using forbidden positioned words, i.e. they are given by a (recursively enumerable) list of pairs of the form $(i, w)$, meaning that $w$ is forbidden to appear at position $i$. This definition, while slightly nonstandard, makes it obvious that effective subshifts are indeed $\Pi_{1}^{0}$ classes, and from this we can obtain a large number of results on what points of effective subshifts look like [Kre53, Sho60, JS72b].

However, subshifts have the additional property of shift-invariance: if $x \in X$ then the shift $\sigma(x)$ of $x$ (defined formally by $\left.\sigma(x)_{i}=x_{i+1}\right)$ is also in $X$. Whether this property translates into computability properties on elements of $X$ is the main question.

While our focus in on effectively closed subshifts, note that the above questions also make sense for general subshifts.

Cenzer et al. [CDK08] produced an example of an (nonempty) effectively closed subshift with no computable points. Another example is given by Rumyantsev and Ushakov [RU06]: Forbid all words $x$ of Kolmogorov complexity less than $|x| / 2+c$. This subshift is nonempty if $c$ is sufficiently large. More generally, Miller [Mil12] proved that any $\Pi_{1}^{0}$ set is Medvedev equivalent to an effectively closed subshift.

Definition 3. Let $S, S^{\prime}$ two subsets of $A^{\mathbb{Z}}$. We say that $S \leq_{M} S^{\prime}$ is Medvedev reducible to $S^{\prime}$ if there is a Turing functional $\Phi$ such that $\Phi\left(S^{\prime}\right) \subseteq S$.

Medvedev equivalence is usually introduced in the context of mass problems [Sim11]: $S \leq_{M} S^{\prime}$ if it is easier to find an element of $S$ than to find an element of $S^{\prime}$, in the sense that, if we find an element $y$ of $S^{\prime}$, then we also obtain in the same way an element of $S$ (namely $\Phi(y)$ ).

Theorem 1 [Mil12]. For any $\Pi_{1}^{0}$ set $S$, there is an effectively closed subshift $S^{\prime}$ that is Medvedev-equivalent to $S$.

Note that Medvedev equivalence is a weak notion in the sense that it speaks somehow only about the easiest (in terms of Turing degrees) elements of a set: Two sets $S$ and $S^{\prime}$ which both contain computable points are always Medvedev equivalent.

We can search for something (somewhat) stronger: Given a $\Pi_{1}^{0}$ set $S$, is there a subshift with the same set of Turing degrees? The answer is negative:

Theorem 2 [JV13]. Let $S$ be a subshift. Then either $S$ contains a computable point, or it contains a cone of Turing degrees: There exists a point $x \in S$ s.t. there are points $y \in S$ of arbitrary Turing degree above the degree of $x$.

In particular, if $S$ has no computable point, it contains two points of different but comparable Turing degrees. However we can construct some $\Pi_{1}^{0}$ classes which do not have this property [JS72a], which proves it is rather specific to subshifts. Note that this theorem is true for any subshift, and not only for effectively closed subshifts.

This property is due to the fact that every nonempty subshift contains a point $x$ with a peculiar property, called uniform recurrence: If a word $u$ appears in $x$, there exists a size $n$ s.t. $u$ occurs in every word of size $n$ that appear in $x$. An obvious example of an uniformly recurrent word is a periodic word $w$ ( $w_{i}=w_{i+n}$ for some $n$ and all $i$ ). Another classical example is the Thue-Morse word.

The situation when $S$ has a computable point is completely understood and subsumed in the following theorem:

Theorem 3 [JV13]. For any set $S$ with a computable point, there is a subshift $T$ s.t. $S$ and $T$ have the same Turing degrees.

Moreover, the set of positioned words that do not appear in $S$ is enumerationequivalent to the set of words that do not appear in $T$.
(In particular, if $S$ is a $\Pi_{1}^{0}$ class, then $T$ is effectively closed).
The situation of sets with cones of Turing degrees is less understood. Hochman and Vanier [HV] produced examples of subshifts for which the Turing degree spectrum is an uncountable union of disjoint cones.

## 3 Multidimensional Subshifts

In the previous section, the subshifts with specific properties that are produced are often effectively closed, but never of finite type. Indeed, the theory of subshifts of finite type is well understood in dimension one, and connected with finite automata theory [LM95].

The situation is dramatically different when dealing with multi-dimensional subshifts. The definition of multi-dimensional subshifts is similar to onedimensional subshifts, where a configuration is now an element of $A^{\mathbb{Z}^{d}}$ for some $d$, and the concept of a pattern (an element of $A^{n^{d}}$ ) replaces the concept of a word.

However, subshifts of finite type now become interesting. Indeed, it is undecidable to know if a subshift of finite type (given by a list of forbidden patterns) is empty [Ber64], and there exist nonempty subshifts of finite type where no configuration is computable [Mye74].

The main reason for these theorems is the ease of coding the space-time diagram of a Turing machine as a two-dimensional configuration. While these results where obtained in the late 60 s and early 70 s , they are now better understood in the context of the embedding theorems of the next paragraph.

### 3.1 The Embedding Theorems

The embedding theorems state that one-dimensional effectively closed subshifts may be encoded into multi-dimensional subshifts of finite type. Using this embedding, many of the previous theorems can be prove to hold for multidimensional subshifts of finite type.

To present the theorems, a few definitions are needed.
If $S$ is a subshift over the alphabet $A$ in dimension $d$, and $\pi$ a map from $A$ to $B$, the recoloring $\pi(S)$ is the subset of $B^{\mathbb{Z}^{d}}$ of all configurations $y$ s.t. there exists $y \in S$ s.t. $y_{i}=\pi(x)_{i}$ for all $i$. The recoloring can be thought of as a way somehow to ignore construction lines by recoloring them. Note that however a recoloring subshift is never more complex than the original subshift: Indeed, $\mathcal{B}(\pi(S)) \leq_{e} \mathcal{B}(S)$.

Given a subshift $S \subseteq A^{\mathbb{Z}^{d}}$ in dimension $d$, one can define naturally higher and lower dimensional versions of $S$. The higher dimensional version $S^{\mathbb{Z}^{d^{\prime}-d}}$ is the set of all configurations $x$ of $A^{\mathbb{Z}^{d^{\prime}}}$ for which there exists $y \in S$ s.t. $x_{(i, j)}=y_{i}$ for all $i \in \mathbb{Z}^{d}$ and $j \in \mathbb{Z}^{d^{\prime}-d}$. If $d=1$ and $d^{\prime}-d=1, S^{\mathbb{Z}}$ is therefore the set of all two-dimensional configurations where all rows are identical to some element of $S$. It is easy to see that $\mathcal{B}\left(S^{\mathbb{Z}^{d^{\prime}}}\right) \equiv_{e} \mathcal{B}(S)$.

Then we have the following theorem:
Theorem 4 [Hoc09,AS13,DRS10]. Let $S$ be an effectively closed subshift of dimension $d$ over the alphabet $A$. Then there exist $d^{\prime}$, a recoloring $\pi$, and $a$ subshift of finite type $S^{\prime}$ s.t. $S^{\mathbb{Z}^{d^{\prime}}}=\pi\left(S^{\prime}\right)$.

In this theorem, we can take $d^{\prime}=1$.

There is a relativized version of this statement. Say that $S^{\prime}$ (over the alphabet $B$ ) is of finite type over $S$ (over the alphabet $A \subseteq B$ ) if there exists a finite set $\mathcal{F}$ of forbidden patterns s.t. $S^{\prime}$ is obtained from $S$ by adding these forbidden patterns: $S^{\prime}=X_{\mathcal{F} \cup \mathcal{B}(S)}$. Note that $S^{\prime}$ may have a larger alphabet than $S$. Again it is easy to see that $\mathcal{B}\left(S^{\prime}\right) \leq_{e} \mathcal{B}(S)$.

Theorem 5 [AS09]. Let $S_{1}$ and $S_{2}$ be two subshifts.
Then $\mathcal{B}\left(S_{1}\right) \leq_{e} \mathcal{B}\left(S_{2}\right)$ iff there exist integers $d_{1}, d_{2}$, a recoloring $\pi$, and $S^{\prime}$ of finite type over $S_{2}^{\mathbb{Z}^{d_{2}}}$ s.t. $S_{1}^{\mathbb{Z}^{d_{1}}}=\pi\left(S^{\prime}\right)$

If we start from $S_{2}=\{0\}^{\mathbb{Z}}$, we recover the previous theorem. It is interesting to note that these theorems are analogues of respectively the Highman embedding theorem [Hig61] and the relative Highman embedding theorems for groups [HS88], with subshifts (of finite type/effectively closed) playing the role of groups (finitely presented/recursively presented).

This gives a way to produce subshifts of finite type with complex behaviours: starting from a effectively closed subshift in dimension one with a given property, we obtain this way a subshift of finite type with the same property. Not all properties are preserved by recolorings and higher-dimensional versions, but enough are.

As an example, if we start from a one-dimensional effectively closed subshift with no computable point, we obtain a two-dimensional subshift of finite type with no computable point. If we start from an effectively closed subshift that is Medvedev equivalent to some $\Pi_{1}^{0}$ set $S$, we obtain a two-dimensional subshift of finite type that is Medvedev equivalent to $S$, a result originally from Simpson [Sim14] using a method from Myers [Mye74].

It is therefore reasonable to think of multi-dimensional subshifts of finite type as having similar computational properties as one-dimensional effectively closed subshifts.

### 3.2 Peculiarities of Subshifts of Finite Type

We finish this section by presenting some results that cannot be proven by the embedding theorem, either due to the nature of the subshift $S^{\prime}$ that is constructed, or due to the fact that our computational properties are not invariant under recoloring.

The first property deals with countable subshifts. In the proof of the embedding theorem, the subshift $S^{\prime}$ is uncountable, and this cannot be corrected. In particular, the theorem cannot be used to prove results on countable subshifts of finite type. Nevertheless, we may obtain:

Theorem 6 [JV13]. Let $S$ be a countable $\Pi_{1}^{0}$ set. Then there exists a countable subshift of finite type $T$ s.t. $S$ and $T$ have the same set of Turing degrees.

The second property has to do with periodic configurations. In a subshift of finite type, it is algorithmically decidable to know whether there exists a configuration periodic of period $n$ in all directions, as we only have to test all
possible hypercubes of size $n$. This is not true anymore of the recoloring of a subshift of finite type: $\pi(x)$ might be periodic of period $n$ without $x$ being periodic. In fact it is easy to prove (using e.g. the embedding theorem) that the problem has now become undecidable.

However it is possible to obtain a strong characterization of what may happen for a subshift of finite type

Theorem 7 [JV15]. Let $X$ be a subshift of finite type. Then the set of all $n$ s.t. $X$ contains a configuration of period exactly $n$ in all directions is in NP (when $n$ is encoded in unary).

Conversely, given a unary language $L$ in $\mathbf{N P}$, there exists a subshift of finite type $X$ s.t. the set of all $n$ s.t. $X$ contains a configuration of period exactly $n$ in all directions is exactly $L$.

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# Using Semidirect Product of (Semi)groups in Public Key Cryptography 

Delaram Kahrobaei ${ }^{1(\boxtimes)}$ and Vladimir Shpilrain ${ }^{2}$<br>${ }^{1}$ CUNY Graduate Center and City Tech, City University of New York, New York, USA<br>dkahrobaei@gc.cuny.edu<br>${ }^{2}$ The City College of New York and CUNY Graduate Center, New York, USA<br>shpil@groups.sci.ccny.cuny.edu


#### Abstract

In this survey, we describe a general key exchange protocol based on semidirect product of (semi)groups (more specifically, on extensions of (semi)groups by automorphisms), and then focus on practical instances of this general idea. This protocol can be based on any group or semigroup, in particular on any non-commutative group. One of its special cases is the standard Diffie-Hellman protocol, which is based on a cyclic group. However, when this protocol is used with a noncommutative (semi)group, it acquires several useful features that make it compare favorably to the Diffie-Hellman protocol. The focus then shifts to selecting an optimal platform (semi)group, in terms of security and efficiency. We show, in particular, that one can get a variety of new security assumptions by varying an automorphism used for a (semi)group extension.


## 1 Introduction

The area of public key cryptography started with the seminal paper [2] introducing what is now known as the Diffie-Hellman key exchange protocol.

The simplest, and original, implementation of the protocol uses the multiplicative group of integers modulo $p$, where $p$ is prime and $g$ is primitive $\bmod p$. A more general description of the protocol uses an arbitrary finite cyclic group.

1. Alice and Bob agree on a finite cyclic group $G$ and a generating element $g$ in $G$. We will write the group $G$ multiplicatively.
2. Alice picks a random natural number $a$ and sends $g^{a}$ to Bob.

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3. Bob picks a random natural number $b$ and sends $g^{b}$ to Alice.
4. Alice computes $K_{A}=\left(g^{b}\right)^{a}=g^{b a}$.
5. Bob computes $K_{B}=\left(g^{a}\right)^{b}=g^{a b}$.

Since $a b=b a$, both Alice and Bob are now in possession of the same group element $K=K_{A}=K_{B}$ which can serve as the shared secret key.

The protocol is considered secure against eavesdroppers if $G$ and $g$ are chosen properly. The eavesdropper must solve the Diffie-Hellman problem (recover $g^{a b}$ from $g, g^{a}$ and $g^{b}$ ) to obtain the shared secret key. This is currently considered difficult for a "good" choice of parameters (see e.g. [8] for details).

There is an ongoing search for other platforms where the Diffie-Hellman or similar key exchange could be carried out more efficiently or where security would be based on different assumptions. This search already gave rise to several interesting directions, including a whole area of elliptic curve cryptography [17]. We also refer the reader to [10] or [11] for a survey of proposed cryptographic primitives based on non-abelian (= non-commutative) groups. A survey of these efforts is outside of the scope of the present paper; our goal here is to describe a new key exchange protocol from [4] based on extension of a (semi)group by automorphisms (or more generally, by self-homomorphisms) and discuss possible platforms that would make this protocol secure and efficient. This protocol can be based on any group, in particular on any non-commutative group. It has some resemblance to the classical Diffie-Hellman protocol, but there are several distinctive features that, we believe, give the new protocol important advantages. In particular, even though the parties do compute a large power of a public element (as in the classical Diffie-Hellman protocol), they do not transmit the whole result, but rather just part of it.

We then describe in this survey some particular instantiations of this general protocol. We start with a non-commutative semigroup of matrices as the platform, consider an extension of this semigroup by a conjugating automorphism and show that security of the relevant instantiation is based on a quite different security assumption compared to that of the standard Diffie-Hellman protocol. However, due to the nature of this security assumption, the protocol turns out to be vulnerable to a "linear algebra attack", similar to an attack on Stickel's protocol [16] offered in [15], albeit more sophisticated, see [9, 14]. A composition of conjugating automorphism with a field automorphism was employed in [7], but this automorphism still turned out to be not complex enough to make the protocol withstand a linear algebra attack, see [3,14].

We therefore offer here another platform group that we believe should make the protocol invulnerable to the attacks of [3], [9,14]. The group is a free nilpotent $p$-group, for a sufficiently large prime $p$. We give a formal definition of this group in Sect. 8; here we just say that this is a finite group all of whose elements have order dividing $p^{n}$ for some fixed $n \geq 1$. As any finite group, this group is linear, but Janusz [5] showed that a faithful representation of a finite $p$-group, with at least one element of order $p^{n}$, as a group of matrices over a finite field of characteristic $p$ is of dimension at least $1+p^{n-1}$, which is too large to launch a linear algebra attack provided $p$ itself is large enough. At the same time,
to keep computation in the platform group efficient, the nilpotency class of the group has to be fairly small. We note that, in contrast, the dimension of the classical representations of finitely generated torsion-free nilpotent groups in a matrix group $U T(\mathbb{Z})$ can be rather small (cf. [12]), but for torsion groups with elements of large order the situation is really different. Still, there is the usual trade-off between security and efficiency, so the following parameters have to be chosen carefully to provide for both security and efficiency: (1) the size of $p$; (2) the nilpotency class of the platform group; (3) the rank (i.e., the number of generators) of the platform group. We discuss this in our Sect. 8 .

We mention here another, rather different, proposal [13] of a cryptosystem based on the semidirect product of two groups and yet another, more complex, proposal of a key agreement based on the semidirect product of two monoids [1]. Both these proposals are very different from that of [4]. In particular, the crucial idea of transmitting just part of the result of an exponentiation appears only in [4].

Finally, we note that the basic construction (semidirect product) described in this survey can be adopted, with some simple modifications, in other algebraic systems, e.g. associative rings or Lie rings, and key exchange protocols similar to ours can be built on those.

## 2 Semidirect Products and Extensions by Automorphisms

We include this section to make the exposition more comprehensive. The reader who is uncomfortable with group-theoretic constructions can skip to Subsect. 2.1.

We now recall the definition of a semidirect product:
Definition 1. Let $G, H$ be two groups, let $A u t(G)$ be the group of automorphisms of $G$, and let $\rho: H \rightarrow \operatorname{Aut}(G)$ be a homomorphism. Then the semidirect product of $G$ and $H$ is the set

$$
\Gamma=G \rtimes_{\rho} H=\{(g, h): g \in G, h \in H\}
$$

with the group operation given by

$$
(g, h)\left(g^{\prime}, h^{\prime}\right)=\left(g^{\rho\left(h^{\prime}\right)} \cdot g^{\prime}, h \cdot h^{\prime}\right)
$$

Here $g^{\rho\left(h^{\prime}\right)}$ denotes the image of $g$ under the automorphism $\rho\left(h^{\prime}\right)$, and when we write a product $h \cdot h^{\prime}$ of two morphisms, this means that $h$ is applied first.

In this paper, we focus on a special case of this construction, where the group $H$ is just a subgroup of the group $\operatorname{Aut}(G)$. If $H=\operatorname{Aut}(G)$, then the corresponding semidirect product is called the holomorph of the group $G$. We give some more details about the holomorph in our Sects. 2.1, and 3 we describe a key exchange protocol that uses (as the platform) an extension of a group $G$ by a cyclic group of automorphisms.

### 2.1 Extensions by Automorphisms

A particularly simple special case of the semidirect product construction is where the group $H$ is just a subgroup of the group $\operatorname{Aut}(G)$. If $H=\operatorname{Aut}(G)$, then the corresponding semidirect product is called the holomorph of the group $G$. Thus, the holomorph of $G$, usually denoted by $\operatorname{Hol}(G)$, is the set of all pairs $(g, \phi)$, where $g \in G, \phi \in \operatorname{Aut}(G)$, with the group operation given by $(g, \phi) \cdot\left(g^{\prime}, \phi^{\prime}\right)=$ $\left(\phi^{\prime}(g) \cdot g^{\prime}, \phi \cdot \phi^{\prime}\right)$.

It is often more practical to use a subgroup of $\operatorname{Aut}(G)$ in this construction, and this is exactly what we do in Sect.3, where we describe a key exchange protocol that uses (as the platform) an extension of a group $G$ by a cyclic group of automorphisms.

Remark 1. One can also use this construction if $G$ is not necessarily a group, but just a semigroup, and/or consider endomorphisms (i.e., self-homomorphisms) of $G$, not necessarily automorphisms. Then the result will be a semigroup; this is what we use in our Sect. 6 .

## 3 Key Exchange Protocol

In the simplest implementation of the construction described in our Sect. 2.1, one can use just a cyclic subgroup (or a cyclic subsemigroup) of the group $\operatorname{Aut}(G)$ (respectively, of the semigroup $\operatorname{End}(G)$ of endomorphisms) instead of the whole group of automorphisms of $G$.

Thus, let $G$ be a (semi)group. An element $g \in G$ is chosen and made public as well as an arbitrary automorphism $\phi \in \operatorname{Aut}(G)$ (or an arbitrary endomorphism $\phi \in \operatorname{End}(G))$. Bob chooses a private $n \in \mathbb{N}$, while Alice chooses a private $m \in \mathbb{N}$. Both Alice and Bob are going to work with elements of the form $\left(g, \phi^{r}\right)$, where $g \in G, r \in \mathbb{N}$. Note that two elements of this form are multiplied as follows: $\left(g, \phi^{r}\right) \cdot\left(h, \phi^{s}\right)=\left(\phi^{s}(g) \cdot h, \phi^{r+s}\right)$.

1. Alice computes $(g, \phi)^{m}=\left(\phi^{m-1}(g) \cdots \phi^{2}(g) \cdot \phi(g) \cdot g, \phi^{m}\right)$ and sends only the first component of this pair to Bob. Thus, she sends to Bob only the element $a=\phi^{m-1}(g) \cdots \phi^{2}(g) \cdot \phi(g) \cdot g$ of the (semi)group $G$.
2. Bob computes $(g, \phi)^{n}=\left(\phi^{n-1}(g) \cdots \phi^{2}(g) \cdot \phi(g) \cdot g, \phi^{n}\right)$ and sends only the first component of this pair to Alice. Thus, he sends to Alice only the element $b=\phi^{n-1}(g) \cdots \phi^{2}(g) \cdot \phi(g) \cdot g$ of the (semi)group $G$.
3. Alice computes $(b, x) \cdot\left(a, \phi^{m}\right)=\left(\phi^{m}(b) \cdot a, x \cdot \phi^{m}\right)$. Her key is now $K_{A}=$ $\phi^{m}(b) \cdot a$. Note that she does not actually "compute" $x \cdot \phi^{m}$ because she does not know the automorphism $x=\phi^{n}$; recall that it was not transmitted to her. But she does not need it to compute $K_{A}$.
4. Bob computes $(a, y) \cdot\left(b, \phi^{n}\right)=\left(\phi^{n}(a) \cdot b, y \cdot \phi^{n}\right)$. His key is now $K_{B}=\phi^{n}(a) \cdot b$. Again, Bob does not actually "compute" $y \cdot \phi^{n}$ because he does not know the automorphism $y=\phi^{m}$.
5. Since $(b, x) \cdot\left(a, \phi^{m}\right)=(a, y) \cdot\left(b, \phi^{n}\right)=(g, \phi)^{m+n}$, we should have $K_{A}=$ $K_{B}=K$, the shared secret key.

Remark 2. Note that, in contrast with the "standard" Diffie-Hellman key exchange, correctness here is based on the equality $h^{m} \cdot h^{n}=h^{n} \cdot h^{m}=h^{m+n}$ rather than on the equality $\left(h^{m}\right)^{n}=\left(h^{n}\right)^{m}=h^{m n}$. In the "standard" DiffieHellman set up, our trick would not work because, if the shared key $K$ was just the product of two openly transmitted elements, then anybody, including the eavesdropper, could compute $K$.

## 4 Computational Cost

From the look of transmitted elements in the protocol in Sect.3, it may seem that the parties have to compute a product of $m$ (respectively, $n$ ) elements of the (semi)group $G$. However, since the parties actually compute powers of an element of $G$, they can use the "square-and-multiply" method, as in the standard Diffie-Hellman protocol. Then there is a cost of applying an automorphism $\phi$ to an element of $G$, and also of computing powers of $\phi$. These costs depend, of course, on a specific platform (semi)group that is used with our protocol and on a specific automorphism that is used for a (semi)group extension. In our first, "toy" example (Sect. 5 below), both applying an automorphism $\phi$ and computing its powers amount to exponentiation of elements of $G$, which can be done again by the "square-and-multiply" method. In our example in Sect. $6, \phi$ is a conjugation, so applying $\phi$ amounts to just two multiplications of elements in $G$, while computing powers of $\phi$ amounts to exponentiation of two elements of $G$ (namely, of the conjugating element and of its inverse).

Thus, in either instantiation of our protocol considered in this paper, the cost of computing $(g, \phi)^{n}$ is $O(\log n)$, just as in the standard Diffie-Hellman protocol. Computational cost analysis for the platform group suggested in Sect. 8 is somewhat more delicate; we refer to Sect. 8.1 for more details.

## 5 "Toy Example": Multiplicative $\mathbb{Z}_{p}^{*}$

As one of the simplest instantiations of our protocol, we use here the multiplicative group $\mathbb{Z}_{p}^{*}$ as the platform group $G$ to illustrate what is going on. In selecting a prime $p$, as well as private exponents $m, n$, one can follow the same guidelines as in the "standard" Diffie-Hellman.

Selecting the (public) endomorphism $\phi$ of the group $\mathbb{Z}_{p}^{*}$ amounts to selecting yet another integer $k$, so that for every $h \in \mathbb{Z}_{p}^{*}$, one has $\phi(h)=h^{k}$. If $k$ is relatively prime to $p-1$, then $\phi$ is actually an automorphism. Below we assume that $k>1$.

Then, for an element $g \in \mathbb{Z}_{p}^{*}$, we have:

$$
(g, \phi)^{m}=\left(\phi^{m-1}(g) \cdots \phi(g) \cdot \phi^{2}(g) \cdot g, \phi^{m}\right)
$$

We focus on the first component of the element on the right; easy computation shows that it is equal to $g^{k^{m-1}+\ldots+k+1}=g^{\frac{k^{m}-1}{k-1}}$. Thus, if the adversary chooses
a "direct" attack, by trying to recover the private exponent $m$, he will have to solve the discrete log problem twice: first to recover $\frac{k^{m}-1}{k-1}$ from $g^{\frac{k^{m}-1}{k-1}}$, and then to recover $m$ from $k^{m}$. (Note that $k$ is public since $\phi$ is public.)

On the other hand, the analog of what is called "the Diffie-Hellman problem" would be to recover the shared key $K=g^{\frac{k^{m+n}-1}{k-1}}$ from the triple $\left(g, g^{\frac{k^{m}-1}{k-1}}, g^{\frac{k^{n}-1}{k-1}}\right)$. Since $g$ and $k$ are public, this is equivalent to recovering $g^{k^{m+n}}$ from the triple $\left(g, g^{k^{m}}, g^{k^{n}}\right)$, i.e., this is exactly the standard DiffieHellman problem.

Thus, the bottom line of this example is that the instantiation of our protocol where the group $G$ is $\mathbb{Z}_{p}^{*}$, is not really different from the standard Diffie-Hellman protocol. In the next section, we describe a more interesting instantiation, where the (semi)group $G$ is non-commutative.

## 6 Matrices Over Group Rings and Extensions by Inner Automorphisms

Our exposition here follows [4]. To begin with, we note that the general protocol in Sect. 3 can be used with any non-commutative group $G$ if $\phi$ is selected to be a non-trivial inner automorphism, i.e., conjugation by an element which is not in the center of $G$. Furthermore, it can be used with any non-commutative semigroup $G$ as well, as long as $G$ has some invertible elements; these can be used to produce inner automorphisms. A typical example of such a semigroup would be a semigroup of matrices over some ring.

In the paper [6], the authors have employed matrices over group rings of a (small) symmetric group as platforms for the (standard) Diffie-Hellman-like key exchange. In this section, we use these matrix semigroups again and consider an extension of such a semigroup by an inner automorphism to get a platform semigroup for the general protocol in Sect. 3 .

Recall that a (semi)group ring $R[S]$ of a (semi)group $S$ over a commutative ring $R$ is the set of all formal sums $\sum_{g_{i} \in S} r_{i} g_{i}$, where $r_{i} \in R$, and all but a finite number of $r_{i}$ are zero.

The sum of two elements in $R[G]$ is defined by

$$
\left(\sum_{g_{i} \in S} a_{i} g_{i}\right)+\left(\sum_{g_{i} \in S} b_{i} g_{i}\right)=\sum_{g_{i} \in S}\left(a_{i}+b_{i}\right) g_{i}
$$

The multiplication of two elements in $R[G]$ is defined by using distributivity.
As we have already pointed out, if a (semi)group $G$ is non-commutative and has non-central invertible elements, then it always has a non-identical inner automorphism, i.e., conjugation by an element $g \in G$ such that $g^{-1} h g \neq h$ for at least some $h \in G$.

Now let $G$ be the semigroup of $3 \times 3$ matrices over the group ring $\mathbb{Z}_{7}\left[A_{5}\right]$, where $A_{5}$ is the alternating group on 5 elements. Here we use an extension of the semigroup $G$ by an inner automorphism $\varphi_{H}$, which is conjugation by a matrix
$H \in G L_{3}\left(\mathbb{Z}_{7}\left[A_{5}\right]\right)$. Thus, for any matrix $M \in G$ and for any integer $k \geq 1$, we have

$$
\varphi_{H}(M)=H^{-1} M H ; \varphi_{H}^{k}(M)=H^{-k} M H^{k} .
$$

Now the general protocol from Sect. 3 is specialized in this case as follows.

1. Alice and Bob agree on public matrices $M \in G$ and $H \in G L_{3}\left(\mathbb{Z}_{7}\left[A_{5}\right]\right)$. Alice selects a private positive integer $m$, and Bob selects a private positive integer $n$.
2. Alice computes $\left(M, \varphi_{H}\right)^{m}=\left(H^{-m+1} M H^{m-1} \cdots H^{-2} M H^{2} \cdot H^{-1} M H\right.$. $\left.M, \varphi_{H}^{m}\right)$ and sends only the first component of this pair to Bob. Thus, she sends to Bob only the matrix

$$
A=H^{-m+1} M H^{m-1} \cdots H^{-2} M H^{2} \cdot H^{-1} M H \cdot M=H^{-m}(H M)^{m}
$$

3. Bob computes $\left(M, \varphi_{H}\right)^{n}=\left(H^{-n+1} M H^{n-1} \cdots H^{-2} M H^{2} \cdot H^{-1} M H \cdot M, \varphi_{H}^{n}\right)$ and sends only the first component of this pair to Alice. Thus, he sends to Alice only the matrix

$$
B=H^{-n+1} M H^{n-1} \cdots H^{-2} M H^{2} \cdot H^{-1} M H \cdot M=H^{-n}(H M)^{n} .
$$

4. Alice computes $(B, x) \cdot\left(A, \varphi_{H}^{m}\right)=\left(\varphi_{H}^{m}(B) \cdot A, x \cdot \varphi_{H}^{m}\right)$. Her key is now $K_{\text {Alice }}=\varphi_{H}^{m}(B) \cdot A=H^{-(m+n)}(H M)^{m+n}$. Note that she does not actually "compute" $x \cdot \varphi_{H}^{m}$ because she does not know the automorphism $x=\varphi_{H}^{n}$; recall that it was not transmitted to her. But she does not need it to compute $K_{\text {Alice }}$.
5. Bob computes $(A, y) \cdot\left(B, \varphi_{H}^{n}\right)=\left(\varphi_{H}^{n}(A) \cdot B, y \cdot \varphi_{H}^{n}\right)$. His key is now $K_{B o b}=$ $\varphi_{H}^{n}(A) \cdot B$. Again, Bob does not actually "compute" $y \cdot \varphi_{H}^{n}$ because he does not know the automorphism $y=\varphi_{H}^{m}$.
6. Since $(B, x) \cdot\left(A, \varphi_{H}^{m}\right)=(A, y) \cdot\left(B, \varphi_{H}^{n}\right)=\left(M, \varphi_{H}\right)^{m+n}$, we should have $K_{\text {Alice }}=K_{\text {Bob }}=K$, the shared secret key.

## 7 Security Assumptions

In this section, we address the question of security of the protocol described in Sect. 6.

Recall that the shared secret key in the protocol of Sect. 6 is

$$
K=\varphi_{H}^{m}(B) \cdot A=\varphi_{H}^{n}(A) \cdot B=H^{-(m+n)}(H M)^{m+n}
$$

Therefore, our security assumption here is that it is computationally hard to retrieve the key $K=H^{-(m+n)}(H M)^{m+n}$ from the quadruple $\left(H, M, H^{-m}(H M)^{m}, H^{-n}(H M)^{n}\right)$.

In particular, we have to take care that the matrices $H$ and $H M$ do not commute because otherwise, $K$ is just a product of $H^{-m}(H M)^{m}$ and $H^{-n}(H M)^{n}$.

A weaker security assumption arises if an eavesdropper tries to recover a private exponent from a transmission, i.e., to recover, say, $m$ from $H^{-m}(H M)^{m}$.

A special case of this problem, where $H=I$, is the "discrete log" problem for matrices over $\mathbb{Z}_{7}\left[A_{5}\right]$, namely: recover $m$ from $M$ and $M^{m}$.

As we have mentioned in the Introduction, the protocol in this section was attacked in $[9,14]$ by a "linear algebra attack". This was possible partly because of the special "compact" form of the above security assumptions, and partly because the dimension of a linear representation of the platform semigroup happens to be small enough in this case for a linear algebra attack to be computationally feasible. In the following Sect. 8, we offer another platform that does not have these vulnerabilities.

## 8 Nilpotent Groups and $\boldsymbol{p}$-groups

First we recall that a free group $F_{r}$ on $x_{1}, \ldots, x_{r}$ is the set of reduced words in the alphabet $\left\{x_{1}, \ldots, x_{r}, x_{1}^{-1}, \ldots, x_{r}^{-1}\right\}$. A reduced word is a word without subwords $x_{i} x_{i}^{-1}$ or $x_{i}^{-1} x_{i}$. The multiplication on this set is concatenation of two words, followed by canceling out all subwords $x_{i} x_{i}^{-1}$ and $x_{i}^{-1} x_{i}$ until the word becomes reduced.

It is a fact that every group that can be generated by $r$ elements is the factor group of $F_{r}$ by an appropriate normal subgroup. We are now going to define two special normal subgroups of $F_{r}$.

The normal subgroup $F_{r}^{p}$ is generated (as a group) by all elements of the form $g^{p}, g \in F_{r}$. In the factor group $F_{r} / F_{r}^{p}$ every nontrivial element therefore has order $p$ (if $p$ is a prime). More generally, if $n \geq 2$ is an arbitrary integer, then the order of any element of $F_{r} / F_{r}^{n}$ divides $n$.

The other normal subgroup that we need is somewhat less straightforward to define. Let $[a, b]$ denote $a^{-1} b^{-1} a b$. Then, inductively, let $\left[y_{1}, \ldots, y_{c+1}\right]$ denote [ $\left.\left[y_{1}, \ldots, y_{c}\right], y_{c+1}\right]$. For a group $G$, denote by $\gamma_{c}(G)$ the (normal) subgroup of $G$ generated (as a group) by all elements of the form $\left[y_{1}, \ldots, y_{c}\right]$. If $\gamma_{c+1}(G)=\{1\}$, we say that the group $G$ is nilpotent of nilpotency class $c$.

The factor group $F_{r} / \gamma_{c+1}\left(F_{r}\right)$ is called the free nilpotent group of nilpotency class $c$. This group is infinite; however, the group we define in the following subsection is finite, and we are going to recommend it as the platform for the cryptographic scheme based on a semidirect product.

### 8.1 Free Nilpotent $\boldsymbol{p}$-group

The group $G=F_{r} / F_{r}^{p^{2}} \cdot \gamma_{c+1}\left(F_{r}\right)$ is what we suggest to use as the platform for the key exchange protocol in Sect. 3.

This group, being a nilpotent $p$-group, is finite. Its order depends on $p, c$, and $r$. For efficiency reasons, it seems better to keep $c$ and $r$ fairly small (in particular, we suggest $c=2$ or 3 ), while $p$ should be large enough to make the dimension of linear representations of $G$ so large that a linear algebra attack would be infeasible. As we have mentioned in the Introduction, a faithful representation of a finite $p$-group, with at least one element of order $p^{n}$, as a group of matrices over a finite field of characteristic $p$ is of dimension at least $1+p^{n-1}$ [5], so in
our case it is of dimension at least $1+p$. Thus, if $p$ is, say, a 100 -bit number, a linear algebra attack is already infeasible.

At the same time, we want computation in the group $G$ to be efficient. Also, we want transmitted elements to be in some kind of standard form, usually called a normal form. Here is how a normal form looks like if nilpotency class $c=2$ :

$$
x_{1}^{\alpha_{1}} \cdots x_{i}^{\alpha_{i}} \cdots x_{r}^{\alpha_{r}}\left[x_{1}, x_{2}\right]^{\beta_{1,2}} \cdots\left[x_{i}, x_{j}\right]^{\beta_{i, j}} \cdots\left[x_{r-1}, x_{r}\right]^{\beta_{r-1, r}},
$$

where $\alpha_{i}$ and $\beta_{i, j}$ are integers and in every $\left[x_{i}, x_{j}\right]$ above one has $i<j$. Different collections of $\alpha_{i}$ and $\beta_{i, j}$ produce different elements of $G$ as long as $0 \leq \alpha_{i}, \beta_{i, j}<p^{2}$, so $G$ in this case has at least $p^{2 r+r(r-1)}=p^{r^{2}+r}$ elements, which is a large number even if $r$ is fairly small. At the same time, group operations (i.e., multiplication and inversion) in $G$ are quite efficient. Indeed, multiplying two elements in the above form essentially amounts to re-writing a product $x_{1}^{\alpha_{1}} \cdots x_{r}^{\alpha_{r}} \cdot x_{1}^{\alpha_{1}^{\prime}} \cdots x_{r}^{\alpha_{r}^{\prime}}$ in the normal form. This is because commutators $\left[x_{i}, x_{j}\right]$ commute with any element of $G$ (since $c=2$ ), so collecting all $\left[x_{i}, x_{j}\right]$ in the right place takes (almost) linear time in the length of an input. Now re-writing a product of powers of $x_{i}$ in the normal form is not too hard either because $\left[x_{i}^{a}, x_{j}^{b}\right]=\left[x_{i}, x_{j}\right]^{a b}$ in the group $G$ (again, since $c=2$ ). Thus, re-writing will take at most quadratic time in the length of an input.

Applying an endomorphism (i.e., a self-homomorphism) $\phi$ given as a map $\phi\left(x_{i}\right)=y_{i}$ on the generators is efficient, too. This is due to the fact that in any group $G$ of nilpotency class 2, one has: (1) $a b=b a$ if either $a$ or $b$ (or both) belong to $\gamma_{2}(G) ;(2)[a b, c]=[a, c][b, c]$ and $[a, b c]=[a, b][a, c]$ for any $a, b, c \in G ;(3)(a b)^{n}=a^{n} b^{n}[b, a]^{\frac{n(n-1)}{2}}$ for any $a, b \in G$. Using these identities, one can reduce $\phi(g)$ to the normal form in at most quadratic time in the length of $g \in G$, provided $g$ itself was in the normal form.

The group $G$ has another property useful for our purposes. We note that the subgroup $F_{r}^{p^{2}} \cdot \gamma_{c+1}\left(F_{r}\right)$ of $F_{r}$ is, in fact, fully invariant, i.e., is invariant under any endomorphism of $F_{r}$. This implies that the group $G$ has a lot of endomorphisms because any map on the generators of $G$ can be extended (by the homomorphic property) to an endomorphism of $G$. Thus, if $G$ has $r$ generators and $m$ elements altogether, then it has $m^{r}$ endomorphisms. Even if $r$ is very small (say, $r=3$ ), this number is huge because, as we have just seen, $G$ has at least $p^{r^{2}+r}$ elements, so with a 100-bit $p$, we are going to have at least $2^{3600}$ endomorphisms. Of course, we want our endomorphism $\phi$ not to have short cycles (i.e., if $\phi^{m}=\phi^{n}$, then $|m-n|$ has to be quite large). This is easier to guarantee if $\phi$ is actually an automorphism because then we can sample from automorphisms having a large order, and these correspond to matrices from $G L_{r}\left(\mathbb{Z}_{p^{2}}\right)$ that have large order. Sampling matrices of large order from that group is not completely trivial, but we leave this outside of the scope of this survey. Here we just mention that for most automorphisms of $G$, relevant security assumptions will not have a compact form like that in Sect. 7 because a product of the form $\phi^{m-1}(g) \cdots \phi^{2}(g) \cdot \phi(g) \cdot g$ (see the general protocol in our Sect.3) typically does not simplify much.

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# Towards Computational Complexity Theory on Advanced Function Spaces in Analysis 

Akitoshi Kawamura ${ }^{1}$, Florian Steinberg ${ }^{1,2}$, and Martin Ziegler ${ }^{2,3(\boxtimes)}$<br>${ }^{1}$ The University of Tokyo, Tokyo, Japan<br>${ }^{2}$ TU Darmstadt, Darmstadt, Germany<br>${ }^{3}$ KAIST, Daejeon, The Republic of Korea<br>m@zie.de


#### Abstract

Pour-El and Richards [PER89], Weihrauch [Weih00], and others have extended Recursive Analysis from real numbers and continuous functions to rather general topological spaces. This has enabled and spurred a series of rigorous investigations on the computability of partial differential equations in appropriate advanced spaces of functions. In order to quantitatively refine such qualitative results with respect to computational efficiency we devise, explore, and compare natural encodings (representations) of compact metric spaces: both as infinite binary sequences (TTE) and more generally as families of Boolean functions via oracle access as introduced by Kawamura and Cook ([KaCo10], Sect. 3.4). Our guide is relativization: Permitting arbitrary oracles on continuous universes reduces computability to topology and computational complexity to metric entropy in the sense of Kolmogorov. This yields a criterion and generic construction of optimal representations in particular of (subsets of) $L^{p}$ and Sobolev spaces that solutions of partial differential equations naturally live in.


## 1 Introduction and Motivation

The Type-2 Theory of Effectivity (TTE) compares and studies transformation properties of so-called representations for a given space $X$ : surjective partial mappings $\delta: \subseteq\{0,1\}^{\omega} \rightarrow X$ describing an encoding of $X$ 's elements as infinite binary strings, such as sequences of (indices of) fast converging approximations from a fixed countable dense subset. In particular several natural but different representations of spaces of continuous functions on Euclidean domains have been established as computably equivalent. Partial differential equations, however, exhibit counter-intuitive computability properties when considered on such classical function spaces rather than than the advanced ones suggested by functional analysis: $L^{p}$ and more generally Sobolev spaces $W_{p}^{k}$ [WeZh02]. The qualitative computability theory of such spaces is well established [SZZ15]; and we,

[^16]taking a refined complexity-theoretic perspective, suggest, and justify the choice of, natural representations promising to bridge the gap to numerical practice.

Section 2 recalls notions and qualitative topological characterizations of relatively computable functions on metric spaces. Section 3 collects quantitatively refined notions under time and space bounds. Section 4 reports on Kolmogorov's entropy of a compact metric space. And Sect. 5 connects the latter two in terms of 'ordinary' and second-order representations, the latter introduced in [KaCo10, Sect. 3.4] and recalled in Sect. 6. Justified by these considerations, Sect. 7 finally introduces a natural second-order representation for Sobolev spaces. Proofs are deliberately omitted from this expository abstract.

## 2 Computing on Separable Metric Spaces

Similarly to the classical theory of computing encoding discrete structures (graphs, integers etc.) as finite binary strings, the Type-2 Theory of Effectivity (TTE) studies, and compares notions of, computation over continuous universes by encoding as infinite binary strings. The following concepts are essentially from [Weih00, Sect. 2.1+Sect. 2.3+Sect. 3.1+Sect. 8.1], Item f) from [Schr95]; cmp. also [PER89, Sect. 2].

Definition 1.(a) An Oracle Type-2 Machine $\mathcal{M}^{O}$ is a Turing machine with read-only input tape, read-write working tape, and one-way output tape as well as access to the - possibly empty - oracle $O \subseteq\{0,1\}^{*}$ by means of one-way query tape. $\mathcal{M}^{O}$ is said to compute the partial function $F: \subseteq$ $\{0,1\}^{\omega} \rightarrow\{0,1\}^{\omega}$ if, on input $\bar{w} \in \operatorname{dom}(F)$, it prints $F(\bar{w})$. Its behaviour on $\bar{w} \notin \operatorname{dom}(F)$ may be arbitrary.
(b) A representation of a space $X$ is a partial surjective mapping $\xi: \subseteq\{0,1\}^{\omega} \rightarrow$ $X . A \bar{w}$ with $\xi(\bar{w})=x$ is a $\xi$-name of $x \in X$.
(c) A partial multivalued mapping $f: \subseteq X \rightrightarrows Y$ is a relation $f \subseteq X \times Y$, considered as total function $f: X \ni x \mapsto\{y \in Y:(x, y) \in f\}$. Its domain is $\operatorname{dom}(f)=\{x: f(x) \neq \emptyset\}$. A (partial) single-valued mapping is considered as multivalued with singleton (or empty) values.
(d) For $v$ a representation of $Y, a(\xi, v)$-realizer of $f$ is a partial function $F: \subseteq$ $\{0,1\}^{\omega} \rightarrow\{0,1\}^{\omega}$ with $v \circ F \subseteq f \circ \xi$. Call $f: \subseteq X \rightrightarrows Y$ relativized $(\xi, v)-$ computable iff there exists an oracle type-2 machine $\mathcal{M}^{O}$ computing some $(\xi, v)$-realizer of $f$. We omit $\xi=\mathrm{id}$ in case $X=\{0,1\}^{\omega}$.
(e) A presented separable metric space is a triple $(X, d, \xi)$, where $X$ denotes the carrier set with metric $d: X \times X \rightarrow[0 ; \infty)$ and $\xi: \subseteq \mathbb{N} \rightarrow X$ a partial enumeration of some dense image $(\xi) \subseteq X$.
(f) For a presented separable metric space $(X, d, \xi)$, a $\xi$-name of $x \in X$ is an integer sequence $\left(a_{m}\right)_{m}$ satisfying

$$
\begin{align*}
& \forall m: \quad a_{m} \in \operatorname{dom}(\xi) \wedge d\left(\xi\left(a_{m}\right), x\right)<2^{-m} \wedge \\
& \wedge \forall a^{\prime}<a_{m}: d\left(\xi\left(a^{\prime}\right), x\right) \geq 2^{-m-1} \tag{1}
\end{align*}
$$

The induced representation of $(X, d, \xi)$ is the partial mapping (abusing names also denoted by) $\xi: \subseteq\{0,1\}^{\omega} \rightarrow X$ with $\left\langle\left(\operatorname{bin}\left(a_{m}\right)\right)_{m}\right\rangle \mapsto x$ for every $\bar{a}=\left(a_{m}\right)_{m}$ satisfying Eq. (1).
(g) Here we denote by bin both the binary expansion

$$
\text { bin : }\{0,1\}^{*} \ni\left(v_{0}, \ldots, v_{J-1}\right) \mapsto 2^{J}-1+\sum_{j=0}^{J-1} v_{j} 2^{j} \in \mathbb{N}
$$

and its inverse, where $\mathbb{N}=\{0,1,2, \ldots\}$. Furthermore write

$$
\left\langle\left(v_{1}, \ldots, v_{n}\right)\right\rangle:=\left(1, v_{1}, 1, v_{2}, \ldots, 1, v_{n-1}, 0, v_{n}\right)
$$

for the binary string encoding with delimiter; and also for pairing functions

$$
\left(\{0,1\}^{*}\right)^{*} \ni\left(\vec{v}^{(1)}, \ldots, \vec{v}^{(k)}\right) \mapsto\left\langle\vec{v}^{(1)}\right\rangle \ldots\left\langle\vec{v}^{(k)}\right\rangle \in\{0,1\}^{*}
$$

and $\left(\{0,1\}^{*}\right)^{*} \times\{0,1\}^{\omega} \rightarrow\{0,1\}^{\omega}$ and $\left(\{0,1\}^{*}\right)^{\omega} \rightarrow\{0,1\}^{\omega}$. Finally abbreviate $[N]:=\{0, \ldots, N-1\}$ for $N \in \mathbb{N}$; let $\vec{v}_{<n}$ and $\bar{v}_{<n}$ mean the first $n$ symbols of $\vec{v} \in\{0,1\}^{n+m}$ and of $\bar{v} \in\{0,1\}^{\omega}$, respectively; write $\vec{v}_{n}$ and $\bar{v}_{n}$ for the $n$-th symbol.
(h) For metric spaces $(X, d)$ and $(Y, e)$ a mapping $\mu: \mathbb{N} \rightarrow \mathbb{N}$ is a modulus of continuity to the function $f: X \rightarrow Y$ if, for every $m \in \mathbb{N}$ and $x, x^{\prime} \in X$, $d\left(x, x^{\prime}\right)<2^{-\mu(m)}$ implies e $\left(f(x), f\left(x^{\prime}\right)\right)<2^{-m}$. We write $\mathrm{B}(x, r):=\left\{x^{\prime} \in\right.$ $\left.X: d\left(x, x^{\prime}\right)<r\right\}$ for the open ball of radius $r \geq 0$ around $x \in X$ and $\overline{\mathrm{B}}(x, r):=\left\{x^{\prime} \in X: d\left(x, x^{\prime}\right) \leq r\right\}$ for the corresponding closed ball.

Our prototype presented metric space is the real unit interval $X=[0 ; 1]$ equipped with $\rho: \mathbb{N} \rightarrow X$ enumerating the dyadic rationals $\left\{0,(2 \tilde{a}+1) / 2^{m}: \mathbb{N} \ni\right.$ $\left.\tilde{a} \leq 2^{m-1}, m \in \mathbb{N}_{+}\right\}$in $[0 ; 1)$ without repetition in 'lexicographical' order: $0, \frac{1}{2}, \frac{1}{4}, \frac{3}{4}, \frac{1}{8}, \frac{3}{8}, \frac{5}{8}, \frac{7}{8}, \frac{1}{16}, \ldots ;$ cmp. [BrCo06]. Computing on continuous universes combines recursion-theoretic and topological aspects, reducing to the latter when permitting access to arbitrary oracles; cmp. Items ( $\mathrm{b}+\mathrm{c}$ ) of the following

Fact 2.(a) A function $f: X \rightarrow Y$ admits a modulus of continuity iff it is uniformly continuous. On bounded $X, f$ is Hölder continuous iff it admits a linear modulus of continuity.
(b) A partial function $F: \subseteq\{0,1\}^{\omega} \rightarrow\{0,1\}^{\omega}$ is continuous iff it is computable by some oracle type-2 machine.
(c) Let $(X, d, \xi)$ and $(Y, d, v)$ denote presented metric spaces. A partial function $f: \subseteq X \rightarrow Y$ is continuous iff it is relativized ( $\xi, v$ )-computable.
(d) Reciprocals $(0 ; 1] \ni x \mapsto 1 / x$ are ( $\rho, \rho$ )-computable but, lacking uniform continuity, not within bounded time nor space.
(e) A partial function $f: \subseteq[0 ; 1] \rightarrow \mathbb{R}$ admits a polynomial modulus of continuity iff it is $(\rho, \rho)$-computable by some polynomial-time oracle type-2 machine.
(f) The continuous function $h_{\exp }:[0 ; 1] \ni x \mapsto 1 / \ln (e / x)$ is computable (without oracle) in exponential time but, lacking a polynomial modulus of continuity, not (even with oracle) in sub-exponential time.
(g) A partial $F: \subseteq\{0,1\}^{\omega} \rightarrow\{0,1\}^{\omega}$ admits a polynomial modulus of continuity iff it is computable by some polynomial-time oracle type-2 machine.
(h) There is no representation $\delta: \subseteq\{0,1\}^{\omega} \rightarrow \operatorname{Lip}_{1}([0 ; 1],[0 ; 1])$ of the compact space of uniformly bounded and equicontinuous functions $f:[0 ; 1] \rightarrow[0 ; 1]$ with $\left|f(x)-f\left(x^{\prime}\right)\right| \leq\left|x-x^{\prime}\right|$ rendering application $(f, x) \mapsto f(x)$ uniformly $(\delta \times \rho, \rho)$-computable in relativized subexponential time.

Item (a) is from [KSZ14, Example 2.5], for (b) see [Weih00, Theorems 2.3.7+2.3.8], and for (c) confer [Weih00, Theorem 3.2.11+Definition 3.1.3]. The latter has been generalized from metric to topological so-called QCB-spaces [Schr06, Theorem 2], to weaker representations, as well as from continuity to (levels of Borel) measurability [Zieg07, dBYa10]. For (d) see for instance [Weih00, Theorem 4.3.2.6+Example 7.2.8.3]. [Ko91, Theorem 2.19] asserts one direction of (e) for total functions $f:[a ; b] \rightarrow \mathbb{R}$. Regarding (f) consider [KMRZ15, Fact 3 g$]$; and Lemma 6.3 in [PaZi13] for (g). Claim (h) is contained in [Weih03, Sect. 6]; see also [FHHP15, Theorem 3.1].

## 3 Computational Complexity on Compact Metric Spaces

Items (d) to (h) of Fact 2 refer to the following notions:
Definition 3.(a) For $t: \mathbb{N} \rightarrow \mathbb{N}$, an oracle type-2 machine $\mathcal{M}^{O}$ computing $F: \subseteq\{0,1\}^{\omega} \rightarrow\{0,1\}^{\omega}$ does so in time $t(m)$ if it prints the $m$-th symbol of $F(\bar{v})$ after at most $t(m)$ steps for every $\bar{v} \in \operatorname{dom}(F)$. $F$ is relativized polynomial-time computable if there exists some $d \in \mathbb{N}$ and an oracle type-2 machine computing it in time $t(m)=d \cdot\left(1+m^{d}\right)$.
(b) For $s: \mathbb{N} \rightarrow \mathbb{N}$, an oracle type-2 machine $\mathcal{M}^{O}$ computing $F: \subseteq\{0,1\}^{\omega} \rightarrow$ $\{0,1\}^{\omega}$ does so in space $s(n)$ if it prints the $m$-th symbol of $F(\bar{v})$ after using at most $s(m)$ cells of the working tape (and 'arbitrary' amounts of the input, output, and query tapes) for every $\bar{v} \in \operatorname{dom}(F)$.
(c) Fix $s, t: \mathbb{N} \rightarrow \mathbb{N}$ and a (possibly partial and multivalued) function $f: \subseteq X \rightrightarrows$ $Y$ between represented space $(X, \xi)$ and presented metric space $(Y, e, v)$. An oracle type-2 machine $(\xi, v)$-computes $f$ in time $t(m)$ and space $s(m)$ iff it, for every input of any $\bar{v} \in \operatorname{dom}(\xi)$ with $\xi(\bar{v}) \in \operatorname{dom}(f)$, produces an $v$-name $\left\langle\left(\operatorname{bin}\left(w_{m}\right)\right)_{m}\right\rangle$ of some $y \in f(x)$ such that $\operatorname{bin}\left(w_{m}\right) \in\{0,1\}^{*}$ appears on the output tape within $\leq t(m)$ steps and using $\leq s(m)$ cells of the working tape.
(d) The time and/or space of a machine according to (c) is bounded if there exist mappings $t$ and/or s as above. It is logarithmic/polynomial/exponential if such mappings can be chosen to have asymptotic growth bounded by $\mathcal{O}(\log m), \operatorname{poly}(m):=\mathcal{O}(m)^{\mathcal{O}(1)}$, and $2^{\text {poly }(m)}$, respectively.
(e) A representation $\xi$ of $X$ is polynomially admissible if for every representation $\delta$ of $X$ the following holds: $\delta: \subseteq\{0,1\}^{\omega} \rightarrow X$ has a polynomial modulus of continuity iff there exists a mapping $F: \operatorname{dom}(\delta) \rightarrow \operatorname{dom}(\xi) \subseteq\{0,1\}^{\omega}$ with polynomial modulus of continuity such that $\delta=\xi \circ F$.
(f) The product $\xi \times v$ of $\xi: \subseteq\{0,1\}^{\omega} \rightarrow X$ and $v: \subseteq\{0,1\}^{\omega} \rightarrow Y$ is the mapping $\{0,1\}^{\omega} \ni\left(w_{0}, w_{1}, w_{2}, w_{3}, \ldots\right) \mapsto\left(\xi\left(w_{0}, w_{2}, \ldots\right), v\left(w_{1}, w_{3}, \ldots\right)\right) \in X \times Y$.
(g) A mapping $f: X \rightarrow Y$ between topological spaces is proper if the pre-images $f^{-1}[K]=\{x \in X: f(x) \in K\} \subseteq X$ of compact sets $K \subseteq Y$ are compact.

Item (e) refines the well-known qualitative condition of computable admissibility [Schr06]; see [Weih00, Theorem 3.2.9]. For $X=\{0,1\}^{\omega}=Y$ condition (c) boils down to (a) and (b), but for other represented spaces it may be unrelated to that of computing a $(\xi, v)$-realizer within the given resource bounds [Weih00, Examples 7.2.1+7.2.3]: $\xi$ and $v$ could require/admit very long/short initial segments of names before reaching precision $2^{-m}$. Moreover, said precision is to be met within the given resource bound, regardless of the argument. To avoid counter-examples like Fact 2(d) we focus on proper representations of compact spaces; compare [Schr95, Weih03, Schr04] and [Weih00, Exercise 7.1.2].

## 4 Metric Entropy of Compact Metric Spaces

Theorem 6 will generalize Items (e) and (g) in Fact 2, and the complexitytheoretic characterizations of Items (f) and (h) from Cantor space and the real unit interval to certain compact metric spaces in the spirit of Fact 2(c), based on the following notions essentially dating back to Andrey N. Kolmogorov:

Definition 4. Fix a bounded metric space $(X, d)$.
(a) For $\varepsilon>0$ let $\mathcal{C}(X, d, \varepsilon):=\sup \left\{\operatorname{Card}(C) \mid C \subseteq X, \forall x, x^{\prime} \in C: x=\right.$ $\left.x^{\prime} \vee d\left(x, x^{\prime}\right) \geq \varepsilon\right\}$ denote the size of a largest collection of points fitting into $X$ while avoiding each other by at least distance $\varepsilon$.
(b) For $\varepsilon>0$ let $\mathcal{H}(X, d, \varepsilon):=\inf \{\operatorname{Card}(C) \mid C \subseteq X, \forall x \in X \exists c \in C$ : $d(x, c)<\varepsilon\}$ denote the least number of open balls of radius $\varepsilon$ covering $X$.
(c) The capacity $\lceil(X, d)\rceil: \mathbb{N} \rightarrow \mathbb{N}$ of $(X, d)$ is the truncated binary logarithm of $n \mapsto \mathcal{C}\left(X, d, 2^{-n}\right)$; i.e. $X$ admits $2^{\lceil X\rceil(n)}$, but not $2^{\lceil X\rceil(n)+1}$, points of pairwise distance $\geq 2^{-n}$.
(d) Dually, the entropy $\lfloor(X, d)\rfloor: \mathbb{N} \rightarrow \mathbb{N}$ of $(X, d)$ is the truncated binary logarithm of $n \mapsto \mathcal{H}\left(X, d, 2^{-n}\right)$; i.e. $X$ can be covered by $2^{\lfloor X\rfloor(n)}$ open balls of radius $2^{-n}$, but not by $2^{\lfloor X\rfloor(n)-1}$.

Compare for instance [KoTi59] or [Weih03, Sect. 6] and the related notion of a modulus of total boundedness [Kohl08, Definition 17.106]. Lemma 5(a) asserts that $\lceil X\rceil$ and $\lfloor X\rfloor$ have equal asymptotic growth as long as either one is at most exponential, i.e. $\leq 2^{\text {poly (n) }}$ : such as, e.g., $C_{\mu}(Y,[0 ; 1])$ for both $\mu$ and $\lceil Y\rceil$ polynomials according to Item d) of the following

Lemma 5.(a) Suppose $C \subseteq X$ is maximal w.r.t. $\subseteq$ satisfying $\forall x, x^{\prime} \in C: x=$ $x^{\prime} \vee d\left(x, x^{\prime}\right) \geq \varepsilon$. Then $\bigcup_{c \in C} \mathrm{~B}(c, \varepsilon)=X$.
For $(X, d)$ totally bounded, $\mathcal{C}(X, d, \cdot)$ and $\mathcal{H}(X, d, \cdot)$ are non-increasing total functions $(0 ; \infty) \rightarrow \mathbb{N}$ satisfying $\mathcal{H}(X, d, \varepsilon) \leq \mathcal{C}(X, d, \varepsilon) \leq \mathcal{H}(X, d, \varepsilon / 2)$. In particular it holds $\lfloor(X, d)\rfloor(n) \leq\lceil(X, d)\rceil(n) \leq\lfloor(X, d)\rfloor(n+1)$.
(b) The finite set $X:=\left\{1,2, \ldots, 2^{k}\right\}$ of integers has constant capacity and entropy $\lceil X\rceil(n) \equiv k \equiv\lfloor X\rfloor(n)$. The Euclidean cube/torus $\left[0 ; 2^{k}\right)^{d}$, equipped with the maximum norm, has capacity $\left\lceil\left[0 ; 2^{k}\right)^{d}\right\rceil(n)=(n+k) \cdot d=$ $\left\lfloor\left[0 ; 2^{k}\right)^{d}\right\rfloor(n)$ and thus polynomial entropy.
(c) The compact space from Fact 2(h) equipped with the supremum norm has asymptotically exponential capacity and entropy: $\left\lceil\operatorname{Lip}_{1}([0 ; 1],[0 ; 1])\right\rceil(n)=$ $\Theta\left(2^{n}\right)=\left\lfloor\operatorname{Lip}_{1}([0 ; 1],[0 ; 1])\right\rfloor(n)$. The same holds for $\operatorname{Lip}_{1}([0 ; 1],[0 ; 1]) \subseteq L^{p}$ equipped with the norm $f \mapsto\|f\|_{p}:=\sqrt[p]{\int_{0}^{1}|f(t)|^{p} d t}$ for any fixed $p \geq 1$.
(d) Suppose totally bounded $(X, d)$ has diameter $\operatorname{diam}(X):=\sup \left\{d\left(x, x^{\prime}\right)\right.$ : $\left.x, x^{\prime} \in X\right\} \leq 1$ and super-logarithmic yet at most exponential entropy $\lfloor X\rfloor$ : $\mathbb{N} \rightarrow \mathbb{N}$. Moreover fix some strictly increasing $\mu: \mathbb{N} \rightarrow \mathbb{N}$. W.r.t. sup-norm the space $C_{\mu}(X,[0 ; 1]):=\{f: X \rightarrow[0 ; 1]$ has modulus of continuity $\mu\}$ has $\log \left\lfloor C_{\mu}(X,[0 ; 1])\right\rfloor(n)=\Theta(\lfloor X\rfloor(\mu(n \pm \Theta(1))))$. A set $Y \subseteq C(X,[0 ; 1])$ is relatively compact iff it belongs to $C_{\mu}(X,[0 ; 1])$ for some $\mu$.
(e) Cantor space $2^{\omega}=\{0,1\}^{\omega}$, equipped with the metric $\beta(\bar{v}, \bar{w}):=$ $2^{-\min \left\{n: v_{n} \neq w_{n}\right\}}$ has linear capacity $\left\lceil\left(2^{\omega}, \beta\right)\right\rceil(n)=n+1$; equipped with the topologically equivalent metric $\beta^{\prime}(\bar{v}, \bar{w}):=1 /\left(1+\min \left\{n: v_{n} \neq v_{m}\right\}\right)$ on the other hand it has exponential capacity $\left\lceil\left(2^{\omega}, \beta^{\prime}\right)\right\rceil(n)=2^{n}-1$.
(f) However whenever $d$ and $d^{\prime}$ are strongly equivalent metrics on $X$ in the sense that $d^{\prime} \cdot 2^{-c} \leq d \leq d^{\prime} \cdot 2^{c}$ holds for some $c \in \mathbb{N}$ (such as in case $X$ lives in some finite-dimensional normed real vector space), their induced capacities and entropies differ by at most a constant shift, i.e., it holds $\quad \forall n \geq c$ :

$$
\begin{aligned}
& \left\lceil\left(X, d^{\prime}\right)\right\rceil(n-c) \leq\lceil(X, d)\rceil(n) \leq\left\lceil\left(X, d^{\prime}\right)\right\rceil(n+c) \\
& \left\lfloor\left(X, d^{\prime}\right)\right\rfloor(n-c) \leq\lfloor(X, d)\rfloor(n) \leq\left\lfloor\left(X, d^{\prime}\right)\right\rfloor(n+c)
\end{aligned}
$$

(g) Let $(X, d)$ and $(Y, e)$ be compact metric spaces and $f: X \rightarrow Y$ have modulus of continuity $\mu$. Then the image $f[X] \subseteq Y$ has entropy $\lfloor F[X]\rfloor \leq\lfloor X\rfloor \circ \mu$.

Item (d) quantitatively refines the classical Arzelá-Ascoli Theorem; cmp. [Weih03, Theorem 6.7.3].

## 5 Relativized Complexity and Entropy

The entropy/capacity of compact metric spaces essentially determines the relativized computational complexity of functions on them:

Theorem 6. For a compact metric space $(X, d)$ the following are equivalent:
(i) $X$ has polynomially bounded entropy: $\lfloor X\rfloor(m) \leq p(m)$ for some $p \in \mathbb{N}[m]$.
(ii) $X$ has a proper representation $\delta: \subseteq\{0,1\}^{\omega} \rightarrow X$ with polynomial modulus of continuity.
(iii) $X$ admits a representation $\delta$ rendering the following parameterized partial/fuzzy/ soft equality test relativized $\delta \times \delta$-computable in time polynomial in $m$ :

$$
\begin{equation*}
X \times X \times \mathbb{N} \ni(x, y, m) \mapsto 1^{\omega} \text { for } x=y, \quad \mapsto 0^{\omega} \text { for } d(x, y) \geq 2^{-m} \tag{2}
\end{equation*}
$$

(iv) There exists a representation $\delta$ of $X$ rendering Equation (2) relativized $\delta \times \delta$-computable in space logarithmic in $m$.
(v) There exists a representation $\delta$ of $X$ rendering the metric $d: X \times X \rightarrow$ $[0 ; \infty)$ relativized $(\delta \times \delta, \rho)$-computable in polynomial time
(vi) or in logarithmic space.
(vii) There exists a representation $\delta$ of $X$ and polynomial $q \in \mathbb{N}[m]$ such that every 1-Lipschitz function $f: X \rightarrow[0 ; 1]$ is relativized $(\delta, \rho)$-computable in time $q(m)$
(viii) or space $\mathcal{O}(\log q(m))=\mathcal{O}(\log m)$.

Perhaps surprisingly, the same holds with $\delta$ replaced, in Items (iii) to (viii), by a second-order representation $\Delta$ of $X$ in the following sense:

## 6 Second-Order Complexity Theory

According to Fact 2(h) compact space $\operatorname{Lip}_{1}([0 ; 1],[0 ; 1])$ does not admit a complexity-wise reasonable representation, i.e., encoding as infinite binary strings $\{0,1\}^{\omega} \cong 2^{\{1\}^{*}}$ : essentially due to their restriction to sequential access which requires 'skipping' over the $f$ 's (approximate) values at many arguments $f\left(x^{\prime}\right)$ before reaching the desired $f(x)$; whereas function arguments in practice provide oracle-like random access to their values. This has been formalized by encoding real function arguments as oracles [KaCo10].

Remark 7 Classical oracles are decision problems $O \subseteq\{0,1\}^{*}$, that is, they return a single bit. Function oracles on the other hand return finite strings, that is, they correspond to elements of Baire space $\mathbb{N}^{\mathbb{N}}$ encoded in binary as mappings $\varphi:\{0,1\}^{*} \rightarrow\{0,1\}^{*}$. Now if the answer $\vec{w}=\varphi(\vec{v})$ to a query $\vec{v}$ is 'long', an oracle machine $\mathcal{M}^{\varphi}$ arguably should be allotted more time than the same when run with an oracle $\psi$ giving 'short' answers. This leads to secondorder polynomial resource bounds; see [KaCo96]. Function oracles of polynomial length, on the other hand, can be encoded into decision oracles queried bitwise (and in particular satisfying effective polynomial boundedness [KaPa15]).

For the purpose of this work we focus on the latter:
Definition 8.(a) A second-order representation of a space $X$ is a partial surjective mapping $\Xi: \subseteq 2^{\{0,1\}^{*}} \rightarrow X$, where $2^{Y}$ denotes the set of all subsets $O \subseteq Y$, each identified with its characteristic function $\mathbf{1}_{O}: Y \rightarrow\{0,1\}$.
(b) An oracle Type-2 machine with variable/generic oracle is called contingent and denoted $\mathcal{M}^{?}$. It computes a partial function $F: \subseteq 2^{\{0,1\}^{*}} \times\{0,1\}^{\omega} \rightarrow$ $\{0,1\}^{\omega}$ if, for every $(O, \bar{v}) \in \operatorname{dom}(F), \mathcal{M}^{O}$ on input $\bar{v}$ prints $F(O, \bar{v})$. It does so in logarithmic/polynomials/exponential time/space if the $n$-th symbol of $F(O, \bar{v})$ appears within such resource bounds of time/work tape cells, independently of $(O, \bar{v}) \in \operatorname{dom}(F)$ while permitting unbounded use of the input, output, and query tapes. More precisely $\mathcal{M}^{?}$ may peruse a fixed-depth stack of write-only query tapes where an oracle call refers to, and purges, the top one.
(c) Fix a second-order represented space $(X, \Xi)$, (first-order) represented space $(Y, v)$, and presented metric space $(Z, e, \zeta)$ with induced representation $\zeta$. A contingent oracle machine $\mathcal{M}^{?}(\Xi, v, \zeta)$-computes $f: \subseteq X \times Y \rightrightarrows Z$ in time $t(m)$ and space $s(m)$ iff, for every $O \in \operatorname{dom}(\Xi)$ and $\bar{v} \in \operatorname{dom}(v)$ with $(\Xi(O), v(\bar{v})) \in \operatorname{dom}(f), \mathcal{M}^{O}$ on input $\bar{v}$ produces a $\zeta$-name $\left\langle\left(\operatorname{bin}\left(w_{m}\right)\right)_{m}\right\rangle$ of some $z \in f(\Xi(O), v(\bar{v}))$ such that $\operatorname{bin}\left(w_{m}\right) \in\{0,1\}^{*}$ appears on the output tape within at most $t(m)$ steps and using at most $s(m)$ cells of the working tape (and again 'arbitrary' amounts of the input, output, and query tapes).
(d) For second-order representations $\Xi: \subseteq 2^{\{0,1\}^{*}} \rightarrow X$ and $\Upsilon: \subseteq 2^{\{0,1\}^{*}} \rightarrow Y$, their (binary) product $\Xi \times \Upsilon$ is the mapping

$$
2^{\{0,1\}^{*}} \ni O \mapsto(\Xi(\{\vec{v}: 0 \vec{v} \in O\}), \Upsilon(\{\vec{w}: 1 \vec{v} \in O\})) \in X \times Y
$$

(e) Following up on Definition 1f), the second-order representation induced by a presented metric space $(X, d, \xi)$ is the mapping

$$
\Xi: \subseteq 2^{\{0,1\}^{*}} \ni\left\{\left\langle\operatorname{bin}\left(2^{m}\right), \operatorname{bin}\left(2^{j}\right)\right\rangle: \operatorname{bin}\left(a_{m}\right)_{j}=1\right\} \mapsto x \in X
$$

for every $\xi$-name $\bar{a}=\left(a_{m}\right)_{m} \in \mathbb{N}^{\omega}$ of $x$.
(f) Fix presented metric spaces $(X, d, \xi)$ and $(Y, e, v)$ with induced (first-order) representations $\xi$ and $v$. Justified by Fact 2(c), equip (any fixed compact subset $\mathcal{Z}$ of) the space $C(X, Y)$ of continuous total functions $f: X \rightarrow Y$ with the following second-order representation $v^{\xi}$ : Let

$$
\begin{aligned}
O=\left\{\left\langle\operatorname{bin}(a), \operatorname{bin}\left(2^{m}\right), \operatorname{bin}\left(2^{j}\right)\right\rangle: a \in\right. & \operatorname{dom}(\xi), \\
& \left.\langle\operatorname{bin}(\varphi(a, m))\rangle_{j}=1\right\} \subseteq\{0,1\}^{*}
\end{aligned}
$$

be an $v^{\xi}$-name of $f \in C(X, Y)$ for every mapping $\varphi: \operatorname{dom}(\xi) \times \mathbb{N} \subseteq \mathbb{N} \times \mathbb{N} \rightarrow$ $\operatorname{dom}(v) \subseteq \mathbb{N}$ where $(\varphi(a, m))_{m}$ is an v-name of $f(\xi(a))$.

So Item (c) is about functions with ordinary/first-order represented co-domain, (g) with second-order ones. And $v^{\xi}$ according to Item (f) encodes (approximations in terms of the dense sequence in $Y$ given by $v$ to) the values of $f$ on the dense sequence in $X$ given by $\xi$; cmp. [KaPa15, top of p. 8]. The fixed-depth stacks and subtle semantics of query tapes have been well justified in the discrete setting [Wils88, Buss88, ACN07] as well as in computational analysis [KaOt14]. Definition $8(\mathrm{f})$ does not (yet) incorporate quantitative information about continuity of $f$. In the case $\mathcal{Z}=\operatorname{Lip}_{1}([0 ; 1],[0 ; 1])$ the representation here called $\rho^{\rho}$ is (equivalent to one) well-known [KaCo10,KORZ12,FHHP15,FeZi15]; and renders application $(f, x) \mapsto f(x)$ computable in polynomial-time.

## 7 Representing $L^{p}$ and Sobolev Spaces

Recall that, for compact $X \subseteq \mathbb{R}^{d}, L^{p}(X)=W^{0, p}(X)$ consists of all measurable (but not necessarily continuous) functions $f: X \rightarrow \mathbb{R}$ such that
$\|f\|_{p}<\infty$; and, more generally, $W^{k, p}(X)$ of all $f$ whose weak partial derivatives $\partial^{\vec{j}} f:=\partial x_{1}^{j_{1}} \cdots \partial x_{d}^{j_{d}} f$ up to order $\| \vec{j} \mid:=j_{1}+\cdots+j_{d} \leq k$ belong to $L^{p}(X)$, equipped with the norm $\|f\|_{k, p}:=\max _{|\vec{j}| \leq k}\left\|\partial^{\vec{j}} f\right\|_{p}$. By Lemma 5(c) and the extension of Theorem 6, no (first or) second-order representation of $\operatorname{Lip}_{1}([0 ; 1],[0 ; 1]) \subseteq L_{p}[0 ; 1]$ can simultaneously render both $(f, g) \mapsto|f-g|$ and $f \mapsto\|f\|_{p}$ polynomial-time computable.

Definition 9.(a) Inspired by Definition 1 (h) call $\mu: \mathbb{N} \rightarrow \mathbb{N}$ an $L^{p}$-modulus of $f \in L^{p}[0 ; 1]$ if $\sqrt[p]{\int_{0}^{1}|f(t+h)-f(t)|^{p} d t}<2^{-m}$ whenever $|h|<2^{-\mu(m)}$, with the convention $f(t) \equiv 0$ for $t \notin[0 ; 1]$.
(b) Abbreviate $W_{\mu}^{k, p}[0 ; 1]:=\left\{f \in W^{k, p}[0 ; 1]: \partial^{k} f\right.$ has $L^{p}$-modulus $\left.\mu\right\}$.
(c) Let $\Xi$ denote the second-order representation of $L^{1}[0 ; 1] \supseteq W^{k, p}[0 ; 1]$ s.t. a $\Xi$-name of $f$ is a $\rho^{\rho}$-name of the continuous $[0 ; 1] \ni s \mapsto \int_{0}^{s} f(t) d t \in \mathbb{R}$.

By Fréchet-Kolmogorov, $Y \subseteq L^{p}([0 ; 1])$ is relatively compact iff there exists some $\mu$ with $Y \subseteq W_{\mu}^{0, p}[0 ; 1]$ : Our convention of extending $f$ with zero asserts $W_{\mu}^{0, p}$ to be bounded by $2^{\mu(0)}$. Although harder than Lemma 5 d ), we can prove

Theorem 10. Fix polynomial-time computable $p \geq 1$ and strictly increasing $\mu$.
(a) $\log \left\lfloor W_{\mu}^{0, p}([0 ; 1])\right\rfloor(n)=\mu(n \pm \Theta(1))$.
(b) For any fixed polynomial $\mu$, the embedding $W_{\mu}^{1, p}[0 ; 1] \hookrightarrow C_{n \mapsto \mu(n+1)}[0 ; 1]$ is well-defined and $\left(\Xi, \rho^{\rho}\right)$-computable in polynomial time.
(c) For any fixed $k \in \mathbb{N}$ and polynomial $\mu$, differentiation $\partial: W_{\mu}^{k+1, p}[0 ; 1] \rightarrow$ $W_{\mu}^{k, p}[0 ; 1]$ is well-defined and $(\Xi, \Xi)$-computable in polynomial time.
(d) For any fixed $k \in \mathbb{N}$ and polynomial $\mu$, the embedding $W_{\mu}^{k+1, p}[0 ; 1] \hookrightarrow$ $W^{k, p}[0 ; 1]$ is $(\Xi, \Xi)$-computable in polynomial time.

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# Ergodicity of Noisy Cellular Automata: The Coupling Method and Beyond 

Irène Marcovici ${ }^{(\boxtimes)}$<br>Institut Élie Cartan de Lorraine, Université de Lorraine, Nancy, France<br>irene.marcovici@univ-lorraine.fr


#### Abstract

When perturbating a cellular automaton by a random noise (positive probability of error, for each cell independently), the system is generally expected to be ergodic, meaning that during its evolution, it eventually forgets about its initial condition. For a high noise, this can be shown by coupling. However, for a small noise, ergodicity is often very difficult to prove. We present extensions of the coupling method to small noises when the cellular automaton has some specific properties (hardcore exclusion, nilpotency, permutivity).


Consider a set of cells indexed by $\mathbb{Z}$, each cell containing a letter from a finite symbol set $S$. A cellular automaton (CA) is a dynamical system which acts locally and synchronously on the configuration space $S^{\mathbb{Z}}$. When the updates are random, we obtain a probabilistic cellular automaton (PCA): at each time step, the new content of each cell is randomly chosen, independently of the others, according to a distribution given by the states in a finite neighbourhood of the cell. Examples of PCA are given by noisy CA: the updates are governed by a deterministic rule, which is perturbated by errors with a positive probability.

A PCA is said to be ergodic if it forgets its initial condition, meaning that it has a unique and attractive invariant measure. A variety of tools have been developed to study the ergodicity of PCA. But most of them only allow to handle PCA for which the transition probability to any state given any neighbourhood states is large enough. In particular, ergodicity is often very difficult to prove for noisy CA, even in cases where it appears clear from heuristics or simulations.

In Sect. 1, we recall the coupling method and the notion of envelope PCA, which gives a general framework to prove ergodicity for the high noise regime. Details and more references can be found in a joint publication with A. Bušić and J. Mairesse [1]. The next sections are devoted to three examples of families of CA for which some specific tools allow to prove the ergodicity for small noise. In Sect. 2, we present new results on the noisy version of the hardcore CA. These results stem from a joint work with J.B. Martin, motivated by the study of a percolation game, together with A.E. Holroyd [3]. Then, in Sects. 3 and 4, we consider perturbations of nilpotent and permutive CA. These sections are based on a work initiated with S. Taati and carried on together with M. Sablik. It is interesting to note that nilpotent CA and permutive CA present opposite behaviours within the rich zoology of CA. While nilpotent CA reach in
bounded time a configuration where a single symbol remains, bipermutive CA are expected to converge to the uniform distribution on $S^{\mathbb{Z}}$ from a large class of initial distributions (randomization phenomenon).

## 1 Definitions

Let $S$ be a finite set of symbols. We equip the set $S^{\mathbb{Z}}$ of configurations with the product topology. For a finite subset $K \subset \mathbb{Z}$ and an element $y \in S^{K}$, the set $\left[y_{K}\right]=\left\{z \in S^{\mathbb{Z}} ; \forall k \in K, z_{k}=y_{k}\right\}$ is called a cylinder of base $K$, and we denote by $\mathcal{C}(K)$ the set of all cylinders of base $K$. The set of probability distributions on $S^{\mathbb{Z}}$ for the Borelian $\sigma$-algebra is denoted by $\mathcal{M}\left(S^{\mathbb{Z}}\right)$. For a distribution $\mu \in$ $\mathcal{M}\left(S^{\mathbb{Z}}\right)$, we denote by $\mu\left[y_{K}\right]$ the probability of the cylinder [ $\left.y_{K}\right]$.

Let $m \geq 1$ be an integer, and let $n_{1}, \ldots, n_{m} \in \mathbb{Z}$. A local rule of neighbourhood $\mathcal{N}=\left\{n_{1}, \ldots, n_{m}\right\}$ is a function $f: S^{m} \rightarrow S$. The cellular automaton (CA) of local rule $f$ is the function $F: S^{\mathbb{Z}} \rightarrow S^{\mathbb{Z}}$ defined by:

$$
\forall x \in S^{\mathbb{Z}}, \forall k \in \mathbb{Z}, F(x)_{k}=f\left(x_{k+n_{1}}, \ldots, x_{k+n_{m}}\right)
$$

We denote by $\sigma: S^{\mathbb{Z}} \rightarrow S^{\mathbb{Z}}$ the shift map, defined by $\sigma(x)_{k}=x_{k-1}$. By Curtis-Hedlund-Lyndon theorem, CA are exactly continuous functions that commute with $\sigma$.

For probabilistic cellular automata ( PCA ), the local rule is a function $\varphi$ : $S^{m} \rightarrow \mathcal{M}(S)$, where $\mathcal{M}(S)$ denotes the set of probability distributions on $S$. From a configuration $x \in S^{\mathbb{Z}}$, cell $k$ is updated by a symbol chosen according to the distribution $\varphi\left(x_{k+n_{1}}, \ldots, x_{k+n_{m}}\right)$, independently for different cells. A PCA can be viewed as a Markov chain on $S^{\mathbb{Z}}$. The evolution of a PCA is described by a family of random variables $\left(X^{t}\right)_{t \geq 0}$, where $X^{t}$ represents the configuration at time $t$ when iterating the dynamics from the (deterministic or random) initial condition $X^{0}$. Formally, the PCA of local rule $\varphi$ can also be seen as the function

$$
\begin{aligned}
\Phi: \mathcal{M}\left(S^{\mathbb{Z}}\right) & \rightarrow \mathcal{M}\left(S^{\mathbb{Z}}\right) \\
\mu & \mapsto \Phi \mu
\end{aligned}
$$

defined on cylinder sets by:

$$
\Phi \mu\left[y_{K}\right]=\sum_{\left[x_{K+\mathcal{N}}\right] \in \mathcal{C}(K+\mathcal{N})} \mu\left[x_{K+\mathcal{N}}\right] \prod_{k \in K} \varphi\left(x_{k+n_{1}}, \ldots, x_{k+n_{m}}\right)\left(y_{k}\right),
$$

for any probability distribution $\mu \in \mathcal{M}\left(S^{\mathbb{Z}}\right)$. If the initial configuration $X^{0}$ has distribution $\mu_{0}$, then at time $t$, configuration $X^{t}$ has distribution $\Phi^{t} \mu_{0}$.

Definition 1. Let $\Phi$ be a PCA. The distribution $\pi \in \mathcal{M}\left(S^{\mathbb{Z}}\right)$ is invariant for $\Phi$ if $\Phi \pi=\Phi$. The PCA $\Phi$ is ergodic if it has a unique invariant distribution $\pi$ and if for any initial distribution $\mu_{0} \in \mathcal{M}\left(S^{\mathbb{Z}}\right),\left(\Phi^{t} \mu_{0}\right)_{t \geq 0}$ converges (weakly) to $\pi$.

In this article, we will focus on specific families of PCA obtained when adding random and independent errors in the updates of a deterministic CA.

Let $F$ be a deterministic CA, and let $\Phi$ be a PCA with same symbol set and same neighbourhood as $F$. We say that $\Phi$ is an $\varepsilon$-perturbation of $F$ if its local function is such that for all $x_{1}, \ldots, x_{m} \in S, \varphi\left(x_{1}, \ldots, x_{m}\right)\left(f\left(x_{1}, \ldots, x_{m}\right)\right) \geq 1-\varepsilon$, meaning that there is a deviation from $F$ with probability at most $\varepsilon$.

Let us consider a map $R: S \rightarrow \mathcal{M}(S)$, and let $F$ be a CA of local rule $f$. The noisy version of the CA $F$ with noise $R$ is the PCA of local rule $\varphi=R \circ f$. Starting from a configuration $x \in S^{\mathbb{Z}}$, the CA $F$ is first applied, and then, for each cell $k$ independently, the symbol at cell $k$ is replaced by a symbol distributed according to the distribution $R\left(F(x)_{k}\right)$. We denote simply by $R_{i, j}$ the probability $R(i)(j)$ for a symbol $i$ to be changed into symbol $j$. The noise is said to be positive if for all $i, j \in S, R_{i, j}>0$. The matrix $R$ is a stochastic matrix. The noise preserves the uniform distribution on $S$ if for all $j \in S, \sum_{i \in S} R_{i, j}=1$, meaning that the matrix $R$ is doubly stochastic.

We will also pay special attention to elementary $P C A$, of symbol set $S=$ $\{\mathbf{0}, \mathbf{1}\}$ and neighbourhood $\mathcal{N}=\{0,1\}$. They are characterized by four parameters: the probabilities $\theta_{i j}=\varphi(i, j)(\mathbf{1})$ to update a cell by the symbol $\mathbf{1}$ if its neighbourhood is in state $i, j$, for $i, j \in S^{2}$. If we further assume that $\theta_{\mathbf{0 1}}=\theta_{\mathbf{1 0}}$, the general tools that have been developed to prove the ergodicity allow to handle more than $90 \%$ of the volume of the cube defined by the parameter space [2]. But for example, the noisy hardcore CA we study in Sect. 3 belongs to an open domain of the cube where none of those criteria is valid.

## 2 The Coupling Method

### 2.1 The Envelope PCA

Intuitively, a PCA is ergodic if it "forgets" its initial condition. In some cases, it is possible to prove the ergodicity in a constructive way, by making evolve simultaneously the trajectories from different initial conditions, using a common source of randomness, and showing that the evolutions of all these trajectories are asymptotically the same.

The envelope PCA allows to systematize this idea of coupling. Instead of running the PCA from different initial configurations, we define a new PCA on an extended alphabet, containing a symbol ? representing sites whose values are not known (i.e. which may differ between the different copies) and we run it from a single initial configuration containing only the symbol ?. Each time we are able to make the different copies match on a cell, the symbol ? is replaced by the state $q \in S$ on which the different copies agree. An evolution of the envelope PCA thus encodes a coupling of different copies of the original PCA, with a symbol ? denoting sites where the copies disagree. If the density of symbols ? converges to 0 when time goes to infinity, it means that the PCA is ergodic.

Let us assume that $\Phi$ is a PCA defined on a binary symbol set $S=\{\mathbf{0}, \mathbf{1}\}$, and let $\tilde{S}=\{\mathbf{0}, \mathbf{1}, \boldsymbol{?}\}$. We define a partial order on $\tilde{S}$ by $\mathbf{0} \prec \boldsymbol{?} \succ \mathbf{1}$. The envelope
$P C A \tilde{\Phi}$ of $\Phi$ is the PCA of neighbourhood $\mathcal{N}$ and local function $\tilde{\varphi}: \tilde{S}^{m} \rightarrow \mathcal{M}(\tilde{S})$ defined for $q \in S$ by

$$
\tilde{\varphi}\left(y_{1}, \ldots, y_{m}\right)(q)=\min \left\{\varphi\left(x_{1}, \ldots, x_{m}\right)(q) ; x_{1} \preceq y_{1}, \ldots, x_{m} \preceq y_{m}\right\}
$$

where in the expression above, $x_{1}, \ldots, x_{m}$ are taken in $S$. The probability of a transition to the symbol ? is then given by:

$$
\tilde{\varphi}\left(y_{1}, \ldots, y_{m}\right)(\boldsymbol{?})=1-\tilde{\varphi}\left(y_{1}, \ldots, y_{m}\right)(\mathbf{0})-\tilde{\varphi}\left(y_{1}, \ldots, y_{m}\right)(\mathbf{1}) .
$$

From a configuration $y \in \tilde{S}^{\mathbb{Z}}$, cell $k$ is thus updated by the symbol $q \in S$ with the minimum of the probabilities of transition to the symbol $q$ for $\Phi$, taken over all the values of the neighbourhood of cell $k$ that are compatible with the unknown cells of $y$. With the remaining probability, the cell is updated by a ?

Proposition 1. If the density $\tilde{\Phi}^{t} \delta_{\text {? }}^{2}[$ ? $]$ of symbols ? at time $t$ starting from the initial configuration ? $?^{\mathbb{Z}}$ converges to 0 as $t \rightarrow \infty$, then the PCA $\Phi$ is ergodic.

The fact that symbols ? die out is equivalent to the ergodicity of the envelope PCA $\tilde{\Phi}$, but the ergodicity of the original PCA $\Phi$ does not imply the ergodicity of $\tilde{\Phi}$. Note also that the definition of the envelope PCA can be extended to sets of symbols having more than two elements [1].

### 2.2 Ergodicity Criterion

In this section, we still consider a binary PCA $\Phi$ and its envelope PCA $\tilde{\Phi}$. In the evolution of the envelope PCA, at each time step, a cell is updated by the symbol ? only if it has at least one neighbour in state ?, and in that case, it becomes a ? with probability at most:

$$
\begin{align*}
p_{?} & =\tilde{\varphi}(?, \ldots, ?)(\boldsymbol{?}) \\
& =1-\min _{x_{1}, \ldots, x_{m} \in S} \varphi\left(x_{1}, \ldots, x_{m}\right)(\mathbf{0})-\min _{x_{1}, \ldots, x_{m} \in S} \varphi\left(x_{1}, \ldots, x_{m}\right)(\mathbf{1})  \tag{1}\\
& =\max _{x_{1}, \ldots, x_{m} \in S} \varphi\left(x_{1}, \ldots, x_{m}\right)(\mathbf{0})-\min _{x_{1}, \ldots, x_{m} \in S} \varphi\left(x_{1}, \ldots, x_{m}\right)(\mathbf{0}) \\
& =\max _{x_{1}, \ldots, x_{m} \in S} \varphi\left(x_{1}, \ldots, x_{m}\right)(\mathbf{1})-\min _{x_{1}, \ldots, x_{m} \in S} \varphi\left(x_{1}, \ldots, x_{m}\right)(\mathbf{1}) .
\end{align*}
$$

This quantity measures how much the probability transitions depend on the value of the neighbourhood.

Let us consider the oriented graph $G$ describing the dependences between sites in the space-time diagram of the PCA. The set of vertices of $G$ is $\mathbb{Z} \times \mathbb{N}$, and there is an edge from $(k, t)$ to $(\ell, t+1)$ if $k \in \ell+\mathcal{N}$. For a given parameter $p \in$ $(0,1)$, the directed site percolation on $G$ consists in declaring each site to be open with probability $p$, independently for different sites. One can show that there is a critical value $p_{c}(\mathcal{N}) \in(0,1)$, such that if $p<p_{c}(\mathcal{N})$, then there is almost surely no infinite open (oriented) component (note that $p_{c}(\mathcal{N}) \geq 1 / \operatorname{Card} \mathcal{N}$ ).

By dominating the process of symbols ? in the space-time diagram of the envelope PCA by a directed site percolation of parameter $p_{\text {? }}$, one proves that if $p_{\text {? }}<p_{c}(\mathcal{N})$, then the symbols ? die out. Next proposition follows.

Proposition 2. Let $p_{c}(\mathcal{N})$ be the critical value of the two-dimensional directed site percolation of neighbourhood $\mathcal{N}$. If $p_{?}<p_{c}(\mathcal{N})$, then $\tilde{\Phi}^{t} \delta_{q_{Z}^{Z}}[?] \underset{t \rightarrow+\infty}{ } 0$, so that the PCA $\Phi$ is ergodic.

For a noisy CA with noise $R$ on a binary symbol set, one can check that unless the CA is constant, $p_{\text {? }}=\left|R_{0,1}-R_{1,1}\right|$, which shows ergodicity for $R_{0,1}$ and $R_{1,1}$ close enough to each other.

For elementary PCA, we have $p_{c}(\mathcal{N}) \approx 0.7$. If $p_{\text {? }}=\max _{i, j \in S^{2}} \theta_{i j}-$ $\min _{i, j \in S^{2}} \theta_{i j}$ is smaller than this critical value, then the PCA is ergodic.

### 2.3 Coupling from the Past

If the PCA $\tilde{\Phi}$ is ergodic, that is, if $\tilde{\Phi}^{t} \delta_{?^{Z}}[?] \underset{t \rightarrow+\infty}{\longrightarrow} 0$, then the unique invariant distribution of $\pi$ can be sampled exactly by coupling from the past (Propp-Wilson method). We define the update function $h: \tilde{S}^{m} \times(0,1) \rightarrow \tilde{S}$ of $\tilde{\Phi}$ by:

$$
h\left(y_{1}, \ldots, y_{m}, u\right)= \begin{cases}\mathbf{0} & \text { if } 0 \leq u \leq \tilde{\varphi}\left(y_{1}, \ldots, y_{k}\right)(\mathbf{0}) \\ \mathbf{1} & \text { if } 1-\tilde{\varphi}\left(y_{1}, \ldots, y_{k}\right)(\mathbf{1}) \leq u \leq 1 \\ ? & \text { otherwise }\end{cases}
$$

This function has the property that if $\left(U_{k}\right)_{k \in \mathbb{Z}}$ is a family of independent random variables, uniformly distributed on $(0,1)$, then for any $y \in \tilde{S}^{\mathbb{Z}}$, the configuration $\left(Z_{k}\right)_{k \in \mathbb{Z}}$ defined by $Z_{k}=h\left(y_{k+n_{1}}, \ldots, y_{k+n_{m}}, U_{k}\right)$ is distributed according to $\tilde{\Phi} \delta_{y}$.

Let $K$ be a finite subset of $\mathbb{Z}$. We draw a sequence $\left(u_{k,-t}\right)_{k \in \mathbb{Z}, t \geq 0}$ of independent random samples, uniformly distributed on $(0,1)$. We iterate the envelope PCA $\tilde{\Phi}$ from time $-T$ to time 0 , starting with the configuration $?^{\mathbb{Z}}$ at time $-T$, and always using the sample $u_{k,-t}$ to update cell $k$ at time $-t$, with the help of the update function $h$. If for some time $T$, the resulting configuration obtained at time 0 with this procedure is such that there are no symbols? on the cells of $K$, then the symbols observed on $K$ are distributed according to the marginal of the distribution $\pi$ on $K$.

After having iterated the envelope PCA from some time $-T$, the effect of starting from time $-(T+1)$ can only be to change some ? in the space time diagram into symbols $\mathbf{0}$ and $\mathbf{1}$ (once a symbol $\mathbf{0}$ or $\mathbf{1}$ appears at a cell $(k,-t)$ of the space time diagram, it is fixed). The fact that $\tilde{\Phi}^{t} \delta_{?^{Z}}[?] \rightarrow 0$ ensures that there exists almost surely a time $T$ such that at time $t=0$, there are no more symbols ? on the cells of $K$. Note that when implementing this sampling procedure, it is enough to consider only the cells that are in the dependence cone of $K$.

## 3 The Noisy Hardcore CA

Let us consider the elementary PCA $\Phi$ defined by the parameters $\theta_{00}=1-$ $p, \theta_{\mathbf{0 1}}=\theta_{\mathbf{1 0}}=\theta_{\mathbf{1 1}}=q$, with $0<p+q \leq 1$. This PCA is the noisy version of
the deterministic CA of local rule $f(i, j)=(1-i)(1-j)$, with a noise defined by $R_{\mathbf{1}, \mathbf{0}}=p$ and $R_{\mathbf{0}, \mathbf{1}}=q$. The rule $f$ can be seen as an exclusion rule: a cell becomes a $\mathbf{1}$ if it has no neighbour in state $\mathbf{1}$, and a $\mathbf{0}$ otherwise. For $\Phi$, after applying that deterministic dynamics, each $\mathbf{1}$ is changed into a $\mathbf{0}$ with probability $p$, and each $\mathbf{0}$ is changed into a $\mathbf{1}$ with probability $q$, independently.

Until recently, the question of ergodicity of $\Phi$ was a repertoried open problem. This PCA is closely related to the enumeration of directed lattice animals, which are classical objects in combinatorics. It also appears in the study of a percolation game [3]. With the notations of Sect. 2.2, we have $p_{\text {? }}=1-p-q$. Thus, the criterion of Proposition 2 provides the ergodicity of $\Phi$ for $p+q>0.3$ or so, but is of no help for smaller values of $p+q$. In that case, the comparison with oriented percolation is too rough to prove that symbols ? die out. Nevertheless, one can prove the following (see [3] for complete proof in the case $p=0$ or $q=0$ ).

Proposition 3. The noisy hardcore CA, that is, the elementary PCA $\Phi$ defined by $\theta_{00}=1-p, \theta_{01}=\theta_{10}=\theta_{11}=q$, is ergodic for any parameters $p+q<1$.

Proof (sketch). The local function of $\tilde{\Phi}$ is given by the following probability transitions (time is going up, and symbol $*$ represents any element of $\tilde{S}$ ).


The envelope PCA $\tilde{\Phi}$ is itself the noisy version of the deterministic CA $\tilde{F}$ (which is the deterministic CA obtained when taking $p=q=0$ in the table above), with a noise $\tilde{R}$ that changes any symbol into a $\mathbf{0}$ with probability $p$, and into a 1 with probability $q$.

For a given configuration in $\tilde{S}^{\mathbb{Z}}$, let us weight the occurrences of the symbols ? as follows:

- if a ? is followed by the pattern $\mathbf{0 1}$, then it receives weight 3 ;
- if a ? is followed by a $\mathbf{0}$ and then by something other than a $\mathbf{1}$, it receives weight 2;
- otherwise, a ? receives weight 1.

One can prove that the weight can only decrease under the action of the deterministic CA $\tilde{F}$. Precisely, if $\mu$ is a shift-invariant and reflection-invariant distribution on $\tilde{S}^{\mathbb{Z}}$, then $\tilde{F} \mu[\mathbf{?} 01]+\tilde{F} \mu[\mathbf{?} 0]+\tilde{F} \mu[\boldsymbol{?}] \leq \mu[\boldsymbol{?} 01]+\mu[? 0]+\mu[\boldsymbol{?}]$.

We now add the random noise $\tilde{R}$, and consider the PCA $\tilde{\Phi}$. Let $\mu$ be an invariant distribution of $\tilde{\Phi}$. By some computations, one can prove that we necessary have $\mu[?]=0$, since otherwise, we would get $\tilde{\Phi} \mu[? 01]+\tilde{\Phi} \mu[? 0]+\tilde{\Phi} \mu[?]<$ $\mu[? 01]+\mu[? 0]+\mu[?]$, which would be in contradiction with the fact that $\mu$ is an invariant distribution of $\tilde{\Phi}$.

Thus, there is no shift-invariant and reflection-symmetric stationary distribution in which the symbol ? appears with positive probability. Let us iterate $\tilde{\Phi}$ from the configuration $?^{\mathbb{Z}}$. By a coupling argument, the density $\tilde{\Phi}^{t} \delta_{?^{Z}}[?]$ is decreasing with $t$. If it was not converging to 0 , then we could extract a convergent subsequence of the Cesàro sums of $\tilde{\Phi}^{t} \delta_{?^{z}}$ and obtain an invariant distribution $\mu$ of $\tilde{\Phi}$ satisfying $\mu[?]>0$, which is not possible. Thus, $\tilde{\Phi}^{t} \delta_{? Z}^{Z}[?] \xrightarrow[t \rightarrow+\infty]{ } 0$, and $\Phi$ is ergodic.

## 4 Perturbating a Nilpotent CA

Let $F$ be a nilpotent CA . It means that there exists an integer $N$ such that $F^{N}$ is a constant function, equal to $\alpha^{\mathbb{Z}}$ for some symbol $\alpha \in S$.

If $F$ is not the constant function equal to $\alpha^{\mathbb{Z}}$, then for an $\varepsilon$-perturbation of $F$ with small $\varepsilon$, the value $p$ ? is close to 1 . Thus, Proposition 2 (and its analogous for larger symbol sets) cannot be used to prove the ergodicity of an $\varepsilon$-perturbation of $F$. In that case, the envelope PCA as defined in Sect. 2.1 is not an adapted tool. Nevertheless, once again, the coupling method can be used to prove the ergodicity.
Proposition 4. Let $F$ be a nilpotent CA. There exists $\varepsilon_{c}>0$ such for $\varepsilon<\varepsilon_{c}$, any $\varepsilon$-perturbation of $F$ is ergodic.
Proof. Let $\varepsilon>0$, and let $\Phi_{\varepsilon}$ be an $\varepsilon$-perturbation of $F$. We prove that if $\varepsilon$ is small enough, we can couple all the trajectories of $\Phi_{\varepsilon}$.

Let $K$ be a finite subset of points of $\mathbb{Z}$. We consider a configuration $\left(x_{k, 0}\right)_{k \in \mathbb{Z}}$ obtained at time $t=0$, after iterating $\Phi_{\varepsilon}$ from a given time in the past, using a sequence $\left(u_{k,-t}\right)_{k \in \mathbb{Z}, t \geq 0}$ of independent samples, uniformly distributed in $(0,1)$, and an update function having the following property: if $u_{k,-t}>\varepsilon$, then cell $k$ is updated according to the local rule of the deterministic CA $F$, while if $u_{k,-t} \leq \varepsilon$, the value may differ. We prove that almost surely, there exists a time $T>0$ such that the evolutions from all the possible starting configurations at time $-T$ provide the same sequence $\left(x_{k, 0}\right)_{k \in K}$ at time 0 .

We define recursively the sets $\mathcal{N}_{i}$ by $\mathcal{N}_{0}=\{0\}$, and $\mathcal{N}_{i+1}=\mathcal{N}_{i}+\mathcal{N}=$ $\left\{a+b ; a \in \mathcal{N}_{i}, b \in \mathcal{N}\right\}$ for $i \geq 0$, so that $\mathcal{N}_{t}$ is the neighbourhood of $F^{t}$.

For $k \in \mathbb{Z}$, and times $t, i \geq 0$, we define $V_{i}(k,-t)=\left\{(k+\ell,-t-i), \ell \in \mathcal{N}_{i}\right\}$. It is the set of cells at time $-t-i$ in the space-time diagram from which the state of cell $k$ at time $-t$ may depend.

We also introduce $W(k,-t)=\bigcup_{0 \leq i \leq N-1} V_{i}(k,-t)$, where we recall that $N$ is such that $F^{N}$ is constant, equal to $\alpha^{\mathbb{Z}}$.

We call a cell $(k,-t)$ an error if $u_{k,-t} \leq \varepsilon$. Since $F^{N}$ is a constant function, if there is no error in $W\left(k,-t_{0}\right)$, then the value $x_{k,-t_{0}}$ of cell $k$ at time $-t_{0}$ does not depend on the value of the configuration at time $-t_{0}-N$ (we have $x_{k,-t_{0}}=\alpha$ in all cases).

For $k \in \mathbb{Z}, t \geq 0$, let us define the set:

$$
E(k,-t)=\left\{\begin{array}{l}
\emptyset \text { if there is no error in } W(k,-t) \\
V_{N}(k,-t) \text { if there is at least one error in } W(k,-t)
\end{array}\right.
$$

We define recursively the sets $\left(A_{i}\right)_{i \geq 0}$ by $A_{0}=K \times\{0\}$, and:

$$
A_{i+1}=E\left(A_{i}\right)=\bigcup_{(k,-t) \in A_{i}} E(k,-t) \text { for } i \geq 0
$$

Note that if $(k,-t) \in A_{i}$, then $t=i N$. We have the following property: if $A_{i}$ is empty, then if we iterate $\Phi_{\varepsilon}$ from time $-i N$ to time 0 , using the samples $\left(u_{k,-t}\right)_{k \in \mathbb{Z}, 0 \leq t<i N}$, the values $\left(x_{k, 0}\right)_{k \in K}$ obtained on $K$ at time 0 do not depend on the choice of the configuration $\left(x_{k,-i N}\right)_{k \in \mathbb{Z}} \in S^{\mathbb{Z}}$ from which we start at time $-i N$.

In the figure below, errors are represented by red dots. Blue domains represent cells that are known to be in state $\alpha$ (there are no errors affecting them in the last $N$ time steps). Black domains represent cells for which we need further information in the past to determine their state.


Let us prove that if $\varepsilon$ is small enough, then almost surely, there exists an integer after which all the sets $A_{i}$ are empty.

We set $m_{i}=\operatorname{Card} \mathcal{N}_{i}$. Let $(\ell,-t)$ be an error, with $t=i N+j, 0 \leq j \leq N-1$. We have $(\ell,-t) \in W(k,-i N)$ if and only if $k \in \ell-\mathcal{N}_{j}$. Thus, the number of points $(k,-i N)$ such that $(\ell,-t)$ is an error of $W(k,-i N)$ is bounded by $m_{j} \leq m_{N-1}$. If there is an error in $W(k,-i N)$, then $\operatorname{Card} E(k,-i N)=m_{N}$. It follows that a given error has a contribution of at most $L=m_{N-1} m_{N}$ points to $A_{i+1}$.

Let $M=m_{0}+m_{1}+\ldots+m_{N-1}$. We have Card $W(k,-t)=M$ for any $k \in \mathbb{Z}, t \geq 0$. The number of points of $\bigcup_{k \in A_{i}} W(k,-i N)$ is thus smaller than (Card $\left.A_{i}\right) \times M$, and each point is an error with probability $\varepsilon$, independently. Consequently, Card $A_{i+1}$ is bounded by the sum of $\left(\operatorname{Card} A_{i}\right) \times M$ independent random variables, whose value is $L$ with probability $\varepsilon$, and 0 with probability $1-\varepsilon$. If $\varepsilon<1 / L M$, a comparison with a branching process proves that there is extinction: almost surely, the sets $A_{i}$ are eventually empty. Consequently, $\Phi_{\varepsilon}$ is ergodic (note that the bound given for $\varepsilon$ is rough and can certainly be improved).

## 5 Noisy Permutive Cellular Automata

Let $F$ be a CA of neighbourhood $\mathcal{N}=\{\ell, \ell+1, \ldots, r\}$ and local function $f$ : $S^{m} \rightarrow S$, with $m=r-\ell+1 \geq 2$. We say that $F$ is left-permutive (resp. right-permutive) if, for all $w=w_{\ell+1} \cdots w_{r} \in S^{m-1}$, the mapping:

$$
\begin{aligned}
\tau_{w}: S & \rightarrow S \\
a & \mapsto f(a w) \quad(\text { resp. } f(w a))
\end{aligned}
$$

is bijective. A CA is permutive if it is either left or right-permutive. It is bipermutive if it is both left and right-permutive. For example, if $S=\mathbb{Z}_{n}$ and $a, b, c \in \mathbb{Z}_{n}$, the affine CA defined by $f(x, y)=a x+b y+c$ is left-permutive (resp. rightpermutive) if $a$ (resp. $b$ ) is invertible in $\mathbb{Z}_{n}$.

Let $F$ be a permutive CA. Using the bijections $\tau_{w}$ one can prove that $F$ is surjective. For deterministic CA, surjectivity is equivalent to preserving the uniform distribution $\lambda$ on $S^{\mathbb{Z}}$ (that is, the product of the uniform distribution on $S$ ). Next proposition shows that when adding a noise preserving $\lambda$, the PCA indeed converges to $\lambda$. The proof below is adapted from a work of Vasilyev [2, 4].

Proposition 5. Let $F$ be a permutive $C A$, and let $R$ be a positive noise preserving the uniform distribution. The noisy version $\Phi$ of $F$ with noise $R$ is ergodic, and its unique invariant distribution is the uniform distribution $\lambda$.

Proof. We will prove that for any $N \in \mathbb{N}$, and any initial distribution $\mu$ on $S^{\mathbb{Z}}$, the marginal distribution of $\Phi^{t} \mu$ on $K=\{-N, \ldots, N\}$ converges exponentially to the uniform Bernoulli distribution on $S^{K}$, that we denote by $\lambda_{K}$. Precisely, we will prove that for any $N \in \mathbb{N}$, there exists $\theta<1$ such that for any distribution $\mu$ on $S^{\mathbb{Z}}$, we have: $\forall t \geq 0,\left\|\left(\Phi^{t} \mu\right)_{\mid K}-\lambda_{K}\right\|_{1} \leq 2 \theta^{t}$, where for $u: S^{K} \rightarrow \mathbb{R}$, $\|u\|_{1}=\sum_{x \in S^{K}}|u(x)|$.

Let us first assume that $F$ is left-permutive and that $\mathcal{N}=\{0,1, \ldots, r\}$, and let $w \in S^{r}$. By permutivity of $F$, we have a bijection

$$
\begin{aligned}
\sigma_{w}: S^{K} & \longrightarrow S^{K} \\
x & \longmapsto f(x w),
\end{aligned}
$$

where we still denote by $f$ the map from $S^{K \cup\{N+1, \ldots, N+r\}}$ to $S^{K}$ induced by the local function of the CA $F$.

So, when fixing the word $w$ as a boundary condition on the right of $K$, the noisy CA $\Phi$ maps a word $x \in S^{K}$ to a random word $Z_{K}$ distributed according to a product distribution with marginal distribution $R\left(y_{k}\right)$ at site $k \in K$, where $y=\sigma_{w}(x)$. From a given $x \in S^{K}$, we denote by $P_{w}(x, z)$ the probability for $Z_{K}$ to be equal to some $z \in S^{K}$, so that: $P_{w}(x, z)=\prod_{k \in K} R_{y_{k}, z_{k}}$.

Recall that $\lambda_{K}$ is the uniform distribution on $S^{K}$. The map $\sigma_{w}$ being bijective, it preserves $\lambda_{K}$. By assumption, the noise $R$ also preserves the uniform distribution, so that we obtain $P_{w} \lambda_{K}=\lambda_{K}$.

For any $w \in S^{r}$, the transition matrix $P_{w}$ is positive. Therefore, there exists $\theta_{w}<1$ such that for any probability distributions $\nu, \nu^{\prime}$ on $S^{K}$, we have

$$
\left\|P_{w} \nu-P_{w} \nu^{\prime}\right\|_{1} \leq \theta_{w}\left\|\nu-\nu^{\prime}\right\|_{1}
$$

the above inequality being true in particular for $\theta_{w}=1-\varepsilon_{w}$, where $\varepsilon_{w}=$ $\min \left\{P_{w}(i, j) ; i, j \in S\right\}$.

Let us set $\theta=\max \left\{\theta_{w} ; w \in S^{r}\right\}$. It follows that for any sequence $\left(w_{t}\right)_{t \geq 0}$ of words of $S^{r}$, we have: $\left\|P_{w_{t-1}} \ldots P_{w_{1}} P_{w_{0}} \nu-P_{w_{t-1}} \ldots P_{w_{1}} P_{w_{0}} \nu^{\prime}\right\|_{1} \leq \theta^{t}\left\|\nu-\nu^{\prime}\right\|_{1}$. In particular, for $\nu^{\prime}=\lambda_{K}$, we obtain that for any distribution $\nu$ on $S^{K}$ and any sequence $\left(w_{t}\right)_{t \geq 0}$ of words of $S^{r},\left\|P_{w_{t-1}} \ldots P_{w_{1}} P_{w_{0}} \nu-\lambda_{K}\right\|_{1} \leq \theta^{t}\left\|\nu-\lambda_{K}\right\|_{1} \leq 2 \theta^{t}$.

| Time |
| :--- | :--- | :--- | :--- |
| $t=3$ |
| $t=2$ |
| $t=1$ |
| $t=0$ | |  |  |  |  |
| :--- | :--- | :--- | :--- |
|  | $V_{3} \sim P_{w_{2}} P_{w_{1}} P_{w_{0}} \nu$ | $w_{3}$ |  |
|  | $V_{2} \sim P_{w_{1}} P_{w_{0}} \nu$ | $w_{2}$ |  |
|  | $V_{1} \sim P_{w_{0}} \nu$ | $w_{1}$ |  |
|  | $V_{0} \sim \nu$ | $w_{0}$ |  |
| $-N$ |  | $N+r$ |  |

Let now $\mu$ be a distribution on $S^{\mathbb{Z}}$. When iterating $\Phi$, it induces a random sequence of words $\left(W_{t}\right)_{t \geq 0}$ on $\{N+1, \ldots, N+r\}$. Using the above inequality, we get:

$$
\forall t \geq 0,\left\|\left(\Phi^{t} \mu\right)_{\mid K}-\lambda_{K}\right\|_{1} \leq \max _{w_{0}, \ldots, w_{t-1} \in S^{r}}\left\|P_{w_{t-1}} \ldots P_{w_{1}} P_{w_{0}} \mu_{\mid K}-\lambda_{K}\right\|_{1} \leq 2 \theta^{t}
$$

If the neighbourhood of $F$ is not of the form $\mathcal{N}=\{0,1, \ldots, r\}$, then there exists $s \in \mathbb{Z}$ such that $F \circ \sigma^{s}$ is a left-permutive CA having a neighbourhood of that form. The noisy version of $F \circ \sigma^{s}$ is $\Phi \circ \sigma^{s}$, and the previous inequality provides: $\left\|\left(\left(\Phi \circ \sigma^{s}\right)^{t} \mu^{\prime}\right)_{\mid K}-\lambda_{K}\right\|_{1} \leq 2 \theta^{t}$, for any distribution $\mu^{\prime}$. In particular, for $\mu^{\prime}=\sigma^{-s t} \mu$, since $\Phi$ and $\sigma$ commute, we obtain: $\left\|\left(\Phi^{t} \mu\right)_{\mid K}-\lambda_{K}\right\|_{1} \leq 2 \theta^{t}$, which ends the proof. The right-permutive case is analogous.

In a collaboration still in progress with S. Taati and M. Sablik, we investigate the ergodicity of more general noisy surjective CA.

Concerning elementary PCA, it is still a challenging open question whether they are ergodic as soon as $\theta_{i j} \in(0,1)$ for all $i, j \in S$.

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# Types in Programming Languages, Between Modelling, Abstraction, and Correctness <br> Extended Abstract 

Simone Martini ${ }^{(\boxtimes)}$<br>Dipartimento di Informatica-Scienza e Ingegneria, Università di Bologna and INRIA, Bologna, Italy<br>simone.martini@unibo.it

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## 1 Introduction

The notion of type to designate a class of values, and the operations on those values, is a central feature of any modern programming language. In fact, we keep calling them programming languages, but the part of a modern language devoted to the actual specification of the control flow (that is, programming stricto sensu) is only a fraction of the language itself, and two different languages are not much apart under that perspective. What "makes a language" are much more its modelling capabilities to describe complex relations between portions of code and between data. In a word, the central part of a language is made by the abstraction mechanisms it provides to model its application domain(s), all issues the language theorist may well group together in the type chapter of a language definition.

The conquest of the summit by the notion of type is the result of a rather slow process in the history of programming languages. In a previous paper [21] we have sketched some of the earliest history, observing that the concept of type we understand nowadays is not the same it was perceived in the sixties, and that it was largely absent (as such) in the programming languages of the fifties. While the technical term "type" arrives on the scene at the end of the fifties (for sure in the report on Algol $58[26])^{1}$, the use of types as a modelling tool for the "objects of the real world" is the contribution of the sixties (in particular under the influence of McCarthy [22] and Hoare [15]), which will materialize

[^17]in languages like Algol W [30] or Pascal. Moreover, we observed in [21] that the notion of "type" of programming languages, which we now conflate with the concept of the same name of mathematical logic, is instead relatively independent from the logical tradition, until the Curry-Howard isomorphism [18] will make an explicit bridge between them. The connection between these two concepts remains anonymous for a long time - some of the people knew very well the other field, and it is certain that, from mid sixties, the mathematical logic work started influencing programming languages (we think, among other, to Landin, Scott, Strachey, Hoare, McCarthy, Morris etc.). But there is no explicit, mutual recognition - concepts and formal systems are systematically re-discovered in the two fields. The first explicit connection we know of, in a non technical, but explicit, way is [16].

The present paper will elaborate on this story, focusing on that fundamental period covering the seventies and the early eighties. It is there that the types become the cornerstone of the programming language design, passing first from the abstract data type (ADT) movement and blossoming then into the objectoriented paradigm. This will also be the occasion to reflect on how it could have been possible that a concept like ADTs, with its clear mathematical semantics, neat syntax, and straightforward implementation, could have given way to objects, a lot dirtier from any perspective the language theorist may take.

## 2 Modeling and Correctness

A central issue of the story we have told in [21] is the provision of a language mechanism to introduce new data types, in an extensible way (that is, different from a palette of types fixed at language design time). This materializes in the two related proposals of records and typed references [15], and Simula's classes [9]. A further, important realization is that a type is given not only by its class of values, but it is defined together the operations acting on those values (see, e.g., $[8]^{2}$ ). Simula was an extension of Algol 60 designed for discrete event simulation. One of the main concepts in Simula is the class ${ }^{3}$ - the specification of both data and operations, which may have several dynamic instances (objects in the modern terminology, or processes in Simula I), whose life is not required to be dynamically nested. Under this view, Simula classes are a good candidate as a language mechanism for the definition of such types, since they permit the simultaneous definition of data and operations. However, this has to be seen together with the need to enforce some level of correctness, at a syntactic (and if possible, static) way - a central feature of what Priestley [27] calls the "Algol

[^18]research programme ${ }^{4 "}$, where the design of programming languages should assist (or even guide) the programmer in avoiding bugs or, worse, unintended behaviors in a program. While Simula classes may be used to define types as "data plus operations", they do not provide the necessary abstraction to enforce correctness, because the language does not distinguish between a type and its implementation (between, say, a stack, and the list used to implement that stack: a stack may be incautiously manipulated by any operation on lists). The need for this abstraction is the core of much literature on data and programming languages in the early seventies.

## 3 Abstract Data Types

The search for economic and terse linguistic constructs comes together with the need for the definition of a precise semantics for those constructs. Parnas' seminal [25] introduces the term information hiding, meaning that a stable interface towards the rest of the program should protect the design choices which are bound to change (which may thus evolve without affecting the other parts of the program). In Parnas' view this is a general design methodology, which applies to types, modules, packages, etc. It is a crucial forward step of the programming language community that this hiding should be enforced by linguistic abstraction mechanisms, and not merely guaranteed by a design methodology. Looking at the published literature (e.g., Morris [24], Hoare [17], Goguen [12], Liskov and Zilles [20], Reynolds [29]-who also explicitly introduces the expression "representation independence" ${ }^{5}$ - , Guttag [14], etc.), we see that around 1972-1973 the time is ripe for a substantial achievement. If Hoare [17] uses an axiomatic settings, starting from Goguen [12] the (informal) abstraction requirement is described semantically by freeness - a data type is, in its mathematical semantics, a free algebra over a set of constructors (that is, non-interpreted function names). Freeness means that one cannot uses implementation dependent information on a value, because a values is simply an inductive construction over the constructors-hence abstraction. Moreover, this provides a powerful prooftechnique on programs - structural induction [3, 4]. Finally, equations on termsand hence all the good properties of equational logic-provide the axiomatic semantics needed to distinguish between types (algebras) over isomorphic sets

[^19]of constructors, but with different behavior (stacks and queues, say). The notion of abstract data type will make into programming languages with CLU [19], which will have a significant impact on subsequent languages.

At the end of the seventies, thus, it seems that types in programming languages are in a successful, and positive position-type abstraction mechanisms made into good (albeit essentially academic) programming languages, and their linguistic constructs come with a clear and mathematically sound formal semantics. However, in a sort of coup de théâtre, at the peak of their success ADTs will have to give way to another concept-objects, a much dirtier mechanism, able to enforce a lot less correctness than ADTs. Objects, and not ADTs will be the real players of the programming languages of the following decade(s).

## 4 Objects

To understand the fall of ADTs we must contrast correctness with flexibility. The utopia is to have them both at the same time, but realism tells us that they represent, most of the time, conflicting aims. ADTs provide abstraction at the expenses of (reuse and, then, of) compatibility. We cannot give here all the details of an example of the problem ${ }^{6}$. Suppose only to have an ADT $T$ with an operation $f$, which is then extended into a new type $T 1$, sharing the same values of $T$ but with an extended set of operations and for which the operation $f$ is also redefined. It is now natural (and convenient) to assume that the language enforces that $T 1$ is compatible with $T$ (a value of $T 1$ may appear in any context requiring $T$ ). The problem now is that in an expression like $f(t)$ the choice of which code for $f$ will be executed (the one for $T$ or the one for $T 1$ ) depends on the static context, that is, on the static type of $t$. Thus, if $t$ (of static type $T$ ) references a value $t 1$ of type $T 1$ (which may correctly happen, in view of the compatibility of $T 1$ with $T$ ), it is the erroneous $f$ to be applied to a value of $t 1$-abstraction breaks, which means that compatibility of $T 1$ and $T$ must be abandoned, which results in a drastic programming burden.

From this perspective the solution is easy-allow for a dynamic choice for the selection of the code for $f$ in $f(t)$, depending not from the static typing of $t$, but from the actual type of the value referenced by $t$. In the programming language jargon, do not use functions and overloading, and use instead methods and dynamic lookup. Object oriented programming may be seen as the result of this observation. It is a paradigm where: (i) there is mechanism which, under certain conditions, supports the inheritance of the implementation of certain operations from other, analogous constructs; (ii) there is a notion of compatibility defined in terms of the operations admissible for a certain construct; (iii) operations on values are dynamically selected on the basis of the "actual type" of the arguments to which they are applied. These features together allow for the flexibility of the paradigm when used in actual programming of large scale systems. But, at the same time, these features together cannot be given semantics in the clean framework of algebraic types, at least in their simple formulation. Providing a

[^20]sound semantics and a formal treatment of objects will be a challenge for almost fifteen years (see, e.g., $[1,2,28]$ for the references therein). Finally, we will have to explain why types are so central for objects, in view of the lack of types in languages like Smalltalk [13]. The distinction made in [6] (and the discussion in [5]) will guide our reflection.

## 5 Conclusions

The history of computer science is innervated by the continuous tension between formal beauty and technological effectiveness. Types in programming languages are an evident example of this dialectics. They are introduced for a better verification of the correctness of programs, and yet - contrary to mathematical logicthey must be experienced by the working programmer as an enabling feature ${ }^{7}$, allowing for simpler writing of programs.

In its formal approach, computer science never used ideological glasses (types per se; constructive mathematics per se; linear logic per se; etc.), but exploited what it found useful for the design of more elegant, economical, usable artifacts. This eclecticism (or even anarchism, in the sense of epistemological theory) is one of the distinctive traits of the discipline, and one of the reasons of its success.

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# AFCAL and the Emergence of Computer Science in France: 1957-1967 

Pierre-Éric Mounier-Kuhn ${ }^{1}$ and Maël Pégny ${ }^{2(\boxtimes)}$<br>${ }^{1}$ CNRS, Université Paris-Sorbonne, Maison des Sciences de l'Homme, Paris, France mounier@msh-paris.fr<br>${ }^{2}$ CNRS, Université de Paris 1 Panthéon-Sorbonne, IHPST, Paris, France maelpegny@gmail.com


#### Abstract

Founded in 1957, the Association Française de Calcul (AFCAL) was the first French society dedicated mainly to numerical computation. Its rapid growth and amalgamation with sister societies in related fields (Operations Research, Automatic Control) in the 1960s resulted in changes of its name and purpose, including the invention and adoption of the term informatique in 1962-1964, then in the adoption of cybernétique in 1967. Our paper aims at explicating the motives of its creation, its evolving definition and the functions it fulfilled. We seek to understand how this association, altogether a learned and a professional society, contributed to the emergence and recognition of Computing as an academic discipline in France. The main sources are the scattered surviving records of AFCAL, conserved in the archives of the Observatoire de Paris, of the Institut de Mathématiques appliquées de Grenoble (IMAG) and of the CNRS' Institut Blaise Pascal in Paris, as well as AFCAL's first congress and journal, Chiffres.


## 1 Introduction

The Association Française de Calcul (AFCAL) was founded in 1957 by French academics and industry engineers who wanted to join forces, to form a specialist community promoting applied mathematics and computing technologies in the context of the post-war modernization. With this technoscientific and political agenda, the new society soon attracted hundreds of members and support from various organizations.

It would be anachronistic to retrospectively tag AFCAL as a champion of "computer science". It focused initially on numerical analysis, a sub-discipline of applied mathematics, and on the use of calculating machines. The idea that electronic computers and their programs could become a new, distinct scientific academic field did not emerge until the 1960s. One of the main questions addressed in this paper is precisely how AFCAL participated in this emergence: what convergence of agenda and which internal tensions came to shape digital computation as a new academic discipline ${ }^{1}$.

[^22]The archives of the association disappeared accidentally in a flood in the 1980s, making our research a bit difficult. The main sources are scattered surviving records of AFCAL, conserved in a variety of institutions: the archives of the Observatoire de Paris keep a box of papers from AFCAL's early years ${ }^{2}$, those of the Institut de Mathématiques appliquées de Grenoble and of the CNRS Institut Blaise Pascal in Paris contain correspondences and reports exchanged between founding members. There are certainly more records waiting to be discovered in other organizations or with private persons. Published sources include AFCAL journal Chiffres (from 1958) and its conferences (from 1960). Finally, the first meeting on the history of computing in France included several papers addressing the evolution of French societies in this field ${ }^{3}$.

In this sketch of our paper ${ }^{4}$, we will highlight three episodes of the association's beginnings: the circumstances and motives of its creation; its participation to the international congress ICIPS and the international association IFIP; the functions of the association and their evolutions. In this last part, we will examine the role of the association in the emergence of informatique as a new scientific discipline.

## 2 The Emergence of a Computing Community

While the French academic spheres in the post-war period were dominated by pure mathematics, a growing demand for applied mathematics was expressed by physicists, and by various branches of engineering. From the late 1940s on, several young university professors responded to this demand by creating courses in numerical analysis and computing laboratories (Mounier-Kuhn 2012). In the early 1950s, electronic calculators were acquired from manufacturers and installed in these laboratories or, on a larger scale, in big organizations such as Électricité de France (E.D.F.) and aeronautics companies. In 1955, the first stored-program computers entered service in France, considerably reinforcing these facilities and offering new perspectives to mathematical modeling as well as to data processing.

Within two years, a series of learned societies were founded in related fields, a veritable blooming of professional communities active in Operations Research (SOFRO), Automatic Control (AFRA), Measure and Regulation (AFIC) and Computing (AFCAL). Simultaneously, scientists with political connections established a governmental council, CSRSPT (Conseil Supérieur de la Recherche Scientifique et du Progrès Technique), in order to support the modernization of research and higher education. This body identified automatic control, applied mathematics and computing as a priority sector, under the label Cybernétique.

[^23]
### 2.1 Technoscientific Networking: Applied Mathematics as a Modernization Tool

One of the main actors of this movement was Jean Kuntzmann, an algebraist turned militant of applied mathematics, who had created a small laboratory of numerical analysis at the university of Grenoble to meet the needs of the local Institut Polytechnique ${ }^{5}$. His correspondance reveals his role as the initiator of the creation of AFCAL and sheds light on his motivations.

Kuntzmann's action to foster useful mathematics was supported by the local academic authorities and by the Ministry of Education which favored initiatives toward vocational studies. Most of the financial resources of his laboratory came from contracts with civilian or military clients who appreciated his expertise and, in turn, brought stimulating problems. The synergy between funding, research, teaching and practical computing generated a cumulative growth process. In the mid-1950s, Kuntzmann was recognized in France and abroad as a leading specialist in his field. He was consulted by junior professors who wanted to set up their own computing laboratories (in Lille, Nancy, Barcelona and other universities). His laboratory attracted visiting teachers and doctoral students.

Having systematically informed the major French users about the services his laboratory could provide, and received acknowledgements of their interest, Kuntzmann had built an interactive network of colleagues and partners. In April 1956, he began contacting other French specialists to suggest an association.

His correspondents, men of his generation or younger engineers and professors, shared his project and responded positively. This first poll allowed Kuntzmann to solicit and obtain the patronage of academic notabilities at a higher hierarchical level, the dean of the Paris faculty of science and the director of Institut Henri Poincaré ${ }^{6}$. He did not explain his purpose in detail, as these men obviously knew the usefulness of a learned society: Kuntzmann's correspondence contains no negative answer, not a single letter from a colleague objecting to his project.

Already a member of several learned societies, Kuntzmann had a precise idea of the one he wanted to create. He belonged to the Société française de mathématiques, which was still too general-purpose and exclusively academic to be of real use in his project. He also participated in the International Association for Analog Calculation, which by contrast covered only part of the domain. The closest model was the American Association for Computing Machinery (ACM) with its specialized journal, to which Kuntzmann subscribed.

In Spring 1956, Kuntzmann also became a member of the German Gesellschaft für Angewandte Mathematik und Mechanik (GAMM). Kuntzmann could not attend the inauguration of the PERM computer in May 1956, but, always keen on completing his documentation on the state of the art, the Grenoblois ordered the proceedings of the important Darmstadt conference held

[^24]the previous year ${ }^{7}$. The GAMM existed since the 1920s and, with its institutional weight, could be a model of what Kuntzmann and his partners wanted to create in France, where no learned society existed in applied mathematics until 1955.

It is revealing that Kuntzmann aimed to create a new society rather than join the cybernetician community, which had been popular since Norbert Wiener's manifesto in 1948, and which possessed its own organization. While Kuntzmann sent one of his doctoral students at the International Congress of Cybernetics (Namur, Belgium, 1956) and ordered its proceedings, he explained in a harsh letter why he refused to join the Association Internationale de Cybernétique: "I have always thought that the grandiose term of Cybernetics covers very diverse matters. Some are mathematical, others are technical. Unfortunately, on top of that comes some philosophy, and even certainly babble (fumisterie). Either Cybernetics shall fall into oblivion or, some day, one will have to separate what is robust from what is not. I reserve my adhesion (and my participation in meetings) for the moment when this decantation is achieved ${ }^{8}$."

We know that this rejection of Cybernetics, or of what it had become in the mid-1950s, was shared by other French computing pioneers. They believed Cybernetics could not provide the identity of their emerging profession.

Kuntzmann knew that an informal group existed already in Paris, gathering periodically a handful of specialists and distinguished amateurs in a seminar on numerical calculation. Somehow its lack of institutional strategy and of scientific ambition reflected the low status of reckoning in the past decades. Kuntzmann courteously suggested to integrate the group in his future association to broaden its scale and scope. The merger was completed swiftly at a meeting in Paris, during a conference on scientific management. It avoided a possible conflict and brought a mailing list of 200 corresponding members to the nascent AFCAL ${ }^{9}$. The Groupe de Calcul Numérique held its seminars at the Paris Institut d'Astrophysique, and the AFCAL naturally followed suit, obtaining an office and a mailbox in this convenient venue. More importantly, the director of the Institut d'Astrophysique, André Danjon, an energetic and openminded astronomer, agreed to become AFCAL's first president, bringing his own legitimacy and political connections.

Danjon had sent his young astronomers to learn programming on the IBM650 computer at IBM France, and was aware of the potential of the new machines. Beyond the technical importance of calculation in science, we can suggest the hypothesis of a solidarity between emerging disciplines. Just like computing in the years to come, astrophysics had to assert its legitimacy against the conservatism of traditional astronomy, which was entirely mathematical and had

[^25]reasons to view astrophysics as an uncertain science, lacking rigor, perhaps not a science at all: "Basically [for them], astrophysics could not be considered a science" (À l'extrême, l'astrophysique ne pouvait être considérée comme une science ${ }^{10}$ ).

Under Danjon's tutelage, AFCAL had two vice-presidents: Prof. Jean-André Ville, a mathematician who had just been put in charge of a computing center at the Paris faculty of science; and Jean Carteron, a Polytechnique engineer who headed the computing center at the R\&D Department within E.D.F., and was equally active in founding the association (EDF sponsored AFCAL generously) (Mounier-Kuhn 2010b). Kuntzmann reserved for himself the responsibility of the future journal, a position of scientific gate-keeper in the new field.

Kuntzmann's work to promote modern theories and practices in applied mathematics was in synergy with the efforts of the computer industry to market their machines and their applications, and with the need of mathematical modeling among users. The lists of AFCAL members (200 in 1957, some 500 six years later) and of the institutions where they worked offer a fairly complete picture of French applied mathematics ${ }^{11}$ : higher education and research, laboratories of major public or private companies, computer makers, consultants in mathematical and programming expertise. These various specialists felt the need to share their experiences and to promote their methods, which was the primary function of the association. It took nearly a year to create the association, as each founding member had many other urgent tasks. The constitutive assembly (May 31 1957), beyond adopting the charter, heard lectures on the present computing needs and resources among users, on recent machines' capabilities, and on the search for better programming methods that AFCAL should foster, a program covering most of the purposes of the association.

In June 1957, Kuntzmann set to create its journal, Chiffres, that is to find a suitable publisher and to hunt for good papers. The journal would also include sections devoted to scientific news, the announcement of courses and seminars, reviews of books and of conferences... In October, Carteron urged Kuntzmann to "give life as soon as possible to the Association" by organizing meetings, and both men gave it all the publicity they could in France and abroad. By November, Kuntzmann had formed an editorial board and gathered sufficient material for a first issue of Chiffres.

From its very beginning on, the association had an international dimension. Firstly, its membership soon included mathematicians working in Switzerland (Charles Blanc, Pierre Banderet, etc.), in Italy (particularly at the Olivetti-Bull joint venture) and in Belgium (Vitold Belevitch, Jacques Burniat). About 10 \% of AFCAL members in the early 1960s worked outside of France. Its audience still remained limited to French-speaking countries or readers, as its publications provided initially no English abstracts.

[^26]
## 3 Creating an Internationale of Computing: IFIP and Its Scientific Agenda

Soon after its creation, the association contributed to establish an internationale of computing specialists. In December 1957, AFCAL was contacted by an American engineer and entrepreneur, Isaac L. Auerbach. As a member of the US Association for Computing Machinery (ACM), he planned to come to Paris in order to organize an international meeting of Information Processing societies from different countries. Isaac Auerbach had just left the Burroughs Corporation to form his own consultant company. Establishing international contacts to help Auerbach Associates conduct market surveys was certainly an incentive. Yet the main rationale for this internationalization was, as in other learned societies, to exchange information and gain political weight collectively.

The French reacted positively and agreed to meet with Auerbach and to cooperate for the preparation of the conference ${ }^{12}$. Just as in the creation process of AFCAL, the movement to join forces and cooperate appears to have been quite smooth: it was somewhat obvious to all actors, and it hardly needed more explanation than a few words in a letter. In June 1958, in Paris again, an ad-hoc committee gathered computer experts from most developed countries to organize the conference.

The International Conference of Information Processing Societies (ICIPS) met in June 1959 at the UNESCO building in Paris, gathering some 1800 participants from 38 countries during a week. For many French scientists, it was the first opportunity to hear foreign scholars (particularly Americans) present their work, and to discover the state of the art in programming techniques and in non-numerical applications of computers, an important milestone in their evolution towards research in computer science. Beside processor architecture or computing and coding methods, sessions were devoted to machine translation, pattern recognition, machine learning or future perspectives in computing. This diffusion of information was precisely the purpose pursued by Isaac Auerbach and Pierre Auger, the physicist who headed the French delegation at Unesco ${ }^{13}$.

The ICIPS conference gave birth to the International Federation for Information Processing (IFIP) (see Auerbach 1986; Carteron 1996). From then on, IFIP set up a similar worldwide conference every three years in a major city (Munich in 1962, New York in 1965, etc.). The mere existence of this global learned society and of its conferences, comparable with those in astronomy or geophysics, contributed to establish a scientific status for information processing. Such a status was explicitly targeted since its creation, as highlighted by the general

[^27]secretary of the 1959 conference, Pierre Auger: "It is absolutely indispensable to have an international organization to respond to the needs of the science of Information Processing ${ }^{14 "}$.

Moreover, in subsequent years the IFIP created specific committees on specific problems: Software theory and practice, Algorithmic languages and calculi, Formal description of programming concepts, Programming languages, Programming methodology. IFIP thus not only constituted an international computing community, it also provided it with a scientific roadmap of research directions.

Being essentially French-speaking was not a serious handicap for the association, at a time when many mathematicians could at least read French as well as German. The association was thus part, from its beginnings, of an international family of sister societies with which it actively interacted. J. Carteron became IFIP's treasurer and general secretary. AFCAL members participated in the various IFIP committees and working groups, discussing research and education problems and drawing inspiration from foreign experiences (Mounier-Kuhn 2014).

## 4 Tensions Over a New Discipline

### 4.1 Applied Mathematics vs Data Processing

Operations research was excluded from AFCAL's scope, as it already had its own society. So was all work exclusively related to hardware, which belonged to the Société Française des Electriciens, the French member of IEEE, and could be published in a new journal, Automatisme. As for data processing, it was a set of office techniques remote from any scientific formalization. The association constituted the study of computation as a specific, hierarchized object, where calculating machines were scientific instruments applying branches of mathematics.

Since the very beginning of AFCAL, however, a tension had appeared between two conceptions of its scope: one centered on applied mathematics, the other pushing to include data processing. And of course this would grow along with the progressive diversification of the society in the following decade. In 1957 already, Carteron had written a letter to president Danjon on the definition of AFCAL's domain. Carteron recommended not to separate numerical computing and information processing: "For example, problems of sorting on magnetic tapes, or of machine translation of language, are part of our preoccupations, as well as the technical aspects of automatic data processing problems ${ }^{15}$ ". The fact that he mentioned these particular problems reveals that he was remarkably aware of recent research directions. There was also a sociological rationale in his proposal: with the advent of electronic computers, data processing would soon become too complex for traditional punch-card technicians; it would require

14 Journal du Congrès, No. 6, p. 2 (italics ours).
${ }^{15}$ J. Carteron's letter to A. Danjon, December 61957 (Archives de l'Observatoire de Paris, Ms 1061 II-2-D).
engineers trained at a higher level, who could be eligible among the advanced elite of AFCAL members.

At the other end of the spectrum, Kuntzmann, like other academic numerical analysts, wanted to involve the association in militantism for a reform of mathematical education, in order to give more room to applied mathematics. In several letters and reports, he advocated for a restructuring of mathematical categories, which would acknowledge the importance of "algorithmic mathematics". Even if he was soon to include a machine translation team in his institute, and courses on data-processing techniques in its curriculum, his priority always remained applied mathematics. Even decades later he kept considering informatique a technical application of mathematics.

In 1960, this debate became a central topic at AFCAL's annual assembly. The members who wanted to broaden the association's scope to data processing were led by Jean Carteron, Henri Boucher, a computer pioneer in the French Ministry of Defence, and Philippe Dreyfus, then a sort of sales evangelist for Bull computers who had already an international reputation. They succeeded in adding to the association's name the initials "TI" (traitement de l'information). Thus renamed, AFCALTI asserted the unity of a field broadened beyond applied mathematics, precisely at the time when it organized a joint conference with SOFRO (Operations Research) and the French chapter of The Institute of Management Sciences (TIMS). This change was still consolidated as Carteron was elected president as successor to Danjon. The journal's title became ChiffresRevue française de traitement de l'information.

### 4.2 Toward a New Scientific Discipline

At the 1962 annual assembly, a neologism, informatique, was defined by Philippe Dreyfus and Robert Lattès as the "technique of the logical and automatic processing of information, the support of human knowledge and communication" (see Ph. Dreyfus 1962). Note that various attempts had already been made to name the computer field, including Informatik in Germany as early as in 1957. Ph. Dreyfus later explained the invention of informatique by the simple need he felt to name his professional occupation, encompassing all types of applications of electronic computers ${ }^{16}$. But there was another reason: Dreyfus had just left Bull to create a software and service bureau firm, in partnership with Robert Lattès, a brilliant mathematician who worked simultaneously for the French atomic energy authority (Commissariat à l'Énergie Atomique) and for a mathematical consultancy firm, SEMA. Informatique was the brand name of their company, just as, at the same time across the Atlantic, Walter Bauer founded Informatics Inc.

According to the president of AFCALTI, Carteron, informatique covered "all disciplines covered by calculation and information processing, from the logical design of computers to numerical analysis through machine translation." (see Carteron 1963) Defined as a "technique" in 1962, informatique thus evolved into a set of "disciplines".

[^28]A further step was taken in the proceedings of the third congress of the association (CNRS 1963). In the brief introduction, the editors asserted that programming was no longer a mere means to communicate with the machine, but that it turned into "a science of languages extending from algebra to decision theory", allowing to define "the modes of expression as well as the construction of compilers [...]. Finally, considerations about logic allowed to monitor the evolution of ideas concerning theoretical informatique". Science, informatique théorique: these bold assertions were thus launched, with no explanatory development, likely to stimulate the discussion and to accelerate an evolution.

It is also a good case for a geography of science. The fact that this happened in Toulouse is not irrelevant. None in Grenoble would have written such provoking phrases under Kuntzmann's direction. The Toulouse computer scientists, by contrast, were mainly physicists, engineers or applied mathematicians coming from marginal positions. Contrary to Kuntzmann, they did not worry about their mathematical legitimacy, and felt more free to embrace a new academic identity. The next year, Dreyfus and Lattès hammered their message again. Aiming at a broader readership than AFCALTI members, they chose Le Monde to publish an article-manifesto: "A new discipline: Informatique" (Lattès and Dreyfus 1964). The context was favourable: several French universities had already created specialized degrees, IBM had just launched its "third generation" computers, operating systems were now a central issue for computer scientists, and the takeover of the French company Bull by General Electric had triggered reactions in industry and government spheres.

The headlines described informatique as a broad set of sciences and techniques, reflecting the variety of AFCALTI's interests. Then Dreyfus and Lattès recalled their definition of 1962, but assertively added a much more scientific dimension: "More than a simple technique, computing is a discipline, a scienceinterface (une science-carrefour) covering a large and disparate sector as much technical as scientific; it is also a mental attitude in approaching problems." They insisted that computing had become a fully-fledged research sector, with specific matters of its own "such as software and certain technologies". In conclusion the article stressed the growing importance of computing for society and economy, which required its integration in all parts of higher education and executive training, "to raise generations of men adapted to the mode of thought now required by the machine."

By that time, informatique was being gradually adopted in French common vocabulary, coexisting with other terms, and triggering discussions (within and outside the association) about its content, its meaning, its position between science and technology. The controversy opposed mainly those who thought in terms of applied mathematics and those who asserted the central place now occupied by computers in different activities. Branch or application of mathematics? Or technique, or even science of computers? or of information? The initial definition given in 1962 was tentative, and anyone was entitled to propose variants of it. This debate would eventually continue over several decades,
drawing a dividing line within academic spheres, and shaping its institutional organisation.

### 4.3 The Evolving Functions of the Association and the Institutionalization of Computing

Finally, we examine how the definition of the functions of the association evolved over time, and draw lessons on the emergence of Computing as a new professional field in France.

The association had multiple functions from the start. Its founders knew that their field and its economic impact were growing fast. One of their initial goals was to increase awareness of the potential applications of computation among institutional actors; some members such as J. Kuntzmann also wanted to involve the association in their campaign for a reform of mathematical education, which would give more room to applied mathematics and numerical computation.

But the association was initially dominated by academics and engineers, and its primary goal was research. A key issue in the definition of its function was thus the definition of its scientific object. This definition underwent important, and somewhat tortuous evolutions during its first ten years. In the beginning, it was dedicated to computation, its methods and the relevant mathematical developments. Analog computation was not excluded, but numerical computation was central.

But as we have seen (Sect.4.1), as computing applications kept growing, a more applied point of view gained traction within the association, and it became more and more inclusive. This expansion was justified by the common interests shared by the different specialists, and by the extremely rapid growth of computer applications. This rate of growth helps to explain the frantic pace of institutional evolution, as the association underwent three amalgamations in less than a decade.

If the term "amalgamation" is adequate at an institutional level, it is nevertheless misleading at the level of research practices. The series of institutional amalgamations did not erase the differences between the various intellectual traditions involved. Considering its scientific object, the evolution of the association might be better described as a movement of "expansion-subdivision". In its March-April 1967 issue, the Chiffres journal was divided in two collections, one for theory and methodology, the other for practice. In 1968 it was further divided into three different collections: mathematics, computer science, operations research. In the 70s, the subdivision process continued, as the association itself was subdivided into subsections and work groups (Brygoo 1988). The association was thus a place where different subdisciplines would interact, not a melting pot where those subdisciplines would lose their original identities.

The association general assembly, and Chiffres were also places where informatique was proposed and discussed as a disciplinary term. Less successful neologisms, such as the 'economic cybernetics' (cybernétique économique), reveal the fluidity of the new professional domain, and the efforts made to cope with it. As Gingras (1991) put it, scientific associations are a privileged place for
specialists of a given domain to articulate a discourse on their activity and intellectual identity, both for themselves and for outsiders. Terminological innovation is thus a way to define and defend the identity of a new group of specialists. The association was thus one of the institutions where the professional identity of informaticien emerged.

To conclude, let us use this story to test the idea that the creation of scientific associations can be a criterium for the emergence of new disciplines. When it was renamed AFIRO, the association was a scientific institution at least including computer science as part of its object. According to this criterium, it played a part in the institutionalization of computer science. But a couple of remarks should be made in order to understand more precisely this institutionalization process. We must distinguish between disciplines qua intellectual tradition and disciplines qua institution. As Gayon (2015) remarked, an intellectual tradition like evolutionary theory could exist for decades without being institutionalized, and natural history is still an academic institution without being an autonomous intellectual tradition anymore. Furthermore, for an association to be the institutionalization of intellectual tradition X, the intellectual tradition X has to be identified as such by actors. In a typical scenario, an intellectual tradition would be formed before it would be equipped with institutions such as scientific societies.

But at the beginning of the AFCAL, it was not obvious that computation, even if its fast development demanded an institutional response, would represent a distinct intellectual tradition, other than just a branch of applied mathematics. Second, even when the intellectual tradition of computer science started to emerge, it did not lead the association to focus exclusively on computer science quite the contrary. The word informatique was included in its title precisely when the scope of the association was broadened to include recherche opérationnelle (operations research), then automatics and control. The association thus defined and covered a much broader field of expertise and applications than the academic discipline of computer science in France would in the near future, since Operations Research and Automatics have autonomous academic institutionalizations outside of computer science departments.

Far from being just an institutionalization of an already existing academic tradition, the association was a forum where, through collaborative work, tensions over agenda and terminological debate, the conscience of the existence of a new discipline, computer science, was going to emerge. Using the existence of the association as a static criterium for the institutionalization of an academic tradition masks its dynamic role in the very constitution of this intellectual tradition.

## 5 Conclusion

The creation of the AFCAL, and much of its evolution, was application driven. The huge expansion of computer applications in the first years of its existence is the driving force behind the growth of its membership, and the fluidity of its
scope. It can be also seen as a decisive factor in the definition of the various association functions. Spreading the knowledge about potential applications of the new technologies among institutional actors, reforming the mathematical education, gathering specialists coming from different domains around a new research object: all these actions can be understood as a fast institutional response to a rapidly evolving field of high socio-economical impact.

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# Computable Reductions and Reverse Mathematics 

Reed Solomon ${ }^{(\boxtimes)}$<br>University of Connecticut, Storrs, CT 06268, USA<br>david.solomon@uconn.edu


#### Abstract

Recent work in reverse mathematics on combinatorial principles below Ramsey's theorem for pairs has made use of a variety of computable reductions to give a finer analysis of the relationships between these principles. We use three concrete examples to illustrate this work, survey the known results and give new negative results concerning $\mathrm{RT}_{k}^{1}$, $\mathrm{SRT}_{\ell}^{2}$ and COH . Motivated by these examples, we introduce several variations of ADS and describe the relationships between these principles under Weihrauch and strong Weihrauch reductions.


## 1 Introduction

The general project of reverse mathematics is to formalize mathematical theorems in second order arithmetic and to determine which set-theoretic axioms are required to prove them. Typically, we take $\mathrm{RCA}_{0}$, which includes the noninduction axioms of PA as well as the $\Delta_{1}^{0}$-comprehension and $\Sigma_{1}^{0}$-induction schemes, to be the base subsystem of second order arithmetic over which we prove our equivalences. $\mathrm{RCA}_{0}$ has the advantage that it allows a reasonable amount of coding and that theorems provable in $\mathrm{RCA}_{0}$ roughly correspond to theorems for which existential objects can be found computably in the parameters. More specifically, an $\omega$-model satisfies $\mathrm{RCA}_{0}$ if and only if the second order part of the model is closed under Turing reducibility and the Turing join. The $\omega$-models of other subsystems of second order arithmetic can be described in similarly natural computability-theoretic terms. These connections allow us to prove results in reverse mathematics using techniques and intuitions from computability theory.

While we will primarily be concerned with $\omega$-models, it is worth noting the importance of non- $\omega$-models. They play a crucial role in, for example, measuring levels of induction or conservation. More importantly for our purposes, they are used to separate Ramsey's theorem for pairs $\left(\mathrm{RT}_{2}^{2}\right)$ from its stable version $\left(\mathrm{SRT}_{2}^{2}\right)$. In the next section, we will define $\mathrm{RT}_{2}^{2}$ and $\mathrm{SRT}_{2}^{2}$ precisely, but for now we use them to make a motivational point. From a computability standpoint, the natural first step to separate these principles would be to prove that for every $\Delta_{2}^{0}$ set $D$, either $D$ or $\bar{D}$ contains an infinite low set. Unfortunately, this statement

[^29]is false (see [6]), so this method fails at its initial step. However, Chong et al. [4] constructed a non- $\omega$-model of $\mathrm{RCA}_{0}$ in which this property holds and they used this model to prove that $\mathrm{SRT}_{2}^{2}$ does not imply $\mathrm{RT}_{2}^{2}$ over $\mathrm{RCA}_{0}$.

One of the underlying motivations for the work presented here is the question of whether $\mathrm{SRT}_{2}^{2}$ and $\mathrm{RT}_{2}^{2}$ can be separated by an $\omega$-model. Because the fundamental property of the Chong, Slaman and Yang model is false in any $\omega$-model of $\mathrm{RCA}_{0}$, it appears that new ideas are needed. There are a range of tools from proof theory and computability theory to compare the strengths of theorems and many of these tools give a finer analysis than the analysis given by provable equivalence over $\mathrm{RCA}_{0}$. In this article, we will use several concrete examples to survey recent results in this direction using reductions which are more commonly used in computable analysis.

We begin by restricting our attention to $\omega$-models and to combinatorial principles which have the $\Pi_{2}^{1}$ form

$$
\forall X(\Phi(X) \rightarrow \exists Y \Psi(X, Y))
$$

where $\Phi(X)$ and $\Psi(X, Y)$ are arithmetic. We refer to a statement $P$ of this form as a problem and we refer to sets $X$ which satisfy $\Phi(X)$ as instances of $P$. Given an instance $X$ of $P$, we call a witness $Y$ such that $\Psi(X, Y)$ a solution to $X$.

Given two problems $P$ and $Q$ of this form, consider how we might show that $P$ holds in every $\omega$-model of $\mathrm{RCA}_{0}+Q$. Fix an $\omega$-model of $\mathrm{RCA}_{0}+Q$ and recall that the second order part of this model is a Turing ideal, so it is closed under Turing reducibility and the Turing join. We fix an instance $X$ of $P$ in this ideal and try to construct an instance $\widehat{X}$ of $Q$ which is also in the ideal. The simplest way to ensure that $\widehat{X}$ is in the ideal is to make it computable from $X$. Because $\widehat{X}$ is in the ideal, it must have a solution $\widehat{Y}$ in the ideal and, because the ideal is closed under the join operation, $\widehat{Y} \oplus X$ is also in the ideal. If we are lucky, we can use $\widehat{Y} \oplus X$ to compute a solution $Y$ to $X$ which, because the ideal is closed under Turing reducibility, will also be in the ideal.

Many proofs of implications $Q \rightarrow P$ in $\mathrm{RCA}_{0}$ proceed in this manner. However, it is worth noting that many variations are possible. For example, the proof could be more complicated in the sense that it requires us to solve several (possibly nested) instances of $Q$ before computing a solution to $X$. On the other hand, the proof could be simpler in the sense that the solution $\widehat{Y}$ to $\widehat{X}$ might already be a solution to $X$. More generally, $\widehat{Y}$ might compute a solution to $X$ without needing to reference the original instance $X$ of $P$.

The following definition gives a framework in which we can begin to address these finer questions about exactly how the proof of $P$ from $Q$ (over $\omega$-models of $\mathrm{RCA}_{0}$ ) proceeds. In the next section, we will illustrate these reductions with a number of concrete examples. Hirschfeldt and Jockusch [9] give many additional examples as well as more general reduction procedures to extend this type of analysis.

Definition 1. Let $P$ and $Q$ be problems.

1. $P$ is $\omega$-reducible to $Q$, denoted $P \leq_{\omega} Q$, if every $\omega$-model of $\mathrm{RCA}_{0}+Q$ is a model of $P$. (This reducibility has also been called computable entailment in a number of earlier articles).
2. $P$ is computably reducible to $Q$, denoted $P \leq_{c} Q$, if every instance $X$ of $P$ computes an instance $\widehat{X}$ of $Q$ such that if $\widehat{Y}$ is any solution to $\widehat{X}$, then there is a solution $Y$ to $X$ computable from $X \oplus \widehat{Y}$.
3. $P$ is strongly computably reducible to $Q$, denoted $P \leq_{s c} Q$, if every instance $X$ of $P$ computes an instance $\widehat{X}$ of $Q$ such that if $\widehat{Y}$ is any solution to $\widehat{X}$, then there is a solution $Y$ to $X$ computable from $\widehat{Y}$.
4. $P$ is Weihrauch reducible to $Q$, denoted $P \leq_{W} Q$, if there are Turing functionals $\Phi$ and $\Delta$ such that if $X$ is an instance of $P$, then $\Phi^{X}$ is an instance of $Q$ and if $\widehat{Y}$ is any solution to $\Phi^{X}$, then $\Delta^{X \oplus \widehat{Y}}$ is a solution to $X$.
5. $P$ is strongly Weihrauch reducible to $Q$, denoted $P \leq_{s W} Q$, if there are Turing functionals $\Phi$ and $\Delta$ such that if $X$ is an instance of $P$, then $\Phi^{X}$ is an instance of $Q$ and if $\widehat{Y}$ is any solution to $\Phi^{X}$, then $\Delta^{\widehat{Y}}$ is a solution to $X$.

It is straightforward to see that the following implications hold between these reductions. Hirschfeldt and Jockusch [9] prove that none of the given arrows reverse.


## 2 Three Examples

In this section, we give three examples of relationships between principles below Ramsey's theorem for pairs to illustrate these reducibilities. To fix notation, let $[\omega]^{2}$ denote the set of pairs $\langle x, y\rangle$ with $x<y$. We frequently use $k$ to denote the set $\{0, \ldots, k-1\}$.

Definition 2. A $k$-coloring of $[\omega]^{2}$ is a function $c:[\omega]^{2} \rightarrow k$ and we write $c(x, y)$ in place of $c(\langle x, y\rangle)$. We say that $c$ is stable if for every $x$, there is a color $i<k$ such that for every sufficiently large $y, c(x, y)=i$. That is, for every $x, \lim _{y} c(x, y)$ exists. For a stable coloring $c$, we say that $x$ has limit color $i$ if $\lim _{y} c(x, y)=i$.

Given a coloring $c:[\omega]^{2} \rightarrow k$, we say that $H \subseteq \omega$ is homogeneous if there is a color $i$ such that $c(x, y)=i$ for all $x, y \in H$ with $x<y$. In this case, we say that $H$ is homogeneous for color $i$. Similarly, if $c$ is a stable coloring, we say that $H$ is limit homogeneous if there is a color $i$ such that $\lim _{y} c(x, y)=i$ for all
$x \in H$. When we are dealing with more than one coloring, we will refer to sets which are homogenous for the coloring c as c-homogeneous sets.

Definition 3. We define the following versions of Ramsey's theorem.

- Ramsey's theorem for pairs $\left(\mathrm{RT}_{2}^{2}\right)$. Every coloring $c:[\omega]^{2} \rightarrow 2$ has an infinite homogeneous set.
- Stable Ramsey's theorem for pairs $\left(\mathrm{SRT}_{2}^{2}\right)$. Every stable coloring c : $[\omega]^{2} \rightarrow 2$ has an infinite homogeneous set.
- Limit Ramsey's theorem for pairs $\left(\mathrm{D}_{2}^{2}\right)$. Every stable coloring $c:[\omega]^{2} \rightarrow 2$ has an infinite limit homogeneous set.
- Ramsey's theorem for singletons ( $\mathrm{RT}_{k}^{1}$ ). Every coloring $c: \omega \rightarrow k$ has an infinite homogeneous (i.e. monochromatic) set. (This principle is often called the pigeonhole principle.)

Each of these statements can be written in the $\Pi_{2}^{1}$ form given above. In each case, the instances of the problem are the colorings of the appropriate type and the solutions to a given problem $c$ are the infinite homogeneous sets of the appropriate type. We introduce one final principle which will be used in our third example.

Definition 4. A set $Y$ is cohesive for a sequence $\left\langle X_{n} \mid n \in \omega\right\rangle$ of subsets of $\omega$ if for every $n$, either $Y \cap X_{n}$ or $Y \cap \overline{X_{n}}$ is finite.

- Cohesive principle (COH). For every sequence $\left\langle X_{n} \mid n \in \omega\right\rangle$ of subsets of $\omega$, there is an infinite cohesive set $Y$.

Example 1. We consider the principles $\mathrm{D}_{2}^{2}$ and $\mathrm{SRT}_{2}^{2}$. $\mathrm{D}_{2}^{2}$ arises naturally in computability theory by considering a set $D \leq_{T} 0^{\prime}$ with a fixed $\Delta_{2}^{0}$ approximation $f(x, s)$. By restricting the domain of $f(x, s)$ to values $x<s$, we can view $f$ as a 2 -coloring of $[\omega]^{2}$. The infinite limit homogeneous sets of this coloring are exactly the infinite subsets of $D$ and of $\bar{D}$.

To determine the relationship between $\mathrm{SRT}_{2}^{2}$ and $\mathrm{D}_{2}^{2}$ under these reductions, notice that for a stable coloring $c:[\omega]^{2} \rightarrow 2$, every homogeneous set is limit homogeneous (but not conversely). It follows that $\mathrm{D}_{2}^{2} \leq_{s W} \mathrm{SRT}_{2}^{2}$ by letting the functionals $\Phi$ and $\Delta$ be the identity. That is, given an instance $c$ of $\mathrm{D}_{2}^{2}$, we map $c$ to itself but view it as an instance of $\mathrm{SRT}_{2}^{2}$. From any solution $H$ to $c$ as an instance of $\mathrm{SRT}_{2}^{2}$ (i.e. an infinite homogeneous set $H$ ), we map $H$ to itself but view it as a solution to the original instance $c$ of $D_{2}^{2}$ (i.e. an infinite limit homogeneous set).

The non-trivial direction is to use $\mathrm{D}_{2}^{2}$ to solve instances of $\mathrm{SRT}_{2}^{2}$. A proof that $R C A_{0}+D_{2}^{2}$ implies $S R T_{2}^{2}$ over $\omega$-models goes as follows. Fix a stable coloring $c:[\omega]^{2} \rightarrow 2$ as an instance of $\mathrm{SRT}_{2}^{2}$. We regard $c$ as an instance of $\mathrm{D}_{2}^{2}$ and obtain an infinite limit homogeneous set $L$ of color $i$. To produce a homogeneous set of color $i$, we thin $L$ to $H=\left\{h_{0}, h_{1}, \ldots\right\}$. Let $h_{0}$ be the least element of $L$. Having defined $h_{n}$, let $h_{n+1}$ be the least element of $L$ such that $h_{n}<h_{n+1}$ and $c\left(h_{j}, h_{n+1}\right)=i$ for all $j \leq n$. (This proof works more generally in $\mathrm{RCA}_{0}+\mathrm{D}_{2}^{2}$,
but there is an induction issue. It uses $B \Sigma_{2}^{0}$ and hence relies on the fact that $\mathrm{D}_{2}^{2}$ implies $B \Sigma_{2}^{0}$ as shown by Chong et al. [3].)

This proof shows that $\mathrm{SRT}_{2}^{2} \leq_{c} \mathrm{D}_{2}^{2}$. From an instance $c$ of $\mathrm{SRT}_{2}^{2}$, we compute an instance of $D_{2}^{2}$ (via the identity function) and then use an arbitrary $D_{2}^{2}$ solution $L$ together with the original $\mathrm{SRT}_{2}^{2}$ coloring $c$ to compute an $\mathrm{SRT}_{2}^{2}$ solution $H$ to $c$. However, notice that the computation of $H$ from $L$ is non-uniform because it depends on knowing the limit color for $L$ and that the thinning process uses the original $\mathrm{SRT}_{2}^{2}$ instance $c$. Thus this proof leaves open the questions of whether $\mathrm{SRT}_{2}^{2} \leq_{W} \mathrm{D}_{2}^{2}$ (can the proof be made uniform?) and whether $\mathrm{SRT}_{2}^{2} \leq_{s c} \mathrm{D}_{2}^{2}$ (is the reference to the original coloring $c$ necessary to compute $H$ from $L$ ?). Both of these reductions fail.

Theorem 5 (Dzhafarov [7]). $\mathrm{SRT}_{2}^{2} \not \mathbb{Z}_{W} \mathrm{D}_{2}^{2}$ and $\mathrm{SRT}_{2}^{2} \not \leq_{s c} \mathrm{D}_{2}^{2}$.
The proof of Theorem 5 uses a generalization of Seetapun's method of cone avoidance for $\mathrm{RT}_{2}^{2}$ and in fact shows something stronger. Let $\mathrm{D}_{<\infty}^{2}$ be the problem for which the instances are given by pairs $\langle c, k\rangle$ where $c:[\omega]^{2} \rightarrow k$ is a stable coloring and the solutions to $\langle c, k\rangle$ are the infinite limit homogeneous sets. The stronger result from Dzhafarov [7] is that $\mathrm{SRT}_{2}^{2} \not \mathbb{Z}_{W} \mathrm{D}_{<\infty}^{2}$ and $\mathrm{SRT}_{2}^{2} \not \mathbb{Z s c}_{s c} \mathrm{D}_{<\infty}^{2}$.
Example 2. We consider the principles $\mathrm{RT}_{k}^{1}$ for varying numbers of colors. These principles differ in a significant way from those in the first example. Given $c: \omega \rightarrow$ $k$, there is an infinite homogeneous set $H$ computable from $c$ by non-uniformly fixing a color $i$ for which there are infinitely many $x$ such that $c(x)=i$. Therefore, when comparing $\mathrm{RT}_{k}^{1}$ and $\mathrm{RT}_{\ell}^{1}$, we immediately have $\mathrm{RT}_{k}^{1} \leq_{c} \mathrm{RT}_{\ell}^{1}$ because $\leq_{c}$ allows non-uniformity and access to the original $\mathrm{RT}_{k}^{1}$ coloring. Furthermore, if $k \leq \ell$, then every $k$-coloring is an $\ell$-coloring and we immediately have $\mathrm{RT}_{k}^{1} \leq_{s W}$ $\mathrm{R} \mathrm{T}_{\ell}^{1}$. Therefore, we only consider potential reductions from $\mathrm{RT}_{k}^{1}$ to $\mathrm{RT}_{\ell}^{1}{ }^{k}$ when $\ell<k$ and the reduction is stronger than $\leq_{c}$.

It is instructive to consider a combinatorial proof of $\mathrm{RT}_{3}^{1}$ from $\mathrm{RT}_{2}^{1}$. Fix a coloring $c: \omega \rightarrow 3$ and use it to define a coloring $d: \omega \rightarrow 2$ by $d(x)=0$ if $c(x)=0$ or $c(x)=1$ and $d(x)=1$ if $c(x)=2$. Let $H$ be an infinite $d$-homogeneous set. If $H$ is $d$-homogeneous for color 1 , then it is a $c$-homogeneous set for color 2 and we are done. On the other hand, if $H$ is $d$-homogeneous for color 0 , then we need another application of $\mathrm{RT}_{2}^{1}$ to get our $c$-homogeneous set. Let $d_{0}$ be the restriction of $d$ to $H$. Applying $\mathrm{RT}_{2}^{1}$ to $d_{0}$ gives an infinite $d_{0}$-homogeneous set $H_{0} \subseteq H$ which is clearly $c$-homogeneous as well. This proof does not use the original coloring to thin out a homogeneous set, but it does have a split between cases (so is non-uniform) and it potentially uses two applications of $\mathrm{RT}_{2}^{1}$ to solve a single instance of $\mathrm{RT}_{3}^{1}$. The following theorem collects a number of recent results related to the connection between these principles.

Theorem 6. The following negative results hold for all $0<\ell<k$.
$-\mathrm{RT}_{k}^{1} \not Z_{s W} \mathrm{RT}_{\ell}^{1}$ (Dorais et al. [5]).

- $\mathrm{RT}_{k}^{1} \not \Sigma_{W} \mathrm{RT}_{\ell}^{1}$ (independently by Hirschfeldt and Jockusch [9], by Patey [11] and by Rakotoniaina [12]).
$-\mathrm{RT}_{k}^{1} \not \mathbf{Z s c}_{s c} \mathrm{RT}_{\ell}^{1}$ (Dzhafarov [7]).

The most interesting result concerning these principles comes from refining the forcing techniques to remove any computable dependence between the instances of $\mathrm{RT}_{k}^{1}$ and $\mathrm{RT}_{\ell}^{1}$.

Theorem 7 (independently by Hirschfeldt and Jockusch [9] and by Patey [11]). For any $\ell<k$, there is a coloring $c: \omega \rightarrow k$ such that every coloring $d: \omega \rightarrow \ell$ (computable from $c$ or not) has an infinite homogeneous set $H$ which does not compute any infinite c-homogeneous set.

The proof of Theorem 7 given in Patey [11] shows that the coloring $c: \omega \rightarrow k$ can even be made low.

Example 3. We consider the relationship between COH and the principles above. The interest in these relationships is partly motivated by the fact that $R C A_{0}$ proves the equivalence of $\mathrm{RT}_{2}^{2}$ and $\mathrm{SRT}_{2}^{2}+\mathrm{COH}$. Cholak et al. [2] introduced this equivalence as an important tool for constructing solutions to instances of $\mathrm{RT}_{2}^{2}$ in two steps by using COH to pass from a general 2 -coloring to a stable 2 -coloring and then using $\mathrm{SRT}_{2}^{2}$ to solve the stable coloring. Therefore, we would like to know whether COH is reducible (by any of these reductions) to $\mathrm{SRT}_{k}^{2}$ for some $k \geq 2$ or to $\mathrm{SRT}_{<\infty}^{2}$. Partial answers to these questions are given by the following theorem.

Theorem 8 (Dzhafarov [7]). $\mathrm{COH} \not \leq{ }_{W} \mathrm{SRT}_{<\infty}^{2}$ and $\mathrm{COH} \not \leq_{s c} \mathrm{SRT}_{2}^{2}$.
Dzhafarov's proof that $\mathrm{COH} \mathbb{Z}_{s c} \mathrm{SRT}_{2}^{2}$ uses a new tree labeling method for constructing solutions to $\mathrm{SRT}_{2}^{2}$. The version of this method given in [7] is closely tied to coloring pairs and it was left as an open question whether this result could be extended to larger exponents. We will return to this question below.

A second motivation for studying the relationships between COH and other Ramsey principles is that COH can be recast as a sequential version of Ramsey's Theorem for Singletons in which the solution is allowed to make finitely many errors. We say that $Y$ is almost homogeneous for a coloring $c: \omega \rightarrow k$ if there is a finite set $F$ such that $Y-F$ is homogeneous for $c$. Under a suitable coding, COH is $s W$-equivalent to the statement that for every sequence $\left\langle c_{k} \mid k \geq 1\right\rangle$ of colorings $c_{k}: \omega \rightarrow k$, there is an infinite set $Y$ such that $Y$ is almost homogeneous for every coloring $c_{k}$. (For the details of this coding, see [8].) From this perspective, we would like to understand the precise relationship between $\mathrm{RT}_{k}^{1}$ and $\mathrm{SRT}_{\ell}^{2}$ as a stepping stone to determine whether COH can be reduced (by $\leq_{s c}, \leq_{c}$ or $\leq_{\omega}$ ) to $\mathrm{SRT}_{\ell}^{2}$ or to $\mathrm{SRT}_{<\infty}^{2}$. As above, $\mathrm{RT}_{k}^{1} \leq_{c} \mathrm{SRT}_{\ell}^{2}$ for any $k$ and $\ell$ because $\leq_{c}$ allows access to the $\mathrm{RT}_{k}^{1}$ coloring, and it is not hard to see that $\mathrm{RT}_{k}^{1} \leq_{s W} \mathrm{SRT}_{\ell}^{2}$ whenever $k \leq \ell$. Therefore, our interest is in comparing $\mathrm{RT}_{k}^{1}$ and $\mathrm{SRT}_{\ell}^{2}$ when $\ell<k$. The following theorem answers this question for $W$-reducibility.

Theorem 9 (Hirschfeldt and Jockusch [9]). For all $\ell<k, \mathrm{RT}_{k}^{1} \not Z_{W} \mathrm{SRT}_{\ell}^{2}$.
In recent work, we were able to answer a number of the questions left open by the results above. Our main result is to show that $\mathrm{RT}_{k}^{1} \not \mathbb{Z}_{s c} \mathrm{SRT}_{\ell}^{2}$ when $\ell<k$ in a strong form analogous to Theorem 7.

Theorem 10 (Dzhafarov et al. [8]). For all $\ell<k$, there is a coloring $c$ : $\omega \rightarrow k$ such that every stable coloring $d:[\omega]^{2} \rightarrow \ell$ has an infinite homogeneous set which does not compute an infinite c-homogeneous set.

The technique for proving this theorem involves an extension and a simplification of the tree labeling method used to build solutions to $\mathrm{SRT}_{2}^{2}$ in Dzhafarov's proof of $\mathrm{COH} \not \mathbb{Z}_{s c} \mathrm{SRT}_{2}^{2}$ in Theorem 8. It remains an open questions whether the coloring $c$ can be low as in Patey's proof of Theorem 7. There are two corollaries below to Theorem 10. The first corollary follows immediately and the second corollary follows by viewing COH as a sequential form of $\mathrm{RT}_{k}^{1}$ allowing finitely many errors.

Corollary 11 (Dzhafarov et al. [8]). For all $\ell<k, \mathrm{RT}_{k}^{1} \leq{ }_{s c} \mathrm{SRT}_{\ell}^{2}$.
Corollary 12 (Dzhafarov et al. [8]). $\mathrm{COH} \not \mathbb{Z}_{s c} \mathrm{SRT}_{<\infty}^{2}$.

## 3 Variations on ADS

In this section, we consider variations of the ascending/descending sequence principle ADS (defined below) which are motivated by the combinatorial relationships between $\mathrm{RT}_{2}^{2}, \mathrm{SRT}_{2}^{2}$ and $\mathrm{D}_{2}^{2}$. Throughout this section, when we say $\left(L, \leq_{L}\right)$ is a linear order, we assume that $L \subseteq \omega$. That is, we assume our algebraic structures are coded in the natural numbers. We use $\leq$ to denote the usual order on $\omega$.

Definition 13. Let $\left(L, \leq_{L}\right)$ be a linear order and let $S \subseteq L$. We say that $S$ is an

- ascending sequence if for all $x, y \in S, x \leq y$ implies $x \leq_{L} y$;
- descending sequence if for all $x, y \in S, x \leq y$ implies $x \geq_{L} y$;
- ascending chain if for all $x \in S$, there are only finitely many $y \in S$ with $y \leq_{L} x$;
- descending chain if for all $x \in S$, there are only finitely many $y \in S$ with $y \geq_{L} x$.

Note that if $S$ is an ascending (descending) chain, then $S$ is isomorphic to $\omega$ (or $\left.\omega^{*}\right)$. However, it is given as a chain in $L$ rather than as an ascending (descending) sequence in the sense that the elements of $S$ do not necessarily appear in ascending (descending) order when enumerated in increasing $\leq$-order.

Definition 14. We define the following principles.

- Ascending/descending sequence principle (ADS): Every infinite linear order has an infinite ascending or descending sequence.
- Ascending/descending chain principle (ADC): Every infinite linear order has an infinite ascending or descending chain.

There principles have the required $\Pi_{2}^{1}$ form given in the introduction and the combinatorial relationship between ADC and ADS is similar to the combinatorial relationship between $D_{2}^{2}$ and $S R T_{2}^{2}$. For example, $A D C$ and ADS have the same instances and an ADS-solution to ( $L, \leq_{L}$ ) is also an ADC-solution to ( $L, \leq_{L}$ ). Therefore, $\mathrm{ADC} \leq_{s W}$ ADS via identity functionals just as $\mathrm{D}_{2}^{2} \leq_{s W} \mathrm{SRT}_{2}^{2}$.

More importantly for our analogy, every ADC-solution $S$ to ( $L, \leq_{L}$ ) can be thinned out to an ADS-solution $H$ to $\left(L, \leq_{L}\right)$. To see why, suppose $S=$ $\left\{s_{0}, s_{1}, \ldots\right\}$ is an infinite ascending chain. We define an infinite ascending sequence $H=\left\{h_{0}, h_{1}, \ldots\right\}$ as follows. Let $h_{0}$ be the $\leq$-least element of $S$. Having defined $h_{n}$, let $h_{n+1}$ be the $\leq$-least element of $S$ such that $h_{n}<h_{n+1}$ and $h_{n}<_{L} h_{n+1}$. A similar argument works in the case when $S$ is a descending chain. As with $\mathrm{SRT}_{2}^{2}$ and $\mathrm{D}_{2}^{2}$, the computation of $H$ from $S$ is non-uniform (because it requires knowing whether $S$ is ascending or descending) and uses the ADSinstance $\left(L, \leq_{L}\right)$. Therefore, the thinning process shows ADS $\leq_{c}$ ADC exactly as with $\mathrm{SRT}_{2}^{2} \leq_{c} \mathrm{D}_{2}^{2}$. Before considering whether this reduction can be made uniform, we introduce two notions of stability in this context.

Definition 15. Let $\left(L, \leq_{L}\right)$ be a linear order. An element $x \in L$ is called small if there are only finitely many $y \in L$ such that $y<_{L} x$. Similarly, $x \in L$ is called large if there are only finitely many $y \in L$ such that $x<_{L} y$. We say that $L$ is stable if every $x \in L$ is either small or large.

If $L$ is stable, then $L$ is isomorphic to either $\omega+k, k+\omega^{*}$ or $\omega+\omega^{*}$. The established definition for the stable version SADS of ADS restricts the instances of the problem to linear orders isomorphic to $\omega+\omega^{*}$, that is, to stable linear orders with infinitely many small elements and infinitely many large elements. In the context of $\omega$-models of $\mathrm{RCA}_{0}$ or of $\leq_{c}$-reductions, neglecting to consider linear orders $L$ of type $\omega+k$ or $k+\omega^{*}$ does not matter because $L$ can (nonuniformly) compute an infinite ascending sequence (in the case of $\omega+k$ ) or descending sequence (in the case of $k+\omega^{*}$ ). In our analogy with $\mathrm{SRT}_{2}^{2}$, the orders $\omega+k$ and $k+\omega^{*}$ correspond to stable colorings $c:[\omega]^{2} \rightarrow 2$ in which there is a fixed color $i$ such that almost every $x$ has limit color $i$. However, in the context of uniform reductions, it is important not to discount these trivial cases. To account for these cases, we consider two stable versions of ADS and of ADC.

Definition 16. We define the following notions of stability for ADS and ADC.

- SADS: Every stable linear order $L$ with infinitely many small and large elements has an infinite ascending or descending sequence.
- SADC: Every stable linear order $L$ with infinitely many small and large elements has an infinite ascending or descending chain.
- General-ADS: Every stable linear order L has an infinite ascending or descending sequence.
- General-ADC: Every stable linear order L has an infinite ascending or descending chain.

It is straightforward to check that these four principles are equivalent under $\leq_{c}$ and that the following relationships hold under $\leq_{s W}$.

- SADS $\leq_{s W}$ General-ADS $\leq_{s W}$ ADS.
- SADC $\leq_{s W}$ General-ADC $\leq_{s W}$ ADC.
- SADC $\leq_{s W}$ SADS.
- General-ADC $\leq_{s W}$ General-ADS.

Taken together with these reductions, the following theorem gives us a complete picture of the relationships between the six principles introduced in this section under $\leq_{W}$ and $\leq_{s W}$.

Theorem 17 (Astor et al. [1]). We have the following negative results concerning $\leq_{W}$.
(1) SADS $\not Z_{W}$ ADC.
(2) General-ADC $\mathbb{L}_{W}$ SADS.
(3) ADC $\not \mathbb{Z}_{W}$ General-ADS.

The first statement in Theorem 17 is proved by a Seetapun-style forcing construction while the second statement is proved with a simpler forcing notion. The last statement follows from the fact that computable instances of SADC have low solutions and that there are computable instances of ADC which have no low solutions (as shown in [10]).

The relationships between these six principles under $\leq_{W}$ and $\leq_{s W}$ are summarized by the following diagram. The arrows indicate an $\leq_{s W}$ reduction while the missing arrows indicate the failure of $\mathrm{a} \leq_{W}$ reduction.


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## Contributed Papers

# Busy Beavers and Kolmogorov Complexity 

Mikhail Andreev ${ }^{(\boxtimes)}$<br>Moscow Lomonosov State University, Moscow, Russia<br>amishaa@mail.ru


#### Abstract

The idea to find the "maximal number that can be named" can be traced back to Archimedes (see his Psammit [1]). From the viewpoint of computation theory the natural question is "which number can be described by at most $n$ bits"? This question led to the definition of the so-called "busy beaver" numbers (introduced by T. Rado). In this note we consider different versions of the busy beaver-like notions defined in terms of Kolmogorov complexity. We show that these versions differ depending on the version of complexity used (plain, prefix, or a priori complexities) and find out how these notions are related, providing matching lower and upper bounds.


## 1 Introduction

In 1962 Tibor Radó [5] suggested to consider, for each natural $n$, the maximal integer that can be printed by a terminating computation of a Turing machine that has at most $n$ states. The alphabet of the machine is assumed to be binary (blank and non-blank symbols). The machine starts on the empty tape and stops at some time. After that we count the number of non-blank symbols on the tape. Radó proved that this function grows faster that any computable function (of $n$ ). The same is true for other functions defined in a similar way (e.g., the maximal number of steps in a terminating computation of a machine with $n$ states on the empty tape, or the maximal shift of its working head). Still these definitions look too machine-dependent: even small changes in the model (say, allowing two tapes or one-sided tape) could give different (but still fast-growing) functions.

A more invariant approach becomes possible if we use the notions for algorithmic information theory (Kolmogorov complexity theory). We assume here that the reader is familiar with the basic notions of this theory (see, e.g., [7] or [4], or the short introduction in [6]). We consider the maximal number that has complexity at most $n$, i.e., the maximal number that is an output of some program of length at most $n$. Here we assume that the programming language is an optimal decompressor in the sense of algorithmic information theory (that leads to a minimal complexity function; see [7] or [4] for the formal definitions). It is easy to show (see, e.g., [7, Sect.1.2]) that we get the same function (up to $O(1)$-change in the argument) if we consider the maximal running time of the optimal decompressor on programs of length at most $n$. (The latter definition depends on the choice of interpreter for the optimal programming language and
the computation model used to define the running time, but for every choice we get the same function up to $O(1)$-change in the argument.)

In other words, we fix optimal (plain) decompressor $D$ and denote the complexity with respect of this decompressor $D$ by $C(\cdot)$ (the plain Kolmogorov complexity). Then $B(n)=\max \{N \mid C(N) \leqslant n\}$, so $B(n)$ is the maximum value of $D$ on arguments of length at most $n$ (we consider inputs as binary strings and outputs as natural numbers). Define $B B(n)$ as the maximum computation time for $D$ on the same inputs (for arbitrary fixed machine computing $D$ in arbitrary fixed computation model). As we have mentioned, the following statement holds: $B(n-c) \leqslant B B(n) \leqslant B(n+c)$ for some constant $c$ and for all $n$ (see [7]).

Additive constant in the argument is unavoidable, since the function $C(N)$ is defined only up to an $O(1)$ additive term (when you replace one optimal decompressor by another, an additive $O(1)$ term appears). So we will not distinguish $B(n)$ and $B B(n)$ and will use the notation $B(n)$ in the sequel for this plain busy beaver function.

One can repeat the same definitions for prefix-free decompressors and prefixfree Kolmogorov complexity (see [6, 7] for the definitions). We define the prefix busy beaver function $B P(n)=\max \{N \mid K(N) \leqslant n\}$.

Again one can consider the maximal computation time of an optimal prefixfree decompressor (as defined in [7, Sect.4.4]) on inputs of size at most $n$, and again we get two functions that are the same (up to an additive $O(1)$-term in the argument), for the same reasons. ${ }^{1}$

So we may forget about computation time, and consider the functions $B$ and $B P$ defined as explained above. We will compare the growth rate of the functions $B$ and $B P$ and show that these functions are different ( $B$ grows faster than $B P$ ). We also compare these functions with an intermediate function $B P^{\prime}$ that will be defined in terms of the a priori probability.

Let us first recall the definition of a priori probability. A priori probability $\mathbf{m}(k)$ of number $k$ can be defined as the $k$-th term of a maximal (up to multiplicative constant) converging lower semicomputable non-negative series. Levin showed that such a series exists, and proved that $\mathbf{m}(k)=2^{-K(k)+O(1)}$ (see, e.g., [7, Chap. 4] for the details). Now we consider the modulus of convergence for this series: for every $n$ we find minimal $N$ such that $\sum_{k>N} \mathbf{m}(k)<2^{-n}$. Denote this $N$ by $B P^{\prime}(n)$. The difference between $B P$ and $B P^{\prime}$ can be explained as follows: after $B P(n)$ all terms of the series $\sum \mathbf{m}(k)$ are small enough (less than $\left.2^{-n}\right)$, and after $B P^{\prime}(n)$ the tail of this series is small enough. Obviously $B P(n) \leqslant B P^{\prime}(n)$, or, more accurately, $B P(n) \leqslant B P^{\prime}(n+c)$ due to $O(1)$ additive

[^30]terms in both definitions (Kolmogorov complexity is defined up to an $O(1)$ additive term, and a priori probability is defined up to an $\Theta(1)$ factor).

This three functions share basic computational properties with classical busy beaver function: they are not computable and grow faster than any computable function. All three functions are computable with oracle for the halting problem (as well as the classical busy beaver function.

In this article we compare growth rates of these functions. Theorem 1 shows that all three functions are relatively close to each other: all three functions are equal up to at most $(1+\varepsilon) \log n$ argument shift. Theorem 2 shows that the bound provided by Theorem 1 is quite tight. For example, one cannot remove $\varepsilon$ from the previous statement: a gap greater than $\log n$ appears between $B P$ and $B P^{\prime}$ for some values of $n$, as well as between $B P^{\prime}$ and $B$ for some (other) values of $n$.

Theorem 1. (i) There exists a constant $c$ such that $B P(n) \leqslant B P^{\prime}(n+c)$ and $B P^{\prime}(n) \leqslant B(n+c)$ for all $n$.
(ii) There exists a constant $c$ such that $B(n) \leqslant B P(n+K(n)+c)$ for all $n$.
(iii) Let $\left(x_{n}, y_{n}\right)$ be a sequence of pairs of natural numbers such that $x_{n} \leqslant y_{n}$, the sequence $x_{n}$ is lower semicomputable, and the sequence $y_{n}$ is upper semicomputable. Assume that $\sum_{n} 2^{x_{n}-y_{n}}<+\infty$. Then there exists $c$ such that $B\left(x_{n}\right) \leqslant B P\left(y_{n}+c\right)$ for all $n$.

This theorem uses the notion of lower and upper semicomputable sequences. Recall that a sequence $y_{n}$ of real numbers is lower semicomputable if $y_{n}$ is a (point-wise) limit of some total computable non-decreasing (in $k$ ) rational-valued function of two arguments: $y_{n}=\lim _{k} y(n, k)$; upper semicomputability is defined in a symmetric way using non-increasing functions. If $y_{n}$ are natural numbers, the function $y(\cdot, \cdot)$ can be chosen in such a way that its values are also natural numbers, and convergence means that for each $n$ the equality $y_{n}=y(n, k)$ is true for all sufficiently large $k$. See [7] for the details. ${ }^{2}$

Items (i) and (ii) are rather simple, and (iii) is a more symmetric way to present (ii) (as we will see later). Note that (ii) is a special case of (iii) if we let $\left(x_{n}, y_{n}\right)=(n, n+K(n))$. Another special case of (iii) is obtained if we let $\left(x_{n}, y_{n}\right)=(n-K(n), n)$, so $B(n-K(n)) \leqslant B P(n+c)$ for some $c$ and all $n$.

The statement about $(1+\varepsilon) \log n$ mentioned above can be obtained as a corollary of (ii) since $K(n) \leqslant(1+\varepsilon) \log n$ for $\varepsilon>0$ (note that the series $\sum 2^{-(1+\varepsilon) \log n}=\sum\left(1 / n^{1+\varepsilon}\right)$ converges $)$.

Items (ii) and (iii) are not completely symmetric: why do we add $c$ to the right side, instead of subtracting it from the left side? We can formulate symmetric statements:
(ii') $B(n-K(n)-c) \leqslant B P(n)$ for some $c$ and all $n$;
(iii') Under the same assumptions as in (iii) we have $B\left(x_{n}-c\right) \leqslant B P\left(y_{n}\right)$ for some $c$ and all $n$.

These statements are also true; we will return to them after we prove Theorem 1 (they are easy corollaries of it).

[^31]The next results say that if $\sum_{n} 2^{x_{n}-y_{n}}=+\infty$ (for lower semicomputable $x_{n}$ and upper semicomputable $y_{n}$ ) then (iii) is not true anymore. Moreover, in this case a large gap may appear both between $B$ and $B P^{\prime}$ and between $B P^{\prime}$ and $B P$ (but in different places).

Theorem 2. Assume that $\left(x_{n}, y_{n}\right)$ is a sequence of different pairs of natural numbers, $x_{n} \leqslant y_{n}$, the sequence $x_{n}$ is enumerable from below, and the sequence $y_{n}$ is enumerable from above. Assume also that $\sum 2^{x_{n}-y_{n}}=+\infty$. In this case
(i) there exists $n$ such that $B\left(x_{n}\right)>B P^{\prime}\left(y_{n}\right)$;
(ii) there exists $n$ such that $B P^{\prime}\left(x_{n}\right)>B P\left(y_{n}\right)$.

There is no constant $c$ in this theorem (in contrast to the previous one), but one can easily put it on any side (or even both): changing all $x_{n}$ or all $y_{n}$ by an additive constant does not change the divergence condition. For example, it is true that for all $c$ there exists $n$ such that $B\left(x_{n}\right)>B P^{\prime}\left(y_{n}+c\right)$ or for all $c$ there exists $n$ such that $B P^{\prime}\left(x_{n}-c\right)>B P\left(y_{n}\right)$, and so on.

Using Theorems 1 and 2 one can easily deduce that for every upper semicomputable sequence $a_{n}$ the following six conditions are equivalent:

- $B P(n) \leqslant B P^{\prime}\left(n+a_{n}+c\right)$ for some $c$ and for all $n$;
- $B P^{\prime}(n) \leqslant B\left(n+a_{n}+c\right)$ for some $c$ and for all $n$;
- $B P(n) \leqslant B\left(n+a_{n}+c\right)$ for some $c$ and for all $n$;
- $B P\left(n-a_{n}\right) \leqslant B P^{\prime}(n+c)$ for some $c$ and for all $n$;
- $B P^{\prime}\left(n-a_{n}\right) \leqslant B(n+c)$ for some $c$ and for all $n$;
- $B P\left(n-a_{n}\right) \leqslant B(n+c)$ for some $c$ and for all $n$;

Moreover, all these conditions are equivalent to the condition $\sum 2^{-a_{n}}<+\infty$ (which, in its turn, is equivalent to $a_{n} \geqslant K(n)-O(1)$, see [7]).

The meaning of Theorems 1 and 2 can be explained as follows. In these results we compare slow-growing functions that are inverse to the functions $B, B P$, and $B P^{\prime}$. We show that they are equal up to $(1+\varepsilon)$ times the logarithm of their values, and that this $\varepsilon$ cannot be omitted: without it both inequalities between neighbor functions may be violated. As Theorem 1 shows, these big gaps cannot happen at the same places (otherwise the total gap between lowest and highest functions exceeds the upper bound).

Statement (ii) from Theorem 2 has been proven by Gács [3] for the case $\left(x_{n}, y_{n}\right)=\left(n-a_{n}, n\right)$ (and the general case may be derived as a consequence of this special one, as we will see later), so our main result is item (i) from Theorem 2. Still we provide all the proofs in the next section for uniformity and reader's convenience.

How can we modify our definitions? One can look at the maximal $N$ such that $C(N \mid n) \leqslant n$ or such that $K(N \mid n) \leqslant n$. But we do not get new notions in this way: this quantity is still equal to $B(n)$ up to a $O(1)$-change in the argument. Indeed, the conditional complexity $C(x \mid n)$ is bounded by the unconditional complexity $C(x)$; on the other hand, if $C(N \mid n)=n$, then the conditional program of length $n$ for $N$ may be considered as a conditional prefix-free program with the same condition $n$ (if $n$ is given as a condition, we know when to stop reading the
program of length $n$ ). Moreover, this program also can be used as unconditional program for $N$, since $n$ (its length) is determined by the program. In general, $K(x \mid C(x))=C(x \mid C(x))=C(x)$ (up to $\mathrm{O}(1)$ additive term), see [7].

To finish our introduction, let us mention that $B P^{\prime}$ can be equivalently defined as the modulus of convergence for computable non-negative series of rational numbers with Martin-Löf random sums.

Theorem 3. Let $\sum a_{n}$ be a computable series of rational non-negative numbers whose sum is Martin-Löf random. Let $N(\varepsilon)$ be the modulus of convergence of this series, i.e., the minimal value of $N$ such that $\sum_{n>N} a_{n}<\varepsilon$. Then $B P^{\prime}(n-c) \leqslant$ $N\left(2^{-n}\right) \leqslant B P^{\prime}(n+c)$ for some $c$ and all $n$.

The first inequality was proven in [2, Theorem 19], while the second one follows from the definition of the a priori probability (recall that $\mathbf{m}$ is bigger than any computable converging sequence, up to $O(1)$ factor). In [2] it was also shown that if $N(\varepsilon)$ is the modulus of convergence for some computable converging series $\sum a_{n}$ with non-negative terms, and $B P(n-c) \leqslant N\left(2^{-n}\right)$ for some $c$ and all $n$, then the same property holds for $B P^{\prime}$ (for a different value of $c$ ).

## 2 Upper Bounds

In this section we prove Theorem 1 .
(i) The inequality $B P(n) \leqslant B P^{\prime}(n+c)$ follows directly from definitions. If we define $\mathbf{m}(n)$ exactly as $2^{-K(n)}$, it is true even without $c$-term.

Now we prove that $B P^{\prime}(n) \leqslant B(n+c)$ for some $c$ and for all $n$. To do this, we construct an algorithm that, given $n$, enumerates at most $2^{n}$ different integers, and the last of them is bigger than $B P^{\prime}(n)$. The $n$-bit string that is the bit representation of the item's number in this enumeration, identifies the last number ( $n$ is known, being the length of this string), so we get the required inequality. How the enumeration algorithm works? This algorithm approximates all $\mathbf{m}(n)$ from below in parallel; we assume that at every moment only finitely many approximations are not zeros. As soon as the tail of the current approximation for $\mathbf{m}$, starting from the last enumerated integer, becomes greater than $2^{-n}$ (i.e., the current approximation to $B P^{\prime}$ exceeds the last enumerated integer) we enumerate a new integer that is bigger than all $k$ with non-zero current approximations for $\mathbf{m}(k)$. Obviously this cannot happen more than $2^{n}$ times: every time an integer is enumerated, we leave behind total $\mathbf{m}$-weight at least $2^{-n}$.
(ii) It is well known that $K(x) \leqslant C(x)+K(C(x))+O(1)$ (for example, see [7, Sect. 4.6]). The following slightly more general statement is also true: if $C(x) \leqslant n$, then $K(x) \leqslant n+K(n)+O(1)$. Let us prove it. Starting with a program for $x$ that has length at most $n$, we prepend a block of the form $0^{k} 1$ to it (this block is obviously self-delimited) making the total length exactly $n+2$. Then we prepend a self-delimited code for $n$ (of length $K(n)$ ), and the result is a self-delimited code for $x$ (decode $n$ first, then read exactly $n+2$ symbols, remove
$0^{k} 1$ leading block, then use $C$-decompressor). This generalisation immediately implies that $B(n) \leqslant B P(n+K(n)+c)$ for some $c$ and for all $n$.
(iii) We will show that this inequality is a consequence of (ii). We start by showing that we can assume $x_{n}$ and $y_{n}$ to be computable without loss of generality.

By assumption, the sequences $x_{n}$ [resp. $y_{n}$ ] are lower [resp. upper] semicomputable. For each $n$, consider a uniformly computable sequence of pairs $\left(x_{n}^{i}, y_{n}^{i}\right)$ of integers that monotonically converge to $\left(x_{n}, y_{n}\right)$ as $i \rightarrow \infty$. Combine arbitrarily all these sequences into one sequence, leaving only the first appearance of each pair (removing all duplicates). We get a computable sequence ( $\tilde{x}_{i}, \tilde{y}_{i}$ ); every pair $\left(x_{n}, y_{n}\right)$ appears in this sequence together with finitely many its approximations. Note that $\sum_{i} 2^{\tilde{x_{i}}-\tilde{y}_{i}}$ is at most two times bigger than $\sum_{n} 2^{x_{n}-y_{n}}$ : every time a new approximation for $x_{n}$ or $y_{n}$ appears, the respective term is the sum is increased by factor 2 or more, so the sum for $\tilde{x}_{i}, \tilde{y}_{i}$ is at most twice bigger than the original one, and if the original sum is finite, then the new one is also finite. Note also that the desired inequality for the new sequence implies the same inequality for the original sequence (that is a subsequence of the new one). So we can assume $x_{n}, y_{n}$ is computable without loss of generality.

Now assume that a computable sequence $\left(x_{i}, y_{i}\right)$ is given. Define $f(n)=$ $\min \left\{y_{i}-n \mid x_{i}=n\right\}$; if $n$ does not appear among $x_{i}$, the value $f(n)$ is $+\infty$. The function $f$ is upper semicomputable, and $\sum_{n} 2^{-f(n)}<+\infty$, since the pairs $(n, n+f(n))$ are guaranteed to appear among $\left(x_{i}, y_{i}\right)$. So $f(n) \geqslant K(n)-O(1)$. Therefore, $x_{i}+K\left(x_{i}\right) \leqslant y_{i}+O(1)$ for the pairs with minimal $y_{i}$ (for a given $\left.x_{i}\right)$ and therefore for all pairs. The function $B P$ increases, so we get $B\left(x_{i}\right) \leqslant$ $B P\left(x_{i}+K\left(x_{i}\right)+O(1)\right) \leqslant B P\left(y_{i}+O(1)\right)$ for all pairs. The claim (iii) is proven.

Symmetric results (mentioned above) are also easy to prove:
(ii') $B(n-K(n)-c) \leqslant B P(n)$ for some $c$ and for all $n$.
(iii') If $x_{n}$ and $y_{n}$ satisfy the same assumptions as in (iii), then $B\left(x_{n}-c\right) \leqslant$ $B P\left(y_{n}\right)$ for some $c$ and for all $n$.

To prove (ii') we use (ii) for a smaller argument: $B(n-K(n)-e) \leqslant$ $\leqslant B P(n-K(n)-e+K(n-K(n)-e)+c)$ holds for some $c$ and all $n, e$. Now we want to choose the constant $e$ in such a way that the argument in the right hand side is at most $n$ for all $n$ (recall that function $B P$ is monotone): $n-K(n)-e+K(n-K(n)-e)+c \leqslant n$.
Indeed, $K(n-K(n)-e) \leqslant K(n, K(n))+K(e)+O(1) \leqslant K(n)+K(e)+O(1)$, and $e-K(e)$ can be made arbitrary large for large enough $e$ (larger than sum of $O(1)$ terms in the inequalities).

To derive (iii') from (ii'), one can use the same technique as used to deduce (iii) from (ii). The only difference is that one should group pairs with the same $y_{i}$ (instead of $x_{i}$, as we did in the proof).

## 3 Lower Bounds

In this section we prove Theorem 2.

### 3.1 Proof of the Claim (i)

We have a sequence of different pairs $\left(x_{n}, y_{n}\right)$ of integers such that $x_{n} \leqslant y_{n}$. We assume that $x_{n}$ is lower semicomputable, $y_{n}$ is upper semicomputable and $2^{x_{n}-y_{n}}=+\infty$. We need to show that there exists $n$ such that $B\left(x_{n}\right)>B P^{\prime}\left(y_{n}\right)$.

First we will reduce this statement to its special case where $\left(x_{n}, y_{n}\right)=(n, n+$ $a_{n}$ ), and $a_{n}$ is some upper semicomputable sequence of natural numbers (the value $+\infty$ is also allowed).

For this reduction we use the same trick as in the previous section. First we replace $\left(x_{n}, y_{n}\right)$ by its approximations $\left(x_{n}^{i}, y_{n}^{i}\right)$, and then combine all these approximations into one computable sequence by removing the duplicates. The sum of $2^{x_{i}-y_{i}}$ may only increase (we add new elements), there are no duplicates (we removed them) and if $B\left(x_{n}^{i}\right)>B P^{\prime}\left(y_{n}^{i}\right)$ then $B\left(x_{n}\right)>B P^{\prime}\left(y_{n}\right)$ since we use monotone approximations and the busy beaver functions are monotone. So we may assume without loss of generality that the sequence $\left(x_{n}, y_{n}\right)$ is a computable sequence of different integer pairs.

Let $a_{n}=\min \left\{y_{i}-n \mid x_{i}=n\right\}$. The sequence $a_{n}$ is enumerable from above (since the sequence $\left(x_{i}, y_{i}\right)$ is computable). Note that $\sum_{n} 2^{-a_{n}} \geqslant \frac{1}{2} \sum_{i} 2^{x_{i}-y_{i}}$. Indeed, if we group pairs with $x_{i}=n$, the sum of this group is bounded by a geometric sequence with common ratio $1 / 2$, so the sum can be replaced by the maximal element (up to a 2 -factor). Therefore, $\sum_{n} 2^{-a_{n}}=+\infty$, and all pairs $\left(n, n+a_{n}\right)$ appear among $\left(x_{i}, y_{i}\right)$, so we get the desired reduction.

Now we use the following lemma: if $a_{n}$ is an upper semicomputable sequence of integers and $\sum_{n} 2^{-a_{n}}=+\infty$, there exists a computable sequence $\tilde{a}_{n} \geqslant a_{n}$ such that $\sum_{n} 2^{-\tilde{a}_{n}}=+\infty$. Indeed, we can approximate $a_{n}$ from above until some finite part of the series $\sum 2^{-a_{n}}$ exceeds 1 , then fix the current approximations for this part and call them $\tilde{a}_{n}$. Then the same argument is used for the tail, etc. This argument show that we may assume without loss of generality that $a_{n}$ is a computable sequence.

It remains to prove the following statement: if $a_{n}$ is a computable sequence of integers and $\sum 2^{-a_{n}}=+\infty$, then there exists $n$ such that $B(n)>B P^{\prime}\left(n+a_{n}\right)$. In other words, we need to show that there exists some $u$ such that $C(u) \leqslant n$ and $\sum_{i \geqslant u} \mathbf{m}(i)<2^{-n-a_{n}}$.

To prove an upper bound for $C(u)$, we need to construct a decompressor that provides a short description for $u$. However, this gives a bound with some additive constant term, so we need to construct a decompressor $D$ such that for every $d$ there exist $n$ and $u$ such that

$$
C_{D}(u) \leqslant n-d \quad \text { and } \quad \sum_{i \geqslant u} \mathbf{m}(i)<2^{n-a_{n}} .
$$

where $C_{D}(u)$ is the minimal length of $p$ such that $D(p)=u$.
To prove this, we use the game technique. Consider a game where Alice plays with Bob. They make alternating moves. Alice enumerates sets $D_{0}, D_{1}, \ldots$; at each move she adds finitely many integers to finitely many $D_{i}$ (so her move is a finite object). The set $D_{i}$ may contain at most $2^{i}$ elements. Bob approximates from below some sequence $\mu(0), \mu(1), \ldots$; initially all $\mu(i)$ are zeros, and at each
step Bob may increase finitely many of them by some rational numbers, but the sum $\sum \mu(i)$ should not exceed 1.

Assuming that both players respect the rules, Alice wins if (for limit values of $D_{i}$ and $\left.\mu(i)\right)$ for every $d$ there exists $n$ and $u$ such that $u \in D_{n-d}$ and $\sum_{i \geqslant u} \mu(i)<2^{-n-a_{n}}$. One may reformulate this statement eliminating $u$ : for every $d$ there exist $n$ such that

$$
\begin{equation*}
\sum_{i \geqslant \max \left(D_{n-d}\right)} \mu(i)<2^{-n-a_{n}} . \tag{*}
\end{equation*}
$$

We will prove that Alice has a computable winning strategy in this game. This implies the desired result. Indeed, we may let Alice use this strategy against the "blind" strategy of Bob that approximates from below the a priori probability function $\mu(i)=\mathbf{m}(i)$. Then the behavior of Alice is computable, the sets $D_{i}$ are enumerable and we construct a decompressor $D$ that maps $k$-bit string $p$ into $p$ th element in the enumeration of $D_{k}$ (in the last sentence binary string $p$ is identified with an integer it represents in the binary notation). This decompressor has the required property.

So why Alice has a computable strategy? She should guarantee the existence of a suitable $n$ for each $d$. This is done independently for each $d$; Alice chooses for each $d$ some interval $\left[l_{d}, r_{d}\right]$ where $n$ with the required properties exist. This intervals are chosen in such a way that there are no collisions (for different $d$ the values of $n-d$ cannot be the same, i.e., the intervals $\left[l_{d}-d, r_{d}-d\right]$ are disjoint). The intervals should be large enough: the sum of $2^{-a_{n}}$ over $n$ in $\left[l_{d}, r_{d}\right]$ should exceed $2^{d+1}$ (we will see that this is enough for our purposes). Since we assume that $a_{n}$ is a computable sequence and $\sum 2^{-a_{n}}=+\infty$, we can choose $\left[l_{d}, r_{d}\right]$ in a computable way.

How Alice constructs $D_{n-d}$ for $n \in\left[l_{d}, r_{d}\right]$ ? It is done in a straightforward way. Alice chooses some $n$ (say, the minimal value $n=l_{d}$ ) and tries to achieve (*) by adding large elements to $D_{n-d}$. More precisely, if $(*)$ is violated, Alice takes some number $k$ that is greater that all non-zero terms in $\mu$ (i.e., $\mu\left(k^{\prime}\right)=0$ for all $k^{\prime} \geqslant k$ ) and adds $k$ to $D_{n-d}$. Then Bob may increase $\mu$-values; as soon as $(*)$ is violated again, Alice repeats this procedure, and so on. At some point (after $2^{n-d}$ steps) a maximal cardinality of $D_{n-d}$ is reached. But at that time Bob has used at least $2^{n-d} 2^{-n-a_{n}}=2^{-d-a_{n}}$ of his reserve (each time a tail of size $2^{-n-a_{n}}$ is cut). Then Alice switches to next value of $n$, and forces Bob to lose or to use $2^{-d-a_{n}}$ again for this new value of $n$. Ultimately Bob will lose the game since the sum of $2^{-d-a_{n}}=2^{-d} 2^{-a_{n}}$ over $n$ in $\left[l_{d}, r_{d}\right]$ exceeds 1. (A technical correction: we required that the limit value of $\sum \mu(i)$ is strictly less that some threshold; it is not enough to know that all the approximations are strictly less than this threshold (only a non-strict inequality is guaranteed). To remedy this problem, we may use an additional factor of 2 - so we require the sum of $2^{-a_{n}}$ over $n \in\left[l_{d}, r_{d}\right]$ to be greater than $2^{d+1}$, not $2^{d}$.) Claim (i) is proven.

### 3.2 Proof of the Claim (ii)

We again consider a sequence of different pairs $\left(x_{n}, y_{n}\right)$ such that $x_{n} \leqslant y_{n}$, the sequence $x_{n}$ is lower semicomputable, the sequence $y_{n}$ is upper semicomputable and $\sum 2^{x_{n}-y_{n}}=+\infty$. We want to prove (following Gács) that there exists $n$ such that $B P^{\prime}\left(x_{n}\right)>B P\left(y_{n}\right)$

We can use the same reasoning as in (i) with minor modifications to show that we can assume without loss of generality that $\left(x_{n}, y_{n}\right)=\left(n-a_{n}, n\right)$ for some computable sequence of non-negative integers $a_{n}$ with $\sum_{n} 2^{-a_{n}}=+\infty$. This time we need to group terms with the same $y_{i}$, not $x_{i}$. We need to prove then that there exists $n$ such that $B P(n)<B P^{\prime}\left(n-a_{n}\right)$. In other words, we need to prove that there exist $n$ and $u$ such that $\mathbf{m}(i)<2^{-n}$ for all $i \geqslant u$, but $\sum_{i \geqslant u} \mathbf{m}(i)>2^{-n+a_{n}}$ (all terms in the $u$-tail are small but their sum is big).

To show that the sum of $\mathbf{m}$-tail is big, we need to construct a lower semicomputable semimeasure for which this sum is big, and then use the maximality of $\mathbf{m}$. Again a constant appears, so we need to prove a stronger statement: there exists a lower semicomputable semimeasure $\alpha$ such that for every $d$ there are $n$ and $u$ with the following property:

$$
\sum_{i \geqslant u} \alpha(i)>2^{-n+a_{n}+d} \quad \text { but } \quad \mathbf{m}(i)<2^{-n} \text { for all } i \geqslant u .
$$

Again we may use the game approach and imagine that Alice approximates from below some semimeasure $\alpha$ while Bob approximates from below some semimeasure $\beta$, and the claim above (with $\beta$ instead of $\mathbf{m}$ ) is the winning condition for Alice. We will construct a computable strategy for Alice in this game; applying it against the blind strategy of Bob (who approximates $\mathbf{m}(\cdot)$ from below), we get the required statement.

Let us note first that it is enough to construct (for every $d$ ) a winning strategy in the similar game where winning condition is required only for this $d$. Indeed, we may use $2 d$-strategy to win the $d$-game with $\sum_{i} \alpha(i) \leqslant 2^{-d}$ (using $2 d$-strategy with factor $2^{-d}$ ). Then we can use all the strategies (for $d$-games for all $d$ ) in parallel against Bob and sum up all the increases, since the winning condition is monotone and the strategies can only help each other. In this way Alice keeps the total sum less than $\sum_{d} 2^{-d} \leqslant 1$ and wins all games.

So how could Alice win the $d$-game? She should increase her weights gradually by using small weights far away where Bob has only zeros. As soon as her total weight exceeds $2^{-n+a_{n}+d}$ for some $n$, Bob has to react and assign weight at least $2^{-n}$ for some $i$. Then Alice continues to increase the weights (on the right of the place used by Bob), and again after $2^{-n+a_{n}+d}$ new Alice's weight Bob should react by assigning weight at least $2^{-n}$ at some other place. If Alice uses this strategy with small weights (see the discussion below) until her total weight reaches 1, and waits each time until Bob violates the winning condition for Alice, we have the following property of Bob's weights: ${ }^{3}$
${ }^{3}$ Technically speaking, Bob is obliged to react only if the Alice's tail is strictly greater than $2^{-n+a_{n}+d}$. But this leads only to a constant factor that is not important, so we ignore this problem.
for each $n$ there are at least $2^{n-a_{n}-d}$ Bob's weights $\beta(\cdot)$ that exceed $2^{-n}$.
Note that Alice's actions are the same for all $n$; it is Bob who should care about all $n$ and provide a large enough weight at the moments where Alice is in the winning position (for some $n$ ).

What is the total weight Bob uses in this process? The property above guarantees that Bob uses at least $2^{-a_{n}-d}$ to prevent Alice from winning for given $n$. The sum of these quantities for all $n$ is infinite according to our assumption (so at some point Bob will be unable to increase the weights). However, the same Bob's move can be useful on different levels (for different values of $n$ ), so we need the following technical lemma valid for every series $\sum_{i} \beta(i)$ with non-negative values:

$$
\sum_{j} 2^{-j} \cdot \#\left\{i: \beta(i) \geqslant 2^{-j}\right\} \leqslant 2 \sum_{i} \beta(i) .
$$

Indeed, each $\beta(i)$ from the right hand side appears in the left hand side as the sum of $2^{-j}$ for all $j$ such that $2^{-j} \leqslant \beta(i)$, and this sum does not exceed $2 \beta(i)$.

To finish the description of Alice's strategy, we need to say how small should be the weight increases used by Alice. We know that the sum $\sum_{n} 2^{-a_{n}-d}$ is infinite, so there is a finite part of this sum that is large (greater than 4, to be exact). Alice then may use weights $2^{-s}$ where $s$ is some integer greater that all $n+a_{n}+d$ for $n$ that appear in this finite part.

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# The Domino Problem for Self-similar Structures 

Sebastián Barbieri ${ }^{1(\boxtimes)}$ and Mathieu Sablik ${ }^{2}$<br>${ }^{1}$ LIP, ENS de Lyon - CNRS - INRIA - UCBL - Université de Lyon, Lyon, France<br>sebastian.barbieri@ens-lyon.fr<br>${ }^{2}$ Institut de Mathmatiques de Marseille, UMR 7373, Aix-Marseille Université, Marseille, France<br>mathieu.sablik@univ-amu.fr


#### Abstract

We define the domino problem for tilings over self-similar structures of $\mathbb{Z}^{d}$ given by forbidden patterns. In this setting we exhibit non-trivial families of subsets with decidable and undecidable domino problem.


## Introduction

In its original form, the domino problem was introduced by Wang [10] in 1961. It consists of deciding if copies of a finite set of Wang's tiles (square tiles of equal size, not subject to rotation and with colored edges) can tile the plane subject to the condition that two adjacent tiles possess the same color in the edge they share. Wang's student Berger showed undecidability for the domino problem on the plane in 1964 [3] by using a reduction to the halting problem. In 1971, Robinson [8] simplified Berger's proof.

Symbolic dynamics classically studies sets of colorings of $\mathbb{Z}^{d}$ from a finite set of colors which are closed in the product topology and invariant by translation, such sets are called subshifts. Given a finite set of patterns $\mathcal{F}$ (a pattern is a coloring of a finite part of $\left.\mathbb{Z}^{d}\right)$, we associate a subshift of finite type $X(\mathcal{F})$ which corresponds to the set of colorings which does not contain any occurrence of patterns in $\mathcal{F}$. The domino problem can therefore be expressed in this setting: given a finite set of forbidden patterns $\mathcal{F}$, is it possible to decide whether the subshift of finite type $X(\mathcal{F})$ is not empty?

It is well known that there exists an algorithm deciding if a subshift of finite type is empty in dimension one [5] and that there is no such algorithm in higher dimensions. The natural question that comes next is: What is the frontier between decidability and undecidability in the domino problem?

One way to explore this question is to consider subshifts defined over more general structures, such as finitely generated groups or monoids and ask where the domino problem is decidable. This approach has yielded various result in different structures: Some examples are the hyperbolic plane [6], confirming a conjecture of Robinson [9] and Baumslag-Solitar groups [1]. The conjecture in this direction is that the domino problem is decidable if and only if the group is virtually free. The conjecture is known to hold in the case of virtually nilpotent


Fig. 1. Some digitalizations of fractal structures and the status of their domino problem
groups [2]. The main idea of the proof of this result is to construct a grid by local rules in order to use the classical result in $\mathbb{Z}^{2}$.

In this paper we explore another way to delimit the frontier between decidability and undecidability of this problem. In geometry the structures which lie between the line and the plane can have Hausdorff dimension strictly between one and two. In this article we propose a way to define the domino problem in a digitalization of such fractal structures. In Sect. 1 we use self-similar substitutions to define a "fractal" structure where a natural version of the domino problem can be defined. We exhibit a large class of substitutions (including the one which represents the Sierpiński triangle) where the domino problem is decidable (Sect. 2), another class (including the Sierpiński carpet) where the problem is undecidable (Sect.4) and an intermediate class where the question is still open (see Fig. 1 for an example of each of these classes).

## 1 Position of the Problem

### 1.1 Coloring of $\mathbb{Z}^{d}$ and Local Rules

Given a finite alphabet $\mathcal{A}$, a coloring of $\mathbb{Z}^{d}$ is called a configuration. The set of configurations, denoted $\mathcal{A}^{\mathbb{Z}^{d}}$, is a compact set according to the usual product topology. A subshift is a closed set of configurations which is invariant by the shift action. Given a finite subset $S \subset \mathbb{Z}^{d}$, a pattern with support $S$ is an element $p$ of $\mathcal{A}^{S}$. A pattern $p \in \mathcal{A}^{S}$ appears in a configuration $x \in \mathcal{A}^{\mathbb{Z}^{d}}$ if there exists $z \in \mathbb{Z}^{d}$ such that $x_{z+S}=p$. In this case we write $p \sqsubset x$.

Equivalently, a subshift can be defined with a set of forbidden patterns $\mathcal{F}$ as the set of configurations where no patterns of $\mathcal{F}$ appear. We denote it by $X(\mathcal{F})$. If $\mathcal{F}$ is finite, $X(\mathcal{F})$ is called subshift of finite type which can be considered as the set of tilings defined by the local constraints given by $\mathcal{F}$.

### 1.2 Self-similar Structures

We want to extend the condition of coloring to self-similar structures of $\mathbb{Z}^{d}$. This means that only some cells can be decorated by elements of $\mathcal{A}$. To formalize that, a structure is coded as a subset of $\{0,1\}^{\mathbb{Z}^{d}}$ and self-similarity is obtained by a substitution.

Let $\mathcal{A}$ be a finite alphabet. A substitution is a function $s: \mathcal{A} \rightarrow \mathcal{A}^{R}$ where $R=\left[1, l_{1}\right] \times \cdots \times\left[1, l_{d}\right]$ is a $d$-dimensional rectangle. It is naturally extended to act over patterns which have rectangles as support by concatenation. We denote the successive iterations of $s$ over a symbol by $s, s^{2}, s^{3}$ and so on. The subshift generated by a substitution $s$ is the set

$$
X_{s}:=\left\{x \in \mathcal{A}^{\mathbb{Z}^{d}} \mid \forall p \sqsubset x, \exists n \in \mathbb{N}, a \in \mathcal{A}, p \sqsubset s^{n}(a)\right\} .
$$

To obtain self-similar structures, we restrict the notion of substitution to $\{0,1\}$ imposing that the image of 0 consists of a block of 0 s. These substitutions are called self-similar. Self-similar substitutions represent digitalizations of the iterations of the following procedure: start with the hypercube $[0,1]^{d}$, subdivide it in a $l_{1} \times \cdots \times l_{d}$ grid and remove the blocks in the positions $z$ of the grid where $s(1)_{z}=0$. Then repeat the same procedure with every sub-block.

Example 1. Consider $\mathcal{A}=\{\square, \square\}$. The self-similar substitution $s$ such that: $\square \longrightarrow \square$ and $\square \longrightarrow \square$ is called the Sierpiński triangle substitution and is extended by concatenation as shown in Fig. 2.


Fig. 2. First four iterations of the Sierpiński triangle substitution.

### 1.3 Coloring of a Self-similar Structure and Local Rules

Let $\mathcal{A}$ be a finite alphabet where $0 \in \mathcal{A}$ and $s$ be a self-similar substitution. Consider $X_{s} \subset\{0,1\}^{\mathbb{Z}^{d}}$ the associated self-similar structure. A configuration $x \in \mathcal{A}^{\mathbb{Z}^{d}}$ is compatible with $s$ if $\pi(x) \in X_{s}$ where $\pi$ is a map which sends all elements of $\mathcal{A} \backslash\{0\}$ onto 1 and 0 onto 0 . Given a finite set of patterns $\mathcal{F}$ we define the set of configurations on $X_{s}$ defined by local rules $\mathcal{F}$ as

$$
X_{s}(\mathcal{F})=\left\{x \in \mathcal{A}^{\mathbb{Z}^{d}}: \pi(x) \in X_{s} \text { and no pattern of } \mathcal{F} \text { appears in } x\right\}
$$

### 1.4 The Domino Problem on Self-similar Structures

The domino problem for $\mathbb{Z}^{d}$ is defined as the language

$$
\mathrm{DP}\left(\mathbb{Z}^{d}\right)=\{\mathcal{F} \text { finite set of patterns : } X(\mathcal{F}) \neq \emptyset\}
$$

It is the language of all finite sets of patterns over a finite alphabet such that it is possible to construct a configuration without patterns of $\mathcal{F}$.

Classical results which can be found in [5] show that the domino problem for $\mathbb{Z}$ is decidable. In the other hand, we know that for $d>1$ the domino problem for $G=\mathbb{Z}^{d}$ is undecidable (see $[3,8]$ ). This gap of decidability when the dimension increases motivates us to define the domino problem for structures which lay between those groups. Thus given a self-similar substitution $s$ we introduce the $s$-based domino problem as the language

$$
\operatorname{DP}(s):=\left\{\mathcal{F} \text { finite set of patterns }: X_{s}(\mathcal{F}) \neq\left\{0^{\mathbb{Z}^{d}}\right\}\right\}
$$

That is, $\mathrm{DP}(s)$ is the set of all finite sets of forbidden patterns such that there is at least a configuration containing a non-zero symbol. We assume implicitly that $\mathcal{F}$ does not contain any pattern consisting only of 0 s . By a compactness argument, $\operatorname{DP}(s)$ can be equivalently defined as the set of $\mathcal{F}$ such that for arbitrarily big $n \in \mathbb{N}$ the non-zero symbols in $s^{n}(1)$ can be colored avoiding all patterns in $\mathcal{F}$.

## 2 Self-similar Structures with Decidable Domino Problem

In this section we present a family of self-similar substitutions such that their domino problem is decidable. In order to present this result in the most general setting, we introduce the channel number of a self-similar substitution.

Let $\mathbb{H}=\{-1,0,1\}^{d}$ and consider the set $\Lambda \subset\{0,1\}^{\{1,2,3\}^{d}}$ consisting of all $d$-dimensional hypercube patterns of side 3 which appear in $X_{s}$ and that have a 1 in the center. Let $\Lambda_{n}=s^{n}(\Lambda)$ be the set of the images of each $q \in \Lambda$ under $s^{n}$ by concatenation and $S_{n}$ be the support corresponding to the image of position $(2, \ldots, 2)$ of $q$ under $s^{n}$. We define the $n$-channel number $\chi(s, n)$ of $s$ as follows:

$$
\chi(s, n)=\max _{p \in \Lambda_{n}}\left|\left\{z \in S_{n} \mid \exists h \in(z+\mathbb{H}) \cap\left(\operatorname{supp}(p) \backslash S_{n}\right), p_{z}=p_{h}=1\right\}\right|
$$

In other words, it is the maximum number of positions in the support of the pattern $s^{n}(1)$ such that if we surround it either by blocks of 0 or copies of $s^{n}(1)$ appearing in $X_{s}$ there might be two symbols 1, one appearing in $s^{n}(1)$ and another outside, at distance smaller than 1 . We say that $s$ is channel bounded if there exists $K \in \mathbb{N}$ such that for all $n, \chi(s, n) \leq K$. The Sierpiński triangle substitution from Fig. 2 is an example of a channel bounded substitution as colorings of $s^{n}(1)$ can be constructed by pasting three colorings of $s^{n-1}(1)$ and forbidden patterns can only appear locally around the corners.

Theorem 1. For every channel bounded self-similar substitution s the domino problem $\operatorname{DP}(s)$ is decidable.

Proof. Let $\mathcal{F}$ be a set of forbidden patterns over an alphabet $\mathcal{A}$. It suffices to show that if $s$ is channel bounded, it is possible to calculate $N \in \mathbb{N}$ with the property that the existence of any coloring of $s^{N}(1)$ with symbols from $\mathcal{A}$ without any subpattern from $\mathcal{F}$ implies $X_{s}(\mathcal{F}) \neq\left\{0^{\mathbb{Z}^{d}}\right\}$. Indeed, an algorithm could calculate $N$ and try every coloring of $s^{N}(1)$. If there exists one where no pattern in $\mathcal{F}$ appears it returns that $X_{s}(\mathcal{F}) \neq\left\{0^{\mathbb{Z}^{d}}\right\}$, otherwise it returns $X_{s}(\mathcal{F})=\left\{0^{\mathbb{Z}^{d}}\right\}$.

For simplicity, suppose that $\forall p \in \mathcal{F}, \operatorname{supp}(p) \subseteq \mathbb{H}$ and let $K$ be a bound for $\chi(s, n)$ (If the support is $\{-m, \ldots, m\}^{d}$ we can recalculate a new $K$ ). We claim that $N:=2^{|\mathcal{A}|^{\left(3^{d}-1\right) K}}$ suffices. For each $q \in \Lambda$ consider the a coding $J_{q}=\left\{j_{1}, \ldots, j_{k_{q}}\right\}$ with $k_{q} \leq K$ of the positions from the definition of $\chi(s, n)$. That is, $J_{q}$ codes for all $n \in \mathbb{N}$ the set of positions which matter when considering only $q$. Any recursive ordering similar to the one given by a Quadtree works. Consider a coloring of $s^{n}(1)$ without subpatterns in $\mathcal{F}$ and store the symbols of this coloring appearing in $J_{q}$ as a tuple $\left(a_{j_{1}}, a_{j_{2}}, \ldots a_{j_{k_{q}}}\right) \in \mathcal{A}^{J_{q}}$. Therefore all the information concerning the dependency of this coloring with its possible surroundings can be stored on $|\Lambda|$ tuples. Now, given the set of all colorings of $s^{n}(1)$ which do not contain any forbidden pattern we can extract the $|\Lambda|$ tuples from each one of them. All this information for the level $n$ is represented as a subset of $\prod_{q \in \Lambda} \mathcal{A}^{J_{q}}$. By definition this is the only information needed in order to make sure of the existence of a coloring of $s^{n+1}(1)$ with no subpatterns in $\mathcal{F}$. Moreover, the tuples representing those patterns can be obtained from the ones of $s^{n}(1)$ because the positions from the definition of $\chi(s, n+1)$ necessarily appear in the patterns of $s^{n}(1)$. This means it is possible to extract pasting rules which can be codified in a function $\mu_{s}: 2^{\prod_{q \in \Lambda} \mathcal{A}^{J q}} \rightarrow 2^{\prod_{q \in \Lambda} \mathcal{A}^{J q}}$.

This function codes how to construct the tuples of level $n+1$ from the tuples of level $n$. Obviously, $\mu_{s}(\emptyset)=\emptyset$, therefore there are two possibilities: either this function arrives eventually at $\emptyset$ and there are no colorings of $s^{m}(1)$ for some $m \in \mathbb{N}$ or $\mu_{s}$ cycles and thus it's possible to construct colorings of $s^{m}(1)$ for arbitrarily big $m$. By pigeonhole principle this behavior must occur before $\left|2^{\prod_{q \in \Lambda} \mathcal{A}^{J_{q}}}\right| \leq 2^{|\mathcal{A}|^{\left(3^{d}-1\right) K}}$ iterations of $\mu_{s}$.

## 3 The Mozes Property for Self-similar Structures

Most of the proofs of the undecidability of the domino problem on $\mathbb{Z}^{2}$ are based on the construction of a self-similar structure. A Theorem proven by Mozes [7] and later generalized by Goodman-Strauss [4] shows that every $\mathbb{Z}^{d}$-substitutive subshift is a sofic subshift for $d \geq 2$. This theorem fails for the case $d=1$. The importance of this result is the fact that multidimensional substitutions can be realized by local rules. In order to present a family of self-similar substitutions with undecidable domino problem we will make use of an analogue of the theorem shown by Mozes.

Definition 1. A self-similar substitution s satisfies the Mozes property if for every substitution $s^{\prime}$ defined over the same rectangle and over an alphabet $\mathcal{A}$ containing 0 and such that $\forall a \in \mathcal{A} \backslash\{0\}, \pi\left(s^{\prime}(a)\right)=s(1)$ and $s^{\prime}(0)=s(0)$ there exists an alphabet $\mathcal{B}$ containing the symbol 0 , a finite set of forbidden patterns $\mathcal{F} \subseteq \mathcal{B}_{\mathbb{Z}^{d}}^{*}$ and a local function $\Phi: \mathcal{B} \rightarrow \mathcal{A}$ such that $\Phi(0)=0$ and the function $\phi: \mathcal{B}^{\mathbb{Z}^{d}} \rightarrow \mathcal{A}^{\mathbb{Z}^{d}}$ given by $\phi(x)_{z}=\Phi\left(x_{z}\right)$ is surjective from $X_{s}(\mathcal{F})$ to $X_{s^{\prime}}$.

In other words, it's the analogue of saying that $X_{s^{\prime}}$ is a sofic subshift, except that the SFT extension has to be based on $X_{s}$. Currently, we have been able to produce a class of substitutions that satisfy the Mozes property but we have not found a characterization of those who do. An example of a substitution without the Mozes property is the one given by


An interesting example of a substitution satisfying the Mozes property is the Sierpiński carpet shown in Fig. 3. This substitution it not channel bounded as at least one of the four borders of a coloring of $s^{n-1}(1)$ matter when constructing $s^{n}(1)$ and thus $\chi(s, n)$ grows exponentially. In fact this substitution belongs to a bigger class which also satisfies the Mozes property. In the next section we introduce this class and use this previous fact to prove that all the substitutions belonging to it have undecidable domino problem.


Fig. 3. The first iterations of the Sierpiński carpet substitution.

## 4 Self-similar Structures Where the Domino Problem Is Undecidable

In this section we present a family of two-dimensional self-similar substitutions with undecidable $s$-based domino problem. The definition of this class follows.

Definition 2. A self-similar substitution $s$ defined on $\left[1, l_{1}\right] \times\left[1, l_{2}\right]$ contains a grid if there are integers $1 \leq i_{1}<i_{2}<=l_{1}$ and $1 \leq j_{1}<j_{2}<=l_{2}$ such that $j \in\left\{j_{1}, j_{2}\right\}$ or $i \in\left\{i_{1}, i_{2}\right\}$ implies that $s(1)_{(i, j)}=1$.

One example of a self-similar substitution that contains a grid is the Sierpiński carpet. One interesting property of these substitutions is that they satisfy the Mozes property. This follows from a technical construction which uses a layer that looks like a generalized Robinson tiling [8] and stores the information of the simulated substitution and its past hierarchically.

Theorem 2. All self-similar substitutions which contain a grid satisfy the Mozes property.

In what remains of this section we show the following theorem:
Theorem 3. Let s be a self-similar substitution which contains a grid. Then the domino problem $\operatorname{DP}(s)$ is undecidable.

Proof. We claim that an oracle for $\operatorname{DP}(s)$ can be used to decide $\operatorname{DP}\left(\mathbb{Z}^{2}\right)$. This is enough to conclude, as $\operatorname{DP}\left(\mathbb{Z}^{2}\right)$ is undecidable.

Let $s$ be defined on $\left[1, l_{1}\right] \times\left[1, l_{2}\right]$, some values satisfying the grid condition $\left(i_{1}, i_{2}\right)$ and $\left(j_{1}, j_{2}\right)$ and consider a substitution $s^{\prime}$ over the alphabet $\mathcal{A}\left(s^{\prime}\right)=\{\boldsymbol{\square}, \boldsymbol{\square}, \boldsymbol{\Pi}, 0\}$ given by the following rules: Let $C=$ $\left\{\left(i_{1}, j_{1}\right),\left(i_{1}, j_{2}\right),\left(i_{2}, j_{1}\right),\left(i_{2}, j_{2}\right)\right\}, H=\left\{(i, j) \mid j \in\left\{j_{1}, j_{2}\right\}\right\} \backslash C$ and $V=\{(i, j) \mid i \in$ $\left.\left\{i_{1}, i_{2}\right\}\right\} \backslash C$.
$s^{\prime}(\mathbb{\square})_{z}=\left\{\begin{array}{ll}0, & \text { if } s(1)_{z}=0 \\ \boxminus, & \text { if } z \in H \\ \boldsymbol{\Pi}, & \text { if } z \in V \\ \square, & \text { else }\end{array} \quad s^{\prime}(\mathbb{(})_{z}= \begin{cases}0, & \text { if } s(1)_{z}=0 \\ \boldsymbol{D}, & \text { if } z \in V \cup C \quad s^{\prime}(\boxminus)_{z}=\left\{\begin{array}{ll}0, & \text { if } s(1)_{z}=0 \\ \text { ■, } & \text { else }\end{array} \quad \text { if } z \in H \cup C\right. \\ \square, & \text { else }\end{cases}\right.$
For example, in the case where $s$ is the Sierpiński carpet we get:


For any $y \in X_{s^{\prime}} \backslash\left\{0^{\mathbb{Z}^{2}}\right\}$ and $n \in \mathbb{N}$ we have $s^{\prime n}(\square) \sqsubset y$. Indeed, $\square$ appears in the image of every symbol $a \in \mathcal{A}\left(s^{\prime}\right) \backslash\{0\}$. This implies that for every positive integer $n$, a $\square$ must appear at a bounded distance of every non-zero symbol in $s^{\prime n}(a)$. This argument extends inductively because if $s^{\prime n-1}(\square)$ appears at a bounded distance in every $s^{\prime k}(a)$ with $k>n$, it suffices to apply $s^{\prime}$ to obtain that $s^{\prime n}$ ( $\left.\mathbf{\square}\right)$ appears at bounded distance in $s^{\prime k+1}(a)$.

As $s$ satisfies the Mozes property there exists an alphabet $\mathcal{B}\left(s^{\prime}\right)$, a finite set $\mathcal{F}\left(s^{\prime}\right) \subset \mathcal{B}\left(s^{\prime}\right)_{\mathbb{Z}^{2}}^{*}$ and $\Phi: \mathcal{B}\left(s^{\prime}\right) \rightarrow \mathcal{A}\left(s^{\prime}\right)$ such that $\Phi(0)=0$ and the function $\phi: \mathcal{B}\left(s^{\prime}\right)^{\mathbb{Z}^{d}} \rightarrow \mathcal{A}\left(s^{\prime}\right)^{\mathbb{Z}^{d}}$ given by $\phi(x)_{z}=\Phi\left(x_{z}\right)$ is surjective from $X_{s}(\mathcal{F})$ to $X_{s^{\prime}}$.

Consider a finite set of forbidden patterns $\mathcal{F}$ over an alphabet $\mathcal{A}$ defining a $\mathbb{Z}^{2}$ subshift $X(\mathcal{F})$. Without loss of generality $\mathcal{F}$ contains only patterns with supports $\{(0,0),(1,0)\}$ and $\{(0,0),(0,1)\}$ (one can choose a conjugated version of $X(\mathcal{F})$ satisfying this property by using a higher block code. See [5]).

Finally, consider the alphabet $\mathcal{S}:=\mathcal{B}\left(s^{\prime}\right) \times(\mathcal{A} \cup\{0\})$ along with the set of forbidden patterns $\mathcal{G}$ given by the union of the following sets:

- Zeros correspond: $\{(0, a) \mid a \in \mathcal{A}\} \cup\left\{(b, 0) \mid b \in \mathcal{B}\left(s^{\prime}\right) \backslash\{0\}\right\}$.
- First layer forbidden patterns: $\left\{p \times q \mid p \in \mathcal{F}\left(s^{\prime}\right), q \in \mathcal{A}^{\operatorname{supp}(p)}\right\}$. These forbidden patterns make sure that configurations belonging to the first layer of $X_{s}(\mathcal{G})$ belong to $X_{s}\left(\mathcal{F}\left(s^{\prime}\right)\right)$.
- Horizontal forbidden patterns: let $p \in \mathcal{S}\{(0,0),(1,0)\}$ be denoted by $(a, b, c, d)$ if $p(0,0)=(a, c)$ and $p(1,0)=(b, d)$ and $q \in \mathcal{A}^{\{(0,0),(1,0)\}}$ be denoted by $(c, d)$ if $q(0,0)=c$ and $q(1,0)=d$. The set of horizontal forbidden patterns is $\{(a, b, c, d) \mid(a=\boldsymbol{日}, b \in\{\boldsymbol{\theta}, \boldsymbol{\square}\}$ and $c \neq d)$ or $(a=\square, b=$ 日 and $(c, d) \in \mathcal{F})\}$.
- Vertical forbidden patterns: let $p \in \mathcal{S}\{(0,0),(0,1)\}$ be denoted by $(a, b, c, d)$ if $p(0,0)=(a, c)$ and $p(0,1)=(b, d)$ and $q \in \mathcal{A}^{\{(0,0),(0,1)\}}$ be denoted by $(c, d)$ if $q(0,0)=c$ and $q(0,1)=d$. The set of vertical forbidden patterns is given by $\{(a, b, c, d) \mid(a=\boldsymbol{\Pi}, b \in\{\boldsymbol{\square}, \boldsymbol{\square}\}$ and $c \neq d)$ or $(a=\boldsymbol{\square}, b=\boldsymbol{\square}$ and $(c, d) \in \mathcal{F})\}$.

These rules codify the following idea: $\square$ s carry arbitrary symbols from $\mathcal{A}$ in the second layer and the arrows send this information left and up respecting the rules from $\mathcal{F}$, see Fig. 4. By iterating the substitution $s$ it is easy to see that $s^{n}(1)$ actually contains $2^{n}$ vertical and horizontal lines. This means that the intersections of these lines contain symbols of $\mathcal{A}$ which represent a $2^{n} \times 2^{n}$ pattern which contains no forbidden pattern from $\mathcal{F}$. Therefore if $X_{s}(\mathcal{G}) \neq\left\{0^{\mathbb{Z}^{2}}\right\}$ then $X(\mathcal{F}) \neq \emptyset$ by compactness. Conversely if $X(\mathcal{F}) \neq \emptyset$ it is possible to always build the second layer of a point having $s^{\prime n}(1)$ in the first layer.

Suppose there is an algorithm for deciding $\operatorname{DP}(s)$. Then for any $\mathcal{F}$ defining a $\mathbb{Z}^{2}$ subshift the alphabet $\mathcal{S}$ and the rules $\mathcal{G}$ can be built in order to decide if $X_{s}(\mathcal{G}) \neq\left\{0^{\mathbb{Z}^{2}}\right\}$. This is equivalent to deciding if $X(\mathcal{F}) \neq \emptyset$, therefore $\operatorname{DP}\left(\mathbb{Z}^{2}\right)$ can be decided. This yields the desired contradiction.


Fig. 4. On the left a pattern from $X(\mathcal{F})$. In the right its coding in $X_{s}(\mathcal{G})$. The blue squares are arbitrary symbols from $\mathcal{A}$. (Color figure online)

## 5 Generalizations and Perspectives

Here we present some ideas to generalize previous results in order to advance towards a characterization of the self-similar structures where the domino problem is decidable. In the previous sections the information which allows to simulate grids is transferred through straight lines. We can imagine less rigid possibilities.

### 5.1 Connectivity

We propose a way to define the directions in which the information can be transferred in a substitution in $\mathbb{Z}^{2}$. Given a self-similar substitution defined over $\left[1, l_{1}\right] \times\left[1, l_{2}\right]$ we denote by $\mathbb{X}$ the set of coordinates $z$ such that $s(1)_{z}=1$. Let $\mathbb{S}=\{(0,-1),(0,1),(-1,0),(0,1)\}$ and $\mathbb{W}$ contains $\{(-1,-1),(1,1)\}$ if $\left\{(1,1),\left(l_{1}, l_{2}\right)\right\} \in \mathbb{X}$ and $\{(-1,1),(1,-1)\}$ if $\left\{\left(1, l_{2}\right),\left(l_{1}, 1\right)\right\} \in \mathbb{X}$. We say $s$ admits a rigid (respectively flexible) vertical line at $1 \leq v \leq l_{1}$ if there is a non-repeating sequence of vertices $(v, 1)=x_{1}, \ldots x_{n}=\left(v, l_{2}\right)$ such that the differences $x_{j}-x_{j-1}$ belong to $\mathbb{S}$ (respectively $\mathbb{W} \cup \mathbb{S}$ ). We define rigid and flexible horizontal lines for $1 \leq h \leq l_{2}$ analogously. We also say that two lines are weakly disjoint if they share no consecutive pair of vertices in their path.

According to these notions, we distinguish the following four subclasses:

- $s$ has bounded connectivity if $s$ has at most one flexible horizontal and vertical line;
- $s$ has $a$ isthmus if $s(1)$ has at least two weakly disjoint flexible lines in one direction and at most one weakly disjoint flexible line in the other direction;
- $s$ has a weak grid if $s(1)$ has at least two flexible horizontal lines and two flexible vertical lines which are pairwise weakly disjoint.
- $s$ has a strong grid if $s(1)$ has at least two rigid horizontal lines and two rigid vertical lines which are pairwise weakly disjoint.

If $s$ has bounded connectivity the proof of Theorem 1 can be adapted to show decidability. If $s$ has a strong grid it is possible to adapt the proof of Theorem 3 to show the undecidability of the domino problem associated to such substitution, moreover, a generalization of that proof works even in the case of weak grids. Nevertheless we still have no results supporting either direction in the isthmus case. We believe that the Mozes property does not hold in the isthmus case, which would be evidence towards decidability. Figure 5 presents the domino problem of different substitutions according to this classification.

| B. Connectivity | Isthmus | Weak grid | Strong grid |
| :---: | :---: | :---: | :---: |
| DP decidable | Unknown | DP undecidable | DP undecidable |
|  |  |  |  |

Fig. 5. Some examples of substitution according to this classification

### 5.2 Concluding Remarks

We introduced the domino problem on self-similar structures in order to understand the frontier between decidability and undecidability when we go from the line (dimension 1) to the plane (dimension 2). Our results show that there is no decidability threshold on the Hausdorff dimension. Indeed, there are self-similar structures with decidable domino problem and Hausdorff dimension arbitrary near to 2 (obtained by $s_{n}$ ) and self-similar structures with undecidable domino problem and Hausdorff dimension arbitrary near to 1 (obtained by $s_{n}^{\prime}$ ).


Thus, the frontier between decidability and undecidability seems more likely to be based on the presence of a grid where it is possible to implement a computation. To confirm this hypothesis, it remains to study self-similar structures with an isthmus. In the case of an isthmus the substitution presents an unique bridge which links different zones. This impedes a Mozes-like [7] or Goodman-Strauss-like [4] construction. The main problem is that in order to simulate a substitution there is the need to transfer arbitrarily big amounts of information by that isthmus. We believe the study of this class of substitutions will certainly provide new tools to the study of how information can be transfered.

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# Axiomatizing Analog Algorithms 

Olivier Bournez ${ }^{1(\boxtimes)}$, Nachum Dershowitz ${ }^{2}$, and Pierre Néron ${ }^{3}$<br>${ }^{1}$ Laboratoire d'Informatique de l'X (LIX), École Polytechnique, Palaiseau, France<br>bournez@lix.polytechnique.fr<br>${ }^{2}$ School of Computer Science, Tel Aviv University, Ramat Aviv, Israel<br>nachum@cs.tau.ac.il<br>${ }^{3}$ French Network and Information Security Agency (ANSSI), Paris, France<br>pierre.neron@ssi.gouv.fr


#### Abstract

We propose a formalization of generic algorithms that includes analog algorithms. This is achieved by reformulating and extending the framework of abstract state machines to include continuous-time models of computation. We prove that every hybrid algorithm satisfying some reasonable postulates may be expressed precisely by a program in a simple and expressive language.


## 1 Introduction

In [14], Gurevich showed that any algorithm that satisfies three intuitive "Sequential Postulates" can be step-by-step emulated by an abstract state machine (ASM). These postulates formalize the following intuitions: (I) one is dealing with discrete deterministic state-transition systems; (II) the information in states suffices to determine future transitions and may be captured by logical structures that respect isomorphisms; and (III) transitions are governed by the values of a finite and input-independent set of ground terms. All notions of algorithms for "classical" discrete-time models of computation in computer science are covered by this formalization. This includes Turing machines, random-access memory (RAM) machines, and their sundry extensions. The geometric constructions in [18], for example, are loop-free examples of discrete-step continuous-space (real-number) algorithms. The ASM formalization also covers general discrete-time models evolving over continuous space like the Blum-ShubSmale machine model [1].

However, capturing continuous-time models of computation is still a challenge, that is to say, capturing models of computation that operate in continuous (real) time and with real values. Examples of continuous-time models of computations include models of analog machines like the General Purpose Analog

[^32]Computer (GPAC) of Claude Shannon [20], proposed as a mathematical model of the Differential Analyzers, built for the first time in 1931 [7], and used to solve various problems ranging from ballistics to aircraft design - before the era of the digital computer [16]. Others include Pascal's 1642 Pascaline, Hermann's 1814 Planimeter, as well as Bill Phillips' 1949 water-run Financephalograph. Continuous-time computational models also include neural networks and systems built using electronic analog devices. Such systems begin in some initial state and evolve over time; results are read off from the evolving state and/or from a terminal state. More generally, determining which systems can actually be considered to be computational models is an intriguing question and relates to philosophical discussions about what constitutes a programmable machine. Continuous-time computation theory is far less understood than its discretetime counterpart [4]. Another line of development of continuous-time models was motivated by hybrid systems, particularly by questions related to the hardness of their verification and control. In hybrid systems, the dynamics change in response to changing conditions, so there are discrete transitions as well as continuous ones. Here, models are not seen as necessarily modeling analog machines, but, rather, as abstractions of systems about which one would like to establish properties or derive verification algorithms [4]. Some work on ASM models dealing with continuous-time systems has been accomplished for specific cases $[8,9]$. Rust [19] specifies forms of continuous-time evolution based on ASMs using infinitesimals. However, we find that a comprehensive framework capturing general analog systems is still wanting.

Our goal is to capture all such analog and hybrid models within one uniform notion of computation and of algorithm. To this end, we formalize a generic notion of continuous-time algorithm. The proposed framework is an extension of [14], as discrete-time algorithms are a simple special case of analog algorithms. (The initial attempt [5] was not fully satisfactory, as no completeness theorem nor general-form result was obtained. Here, we indeed achieve both.) We provide postulates defining continuous-time algorithms, in the spirit of those of [14], and we prove some completeness results. We define a simple notion of an analog ASM program and prove that all models satisfying the postulates have corresponding analog programs (Lemma 16 and Theorem 20). Furthermore, we provide conditions guaranteeing that said program is unique up to equivalence (Theorem 21 and Corollary 22). All of this seamlessly extends the results of [14] to analog and hybrid systems. The proposed framework covers all classes of continuous-time systems that can be modeled by ordinary differential equations or have hybrid dynamics, including the models in [4] and the examples in [5]. It is a first step towards a general understanding of computability theory for continuous-time models, taken in the hope that it will also lead to a formalization of a "ChurchTuring thesis" for analog systems in the spirit of what has been achieved for discrete-time models $[2,3,10]$. Systems with continuous input signals and other means of specifying continuous behavior are left for future work.

Some of our ideas were inspired by the way the semantics of hybrid systems are given in the approach of Platzer [17]. Among attempts at studying the
semantics of analog systems within a general framework is [22]. Recent results on comparing analog models include [11]. Soundness and (relative) completeness results for a programming language with infinitesimals have also been obtained in [21]. Applications to verification have been explored [15].

## 2 General Algorithms

We want to generalize the notion of algorithms introduced by Gurevich in [14] in order to capture not only the sequential case but also continuous behavior. (For lack of place, we assume some familiarity with [14].) However, when evolving continuously, an algorithm can no longer be viewed as a discrete sequence of states, and we need a notion of evolution that can capture both kinds of behavior. This is based on a notion of a timeline that corresponds to algorithm execution.

Definition 1 (Time). Time $\mathbb{T}$ corresponds to a totally ordered monoid: there is an associative binary operation + , with some neutral element 0 , and a total relation $\leq$ preserved by $+: t \leq t^{\prime}$ implies $t+t^{\prime \prime} \leq t^{\prime}+t^{\prime \prime}$ for all $t^{\prime \prime} \in \mathbb{T}$.

An element of $\mathbb{T}$ will be called a moment. Examples of time $\mathbb{T}$ are $\mathbb{R} \geq 0$ and $\mathbb{N}$. As expected, $t<t^{\prime}$ will mean $t \leq t^{\prime}$ but not $t=t^{\prime}$.

Definition 2 (Timeline). A timeline is a subset of $\mathbb{T}$ containing 0. We let $\mathbb{I}$ denote the set of all timelines.

For a moment $i \in I$ of timeline $I$, we write $\operatorname{Jump}(i)$ if there exists $t \in I$ with $i<t$, and there is no $t^{\prime} \in I$ with $i<t^{\prime}<t$. We write $F l o w(i)$ otherwise: that means that for all $t, i<t$, there is some in-between $t^{\prime} \in I$ with $i<t^{\prime}<t$. A moment $i$ with $\operatorname{Jump}(i)$ is meant to indicate a discrete transition. In this case, we write $i^{+}$for the smallest $t$ greater than $i$. A timeline $I$ is non-Zeno if for any moment $i \in I$, there is a finite number of moments $j \leq i$ with $\operatorname{Jump}(j)$. $\mathbb{I}$ is non-Zeno if all its timelines are.

For timelines $\mathbb{I}=\mathbb{R}^{\geq 0}$, for instance, we have $\operatorname{Flow}(i)$ for all $i \in \mathbb{I}$. For $\mathbb{I}=\mathbb{N}$, we have $\operatorname{Jump}(i)$ for all $i \in \mathbb{I}$, and $i^{+}=i+1$. We intend (for hybrid systems, in particular) to also consider timelines mixing both properties, that is, with Flow $(i)$ for some $i$ and $\operatorname{Jump}(i)$ for other $i$. Formally building such timelines is easy (for example $\bigcup_{n \in \mathbb{N}}[n, n+0.5]$ ). All these examples are non-Zeno.

Definition 3 (Truncation). Given a timeline $I \in \mathbb{I}$ and a moment $i$ of $I$, the truncated timeline $I[i]$ is the timeline defined by $I[i]=\{t \mid i+t \in I\}$.

With timelines in hand, we can define hybrid dynamical systems.
Definition 4 (Dynamical System). A dynamical system $\left\langle\mathcal{S}, \mathcal{S}_{0}, \iota, \varphi\right\rangle$ consists of the following: (a) a nonempty set (or class) $\mathcal{S}$ of states; (b) a nonempty subset (or subclass) $\mathcal{S}_{0} \subseteq \mathcal{S}$, called initial states; (c) a timeline map $\iota: \mathcal{S} \rightarrow \mathbb{I}$, with $\mathbb{I}$ non-Zeno; (d) a trajectory $\operatorname{map} \varphi:(X: \mathcal{S}) \times \iota(X) \rightarrow \mathcal{S}$. We require that, for any state $X$ and moments $i, i+i^{\prime} \in \iota(X)$, one has

$$
\varphi(X, 0)=X, \quad \iota(\varphi(X, i))=\iota(X)[i], \quad \varphi\left(X, i+i^{\prime}\right)=\varphi\left(\varphi(X, i), i^{\prime}\right)
$$

Together, the timeline and trajectory maps associate to each state its future evolution. For a state $X, \iota(X)$ defines the timeline corresponding to the system behavior starting from $X$, and $\varphi(X)$ defines its concrete evolution by associating to each moment in $\iota(X)$ its corresponding state. The third condition ensures that evolution during $i+i^{\prime}$ is similar to first evolving during $i$ and then during $i^{\prime}$; the preceding condition ensures a similar property for timelines (and ensures consistency of the last condition).

Postulate I. An algorithm is a dynamical system.
A vocabulary $\mathcal{V}$ is a finite collection of fixed-arity (possibly nullary) function symbols, some functions of which may be tagged relational. A term whose outermost function symbol is relational is termed Boolean. We assume that $\mathcal{V}$ contains the scalar (nullary) function true. A (first-order) structure $X$ of vocabulary $\mathcal{V}$ is a nonempty set $S$, the base set (domain) of $X$, together with interpretations of the function symbols in $\mathcal{V}$ over $S$ : A $j$-ary function symbol $f$ is interpreted as a function, denoted $\llbracket f \rrbracket_{X}$, from $S^{j}$ to $S$. Elements of $S$ are also called elements of $X$, or values. Similarly, the interpretation of a term $f\left(t_{1}, \ldots, t_{n}\right)$ in $X$ is recursively defined by $\llbracket f\left(t_{1}, \ldots, t_{n}\right) \rrbracket_{X}=\llbracket f \rrbracket_{X}\left(\llbracket t_{1} \rrbracket_{X}, \ldots, \llbracket t_{n} \rrbracket_{X}\right)$.

Let $X$ and $Y$ be structures of the same vocabulary $\mathcal{V}$. An isomorphism from $X$ onto $Y$ is a one-to-one function $\zeta$ from the base set of $X$ onto the base set of $Y$ such that $f\left(\zeta x_{1}, \ldots, \zeta x_{j}\right)=\zeta x_{0}$ in $Y$ whenever $f\left(x_{1}, \ldots, x_{j}\right)=x_{0}$ in $X$.

Definition 5 (Abstract Transition System). An abstract transition system is a dynamical system whose states $\mathcal{S}$ are (first-order) structures over some finite vocabulary $\mathcal{V}$, such that the following hold:
(a) States are closed under isomorphism, so if $X \in \mathcal{S}$ is a state of the system, then any structure $Y$ isomorphic to $X$ is also a state in $\mathcal{S}$, and $Y$ is an initial state if $X$ is.
(b) Transformations preserve the base set: that is, for every state $X \in \mathcal{S}$, for any $i \in \iota(X), \varphi(X, i)$ has the same base set as $X$.
(c) Transformations respect isomorphisms: if $X \cong_{\zeta} Y$ is an isomorphism of states $X, Y \in \mathcal{S}$, then $\iota(X)=\iota(Y)$ and for all $i \in \iota(X), X_{i} \cong_{\zeta} Y_{i}$, where $X_{i}=\varphi(X, i)$, and $Y_{i}=\varphi(Y, i)$.

Postulate II. An algorithm is an abstract transition system.
When $\iota(X)$ is $\mathbb{N}$ (or order-isomorphic to $\mathbb{N}$ ) for all $X$, this corresponds precisely to the concepts introduced by [14], considering that $\varphi(X, n)=\tau^{[n]}(X)$.

It is convenient to think of a structure $X$ as a memory of some kind: If $f$ is a $j$-ary function symbol in vocabulary $\mathcal{V}$, and $\bar{a}$ is a $j$-tuple of elements of the base set of $X$, then the pair $(f, \bar{a})$ is called a location. We denote by $\llbracket f(\bar{a}) \rrbracket_{X}$ its interpretation in $X$, i.e. $\llbracket f \rrbracket_{X}(\bar{a})$. If $(f, \bar{a})$ is a location of $X$ and $b$ is an element of $X$ then $(f, \bar{a}, b)$ is an update of $X$. When $Y$ and $X$ are structures over the same domain and vocabulary, $Y \backslash X$ denotes the set of updates $\Delta^{+}=\left\{\left(f, \bar{a}, \llbracket f(\bar{a}) \rrbracket_{Y}\right) \mid\right.$ $\left.\llbracket f(\bar{a}) \rrbracket_{Y} \neq \llbracket f(\bar{a}) \rrbracket_{X}\right\}$.

We want instantaneous evolution to be describable by updates:

Definition 6. An infinitesimal generator is (a) a function $\Delta$ that maps states $X$ to a set $\Delta(X)$ of updates, and (b) preserves isomorphisms: if $X \cong_{\zeta} Y$ is an isomorphism of states $X, Y \in \mathcal{S}$, then for all updates $(f, \bar{a}, b) \in \Delta(X)$, we have an isomorphic update $(f, \overline{\zeta a}, \zeta b) \in \Delta(Y)$.

We write $\operatorname{Jump}(X)$ and say that $X$ is a $\operatorname{jump}$ when $\operatorname{Jump}(0)$ in timeline $\iota(X)$; otherwise, we write $\operatorname{Flow}(X)$ and say that it is a flow. For states $X$ with $\operatorname{Jump}(X)$, the following is natural:

Definition 7. The update generator is the infinitesimal generator defined on jump states $X$ as $\Delta(X)=\Delta^{+}(X)$, where $\Delta^{+}(X)$ stands for $\varphi\left(X, 0^{+}\right) \backslash X$.

To deal with flow states, we will also define some corresponding infinitesimal generator $\Delta_{\psi}$. Before doing so, let's see how to go from semantics to generators.

An initial evolution over $S$ is a function whose domain of definition is a timeline and whose range is $S$. An initial evolution is said to be initially constant if it has a constant prefix: that is to say, there is some $0<t$ such that the function is constant over $[0$.. $t]$.

Definition 8 (Semantics). $A$ semantics $\psi$ over a class $\mathcal{C}$ of sets $S$ is a partial function mapping initial evolutions over some $S \in \mathcal{C}$ to an element of $S$.

Remark 9. When $\mathbb{T}=\mathbb{R}^{\geq 0}$, an example of semantics over the class of sets $S$ containing $\mathbb{R}$ is the derivative $\psi_{\text {der }}$, mapping a function $f$ to its derivative at 0 when that exists. When $\mathbb{T}=\mathbb{N}$, an example of semantics over the class of all sets would be the function $\psi_{\mathbb{N}}$ mapping $f$ to $f(1)$. More generally, when $0 \in \mathbb{T}$ is such that $\operatorname{Jump}(0)$, an example of semantics over the class of all sets is the function $\psi_{\mathbb{N}}$ mapping $f$ to $f\left(0^{+}\right)$.

Consider a semantics $\psi$ over a class of sets $S$. Let $X$ be a state whose domain is in the class and a location $(f, \bar{a})$ of $X$. Denote by $\operatorname{Evolution}(X,(f, \bar{a}))$ the corresponding initial evolution: that is to say, the function given formally by Evolution $(X,(f, \bar{a})): t \mapsto \llbracket f(\bar{a}) \rrbracket_{\varphi(X, t)}$ for $0 \leq t \leq I_{1}, t \in \iota(X)$, for some $I_{1} \in \iota(X)$, with $I_{1}=0^{+}$for a jump. We use $\psi[X, f, \bar{a}]$ to denote the image of this evolution under $\psi$ (when it exists).

Definition 10 (Infinitesimal Generator Associated with $\psi$ ). The infinitesimal generator associated with $\psi$ maps each state $X$, such that $\psi[X, f, \bar{a}]$ is defined for all locations, to the set: $\Delta_{\psi}(X)=\{(f, \bar{a}, \psi[X, f, \bar{a}]) \mid$ $(f, \bar{a})$ is a location of $X$, Evolution $(X,(f, \bar{a}))$ is not initially constant $\}$.

The update generator $\Delta^{+}$(see Definition 7) is the infinitesimal generator associated with the semantics $\psi_{\mathbb{N}}$ (of Remark 9) over flow states.

From now on, we assume that some semantics $\psi$ is fixed to deal with flow states. It could be $\psi_{\text {der }}$, but it could also be another one (for example: talking about integrals or built using infinitesimals as in [19]). We denote by $\Delta_{\psi}$ the associated infinitesimal generator.

We are actually discussing algorithms relative to some $\psi$, and to be more precise, we should be refering to $\psi$-algorithms. The point is that not every infinitesimal generator is appropriate and that appropriateness is actually relative to a time domain and to the class of allowed dynamics over this time domain. To see this, keep in mind that - when $\Delta_{\psi}$ corresponds to derivative - to be able to talk about derivatives, one implicitly restricts oneself to dynamics that are differentiable, hence non-arbitrary. In other words, one is restricting to a particular class of possible dynamics, and not all dynamics are allowed. Restricting to other classes of dynamics (for example, analytic ones) may lead to different notions of algorithm.

From the update generator $\Delta^{+}$and $\Delta_{\psi}$, we build a generator also tagging states by the fact that they correspond to a jump or a flow:

Definition 11 (Generator of a State). We define the tagged generator of a state $X$, denoted $\Delta_{t}(X)$, as a function that maps state $X$ to $\{\mathcal{F}\} \times \Delta_{\psi}(X)$ when $\operatorname{Flow}(X)$ and $\Delta_{\psi}(X)$ is defined and to $\{\mathcal{J}\} \times \Delta^{+}(X)$ when Jump $(X)$.

Let $T$ be a set of ground terms. We say that states $X$ and $Y$ coincide over $T$, if $\llbracket s \rrbracket_{X}=\llbracket s \rrbracket_{Y}$ for all $s \in T$. This will be abbreviated $X={ }_{T} Y$. The fact that $X$ and $Y$ coincide over $T$ implies that $X$ and $Y$ necessarily share some common elements in their respective base sets, at least all the $\llbracket s \rrbracket_{X}$ for $s \in T$.

An algorithm should have a finite imperative description. Intuitively, the evolution of an algorithm from a given state is only determined by inspecting part of this state by means of the terms appearing in the algorithm description. The following corresponds to the Bounded Exploration postulate in [14].

Postulate III. For any algorithm, there exists a finite set $T$ of ground terms over vocabulary $\mathcal{V}$ such that for all states $X$ and $Y$ that coincide for $T, \Delta_{t}(X)$ and $\Delta_{t}(Y)$ both exist and $\Delta_{t}(X)=\Delta_{t}(Y)$.

A ground term of $T$ is a critical term and a critical element is the value (interpretation) of a critical term.

Definition 12 (Analog Algorithm). An algorithm is an object satisfying Postulates I through III.

## 3 Characterization Theorem

We now go on to define the rules of our programs (adding to those of ASM programs in [14]).

## Definition 13.

- Update Rule: An update rule of vocabulary $\mathcal{V}$ has the form $f\left(t_{1}, \ldots, t_{j}\right):=t_{0}$ where $f$ is a $j$-ary function symbol in $\mathcal{V}$ and $t_{1}, \ldots, t_{j}$ are ground terms over $\mathcal{V}$.
- Parallel Update Rule: If $R_{1}, \ldots, R_{k}$ are update rules of vocabulary $\mathcal{V}$, then par $R_{1} R_{2} \ldots R_{k}$ endpar is a parallel update rule of vocabulary $\mathcal{V}$.
$\Delta_{t}\left(R_{i}, X\right)$ denotes the interpretation of a rule $R$ in state $X$ and is defined as expected: If $R$ is an update rule $f\left(t_{1}, \ldots, t_{j}\right):=t_{0}$ then $\Delta_{t}(R, X)=$ $\{\mathcal{J}\} \times\left(f,\left(\llbracket t_{i} \rrbracket_{X}, \ldots, \llbracket t_{j} \rrbracket_{X}\right), \llbracket t_{0} \rrbracket_{X}\right)$ and when $R$ is par $R_{1}, \ldots, R_{k}$ endpar then $\Delta_{t}(R, X)=\{\mathcal{J}\} \times\left(d_{1} \cup \cdots \cup d_{k}\right)$ where $\Delta_{t}\left(R_{i}, X\right)=\{\mathcal{J}\} \times d_{i}$ for all $i$.

Next, we introduce rules to deal with Flows.

## Definition 14.

- Basic Continuous Rule: A basic continuous rule of vocabulary $\mathcal{V}$ has the form DYnamic $\left(f\left(t_{1}, \ldots, t_{j}\right), t_{0}\right)$ where $f$ is a symbol of arity $j$ and $t_{0}, t_{1}, \ldots, t_{j}$ are ground terms of vocabulary $\mathcal{V}$.
- Flow Rule: If $R_{1}, \ldots, R_{k}$ are basic continuous rules of vocabulary $\mathcal{V}$, then flow $R_{1} R_{2} \ldots R_{k}$ endflow is a flow rule of vocabulary $\mathcal{V}$.

Their semantics are then defined as follows. If $R$ is a basic continuous rule $\operatorname{Dynamic}\left(f\left(t_{1}, \ldots, t_{j}\right), t_{0}\right)$ then $\Delta_{t}(R, X)=\{\mathcal{F}\} \times\left\{\left(f,\left(a_{1}, \ldots, a_{j}\right), a_{0}\right)\right\}$ where each $a_{i}=\llbracket t_{i} \rrbracket_{X}$. If $R$ is a flow rule with constituents $R_{1}, \ldots, R_{k}$, then $\Delta_{t}(R, X)=$ $\{\mathcal{F}\} \times\left(d_{1} \cup \cdots \cup d_{k}\right)$ where $\Delta_{t}\left(R_{i}, X\right)=\{\mathcal{F}\} \times d_{i}$.

Finally, we allow conditionals:

## Definition 15.

- Selection Rule: If $\varphi$ is a ground boolean term over vocabulary $\mathcal{V}$ and $R_{1}$ and $R_{2}$ are rules of vocabulary $\mathcal{V}$ then: if $\varphi$ then $R_{1}$ or else $R_{2}$ endif is a rule of vocabulary $\mathcal{V}$.

Given such a rule $R$ and a state $X$, if $\varphi$ evaluates to true (the interpretation of scalar function true) in $X$ then $\Delta_{t}(R, X)=\Delta_{t}\left(R_{1}, X\right)$ else $\Delta_{t}(R, X)=$ $\Delta_{t}\left(R_{2}, X\right)$.

An ASM program of vocabulary $\mathcal{V}$ is a rule of vocabulary $\mathcal{V}$. The first key result is the following, which can be seen as a completeness result.

Theorem 16 (Completeness). For every algorithm of vocabulary $\mathcal{V}$, there is an $A S M$ program $\Pi$ over $\mathcal{V}$ with the identical behavior: $\Delta_{t}(\Pi, X)=\Delta_{t}(X)$ for all states $X$.

## 4 Extended Statements

We are now very close to formulating our other theorems. First we define an abstract state machine relative to semantics $\psi$.

Definition 17. $A \psi$-abstract state machine $B$ comprises the following: (a) an ASM program $\Pi$; (b) a set $\mathcal{S}$ of (first-order) structures over some finite vocabulary $\mathcal{V}$ closed under isomorphisms, and a subset $\mathcal{S}_{0} \subseteq \mathcal{S}$ closed under isomorphisms; (c) a map ८ and a map $\varphi$ such that $\left\langle\mathcal{S}, \mathcal{S}_{0}, \iota, \varphi\right\rangle$ is an algorithm, where $\Delta_{\psi}$ is fixed to be the infinitesimal generator associated with $\psi$, and for all states $X$ in $\mathcal{S}, \Delta_{t}(\Pi, X)=\Delta_{t}(X)$.

By definition, a $\psi$-abstract state machine $B$ satisfies all the postulates and hence is an algorithm.

Definition 18. An ASM program $\Pi$ is $\psi$-solvable for a set $\mathcal{S}$ of (first-order) structures over some finite vocabulary $\mathcal{V}$ closed under isomorphisms and a subset $\mathcal{S}_{0} \subseteq \mathcal{S}$ closed under isomorphisms if there exists a unique ८ and $\varphi$ such that $\left(\Pi, \mathcal{S}, \mathcal{S}_{0}, \iota, \varphi\right)$ is a $\psi$-abstract state machine.

Definition 19. A semantics $\psi$ is unambiguous if for all sets $\mathcal{S}$ of (first-order) structures over some finite vocabulary $\mathcal{V}$ closed under isomorphisms, and for all subsets $\mathcal{S}_{0} \subseteq \mathcal{S}$ closed under isomorphisms, whenever there exists some ८ and $\varphi$ such that $\left(\Pi, \mathcal{S}, \mathcal{S}_{0}, \iota, \varphi\right)$ is a $\psi$-abstract state machine, then ८ and $\varphi$ are unique.

Our main results follow. (Proofs are relegated to the technical report [6].)
Theorem 20. For every $\psi$-definable algorithm $A$, there exists an equivalent $\psi$-abstract state machine $B$.

Theorem 21. Assume that $\psi$ is unambiguous. For every $\psi$-definable algorithm $A$, there exists a unique equivalent $\psi$-abstract state machine $B$ with the same states and initial states.

Corollary 22. Assume that $\psi$ is unambiguous. For every $\psi$-definable algorithm A, there exists an equivalent $\psi$-solvable ASM program.

To any algorithm $A$ that is $\psi$-definable there corresponds an equivalent $\psi$-abstract state machine $B$, and hence a $\psi$-solvable program $\Pi$. Conversely, a $\psi$-abstract state machine $B$ corresponds to a $\psi$-definable algorithm. However, not every program $\Pi$ is $\psi$-solvable.

When $\psi$-corresponds to $\psi_{\text {der }}$, unambiguity comes from (unicity in) the Cauchy-Lipschitz theorem. The fact that not every program $\Pi$ is $\psi$-solvable is due to the fact that not all differential equations have a solution.

## 5 Examples

The examples in this section are for semantics $\psi_{\text {der }}$. Our settings cover, first of all, analog algorithms that are pure flow, in particular all systems that can be modeled as ordinary differential equations. A very simple, classical example is the pendulum: the motion of an idealized simple pendulum is governed by the second-order differential equation $\theta^{\prime \prime}+\frac{g}{L} \theta=0$, where $\theta$ is angular displacement, $g$ is gravitational acceleration, and $L$ is the length of the pendulum rod. This can indeed be modeled as the program
flow Dynamic $\left(\theta, \theta_{1}\right) \quad$ Dynamic $\left(\theta_{1},-\frac{g}{L} \cdot \theta\right)$ endflow
using the fact that any ordinary differential equation can be put in the form of a vectorial first-order equation, here $\theta_{1}$ corresponding to the derivative of $\theta$.

As a consequence, our formalism covers very generic classes of continuoustime models of computation, including the GPAC, which corresponds to ordinary differential equations with polynomial right-hand sides [12,13]. Recall that the


Fig. 1. A GPAC for sine and cosine (left). Corresponding evolution (right).

GPAC was proposed as a mathematical model of differential analyzers (DAs), one of the most famous analog computer machines in history. Figure 1 (left) depicts a (non-minimal) GPAC that generates sine and cosine. In this picture, $\int$ signifies some integrator, and -1 denotes some constant block. This simple GPAC can be modeled by the program
flow $\operatorname{Dynamic}(x, z)$ Dynamic $(y, x)$ Dynamic $(z,-x)$ endflow
Our proposed model can also adequately describe hybrid systems, made of alternating sequences of continuous evolution and discrete transitions. This includes, for example, a simple model of a bouncing ball, the physics of which are given by the flow equations $x^{\prime \prime}=-g m$, where $g$ is the gravitational constant and $v=x^{\prime}$ is the velocity, except that upon impact, each time $x=0$, the velocity changes according to $v^{\prime}=-k \cdot v^{\prime}$, where $k$ is the coefficient of impact. Every time the ball bounces, its speed is reduced by a factor $k$. This system can be described by a program like
if $x=0$ then $\quad v:=-k \cdot v$
else flow $\operatorname{Dynamic}(x, v)$ $\operatorname{Dynamic}(v,-g \cdot m)$ endflow endif
Our setting is an extension of classical discrete-time algorithms; hence, all classical discrete-time algorithms can also be modeled.

As for examples with semantics other than $\psi_{\text {der }}$ : Observe that one can consider timelines like $\mathbb{Q}$ instead of $\mathbb{R}$. (For such a timeline, we have $\operatorname{Flow}(i)$ for all $i \in \mathbb{Q}$.) One can define a semantics on such a timeline where for every state $X$ we have $\operatorname{Flow}(X)$ by first extending the evolution function to $\mathbb{R}$ (for example by restricting to continuous dynamics) and then using the derivative. Constructions of [19] are also covered by our settings: In some sense, the example at the beginning of the paragraph is the spirit of the constructions from [19], where the timeline is the set of hyperreals obtained by multiplying some fixed infinitesimal by some hyperinteger (using hyperreals and infinitesimals). Notice that there is no need to consider derivatives or similar notions: we could also consider analytic dynamics, and consider a semantics related to the family of Taylor coefficients. Weaker notions of solution, like variational approaches, can also be considered.

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# Generalized Effective Reducibility 

Merlin $\operatorname{Carl}{ }^{(\boxtimes)}$<br>Fachbereich für Mathematik und Statistik der Universität Konstanz, Konstanz, Germany<br>merlin.carl@uni-konstanz.de


#### Abstract

We introduce two notions of effective reducibility for settheoretical statements, based on computability with Ordinal Turing Machines (OTMs), one of which resembles Turing reducibility while the other is modelled after Weihrauch reducibility. We give sample applications by showing that certain (algebraic) constructions are not effective in the OTM-sense and considering the effective equivalence of various versions of the axiom of choice.


## 1 Introduction

From a sufficiently remote point of view, construction problems in mathematics can be seen as multi-valued, class-sized 'functions' from the set-theoretical universe $V$ to itself. Examples of construction problems would be the problem assigning to fields their algebraic closures, to sets their well-orderings, to integrable functions their stem functions, to linear orderings their completions etc. Formally, this makes a construction problem a (class-sized) relation $R \subseteq V \times V$.

A 'solution' to or 'canonification' of a construction problem $R$ is then a (classsized) witness 'function' $F: V \rightarrow V$ such that, for all $x$ in the domain of $R$, we have $R(x, F(x))$ and otherwise $F(x)=\emptyset$. Similarly, we can say that $F$ witnesses the truth of a set-theoretical statement $\phi$ of the form $\forall x \exists y \psi$ if $F$ is a solution for $\{(x, y): \psi(x, y)\}$, the most natural candidates to consider being $\Pi_{2}$-statements, since $\psi$ can be assumed to be absolute between transitive sets in that case.

Fixing an appropriate notion of effectiveness for set-theoretical constructions, we can now ask for specific construction problems $R$ whether there exists an effective solution for $R$ and similarly, whether some statement $\phi$ is 'effectively true'. Moreover, we can ask whether a construction or a statement 'effectively reduces' to another.

In the following, 'effectiveness' will be interpreted to mean computability by Ordinal Turing Machines (OTMs) without ordinal parameters. For the definition and basic results on OTMs, we refer the reader to [Ko1]. It was argued in [Ca] that OTM-computations are appropriate as a formalization of the intuitive notion of a 'transfinite effective procedure'.

For convenience, we assume that our machines work with three tapes, a 'miracle' tape (to be explained below), a scratch tape and an output tape.

Our set-theoretical notation is standard and can e.g. be looked up in [Je].

## 2 Basic Methods and Notions

To apply OTMs to general mathematical constructions, we need to represent arbitrary sets as an input for OTMs, i.e. as sets of ordinals. This can be achieves as follows:

Definition 1. Let $x$ be a set, $t=t c(x)$ the transitive closure of $x, \alpha \in O n$ and $f: \alpha \rightarrow t c(x)$ a well-ordering of $t c(x)$ in the order type $\alpha$. We define $c_{f}(x)$, the $f$-code for $x$, recursively as the following set of ordinals: $c_{f}(x):=\left\{p\left(f^{-1}(y), \beta\right):\right.$ $\left.y \in x \wedge \beta \in c_{f}(y)\right\}$, where $p$ denotes Cantor's ordinal pairing function. We say that $A \subseteq O n$ 'is a code for' or 'codes' the set $x$ if and only if there is some $f$ for which $A=c_{f}(x)$. We write $\operatorname{rep}(\tau, x)$ to indicate that $\tau$ codes $x$.

Remark: By a certain abuse of notation, if $x$ is a set, we will sometimes write $c(x)$ for an 'arbitrary' code for $x$. Note that $|c(x)|=|\operatorname{tc}(x)|$ if $\operatorname{tc}(x)$ is infinite.

We can now talk about OTM-computability of arbitrary functions from $V$ to $V$ :

Definition 2. Let $F: V \rightarrow V$ be a functional class. We say that $F$ is OTMcomputable if and only if there is an OTM-program $P$ such that, for every set $x$ and every tape content $\tau$, if $\operatorname{rep}(\tau, x)$, then $P(\tau)$ converges to output $\sigma$ such that $\operatorname{rep}(\sigma, F(x))$, i.e. $P$ takes representations of $x$ to representations of $F(x)$.

By this definition, the representation of a set $x$ will depend on the choice of a well-ordering of $\operatorname{tc}(x)$. The output of a computation on input $x$ may hence depend on the choice of the representation of $x$. This is fine as long as only the output, but not the object coded by the output, depends on the choice of the input representation.

This allows us to make our notion of 'effectivity' precise:
Definition 3. Let $R \subseteq V \times V$ be a construction problem. Then $R$ is effectively solvable if and only if there is an OTM-computable solution $F$ for $R$. Moreover, a set-theoretical $\Pi_{2}$-statement $\forall x \exists y \phi(x, y)$ (where $\phi$ is $\Delta_{0}$ ) is effective if and only if the construction problem $\{(x, y) \in V \times V: \phi(x, y)\}$ is effectively solvable. ${ }^{1}$ We write $R_{x}$ for $\{y:(x, y) \in R\}$.

One may now inquire whether various well-known construction problems and $\Pi_{2}$-statements are effective. Such questions were studied by Hodges in [Ho2], though with a different notion of effectivity based on Jensen and Karp's primitive recursive set functions. We note here that the two methods Hodges uses also work for our model, which allows us to carry over results.

The following lemma corresponds to Hodges' 'cardinality method', i.e. Lemma 3.2 of [Ho2]:

[^33]Lemma 1. Let $\alpha \in$ On, and let $R \subseteq V \times V$ be such that, for some cardinal $\kappa>\alpha$, there is $x \in V$ such that $|t c(x)|=\kappa, R_{x} \neq \emptyset$ and $\forall y \in R_{x}|y|>\kappa$. Then no witness function for $R$ is OTM-computable in the parameter $\alpha$.
Consequently, if $R$ is such that there are such $\kappa$ and $x$ for every $\alpha \in O n$, then no witness function for $R$ is computable by a parameter-OTM (i.e. an OTM with a fixed tape cell marked with 1).
In particular, if, for some transitive $x$ of infinite cardinality, $R_{x} \neq \emptyset$ and $\forall y \in R_{x}|y|>|x|$ then no witness function for $R$ is parameter-free OTMcomputable.

Proof. Clearly, in less than $\kappa^{+}$many steps, the machine cannot write a code of a structure of cardinality $>\kappa$.

It hence suffices to show that, when $P$ is an OTM-program and $P$ is given a (code $c$ of a) set $x$ of size $\kappa \geq \omega$ for input and the computation halts, then the output of the computation will be of size $\leq \kappa$. This follows if we can show that the computation will take less than $\kappa^{+}$many steps, since $P$ can write at most $\alpha$ many symbols in $\alpha$ many steps. Suppose for a contradiction that $P$ takes $\lambda>\kappa$ many steps, and let $\delta$ be the smallest cardinal $>\lambda$. Let $H$ be the $\Sigma_{1}$-Skolem hull of $\kappa \cup\{c\}$ in $L_{\delta}[c]$ and let $M$ denote the transitive collapse of $H$. We may assume without loss of generality that $c \subseteq \kappa$, so that we have $c \in M$; as $L_{\delta}[c]$ contains the computation of $P$ in the input $c$, so does $H$ and hence there is $S \in M$ such that $M$ believes that $S$ is the computation of $P$ with input $c$. By transitivity of $M$ and absoluteness of computations, $S$ is actually the computation of $P$ with input $c$. Since $S$ is contained in a transitive set of cardinality $\kappa,|S| \leq \kappa$, so the length of the computation is $<\kappa^{+}$, as desired.

There is also an analogue of the 'forcing method' (Lemma 3.7 of [Ho2]), which is given in Lemma 5 below.

Convention: For many of the following results, we will need the existence of generic filters for various partial orderings in $L$ and some of its (symmetric) extensions. To avoid technical complications, we use as a shortcut an extra assumption that guarantees the existence of such filters. $0^{\sharp}$ is more than enough for our purposes, and we assume from now on that it exists. ${ }^{2}$

These lemmata can be seen as expressing the intuition that neither the power set operation on infinite sets nor the use of the axiom of choice are 'effective', not even in a very idealized sense. We note some sample applications.

Lemma 2. None of the following construction problems is effectively solvable: (1) Field to its algebraic closures (2) Linear ordering to its completions (3) Set to its (constructible) power set (4) Set to its well-orderings

Proof. (1) can be proved by an easy adaption of the proof of Theorem 4.1 of [Ho2]. There is only one point that requires a little care, namely the use of

[^34]countable transitive models in that proof: For it might happen that an OTMprogram $P$ that halts in $V$ does not halt in such a model $M .{ }^{3}$ However, a check of Hodge's proof reveals that the countability of the ground model serves no purpose but to guarantee the existence of generic filters. We can hence circumvent this problem by doing the construction over $L$, using $0^{\sharp}$ to guarantee the existence of the required filters. (2) and (3) are easy applications of Lemma 1. (4) follows from Lemma 6 below.

It is, on the other hand, not hard to see that e.g. the construction problem of taking a ring to its quotient field is effectively solvable as in [Ho2]. The intuitions captured by Hodges' approach are hence preserved in our framework.

There are certainly various interesting questions to be asked about the effectivity, or otherwise, of various construction problems or $\Pi_{2}$-statements. However, we want to take the analogy with Turing computability a bit further: Instead of merely asking what problems are solvable, we want to consider what problems/statements are effectively reducible to which others in the sense that, given access to a solution to one as an 'oracle', one can effectively solve the other. A quite straightforward way to make this idea precise is the following:

Definition 4. Assume that the OTM is equipped with an extra 'miracle tape'. Let $F$ be a class function taking sets of ordinals to sets of ordinals. A miracle-OTM-program is defined like an OTM-program, but with an extra 'miracle' command. When this command is carried out, the set $X$ of ordinals on the miracle tape is replaced (in one step) by $F(X)$. We write $P^{F}$ to indicate that $P$ is run and whenever the miracle command is applied to $X$, it is replaced by $F(X) .{ }^{4}$

Definition 5. Let $C_{1}$ and $C_{2}$ be construction problems. Then $C_{1}$ is reducible to $C_{2}$, written $C_{1} \leq C_{2}$, if and only if there is some miracle-OTM-program $P$ such that the following holds: Whenever $F$ is a canonification of $C_{2}$ and whenever $G: V \rightarrow V$ is a class function taking each code for a set $x$ to some code $y$ for $F(x)$ and $x$ is a set and $c$ a code for $x$, we have $P^{G}(c) \downarrow=d$, where $d$ is a code for a set $z$ such that $(x, z) \in C_{1}$.

Remark: Note that we do not demand in the conditions on $G$ that $G(c)$ depends only on $x$ when $c$ is a code for $x$. By demanding that the same reduction works for every $G$, we rule out the possibility of coding extra information into the input representations.

Concerning this notion of reducibility, we observe that certainly a cardinalityraising construction is not reducible to one that is not:

[^35]Lemma 3. Let $C_{1}, C_{2}$ be construction problems. Assume that there are some canonification $F$ of $C_{2}$ and some infinite set $x$ such that, for all sets $y$, (1) if $C_{1}(x, y)$, then $|y|>|t c(x)|$ and (2) if $y$ is infinite, then $|F(y)| \leq|t c(y)|$. Then $C_{1} \not \leq C_{2}$.

Proof. As in the proof of Lemma 1 above, OTM-computable functions cannot raise infinite cardinalities. By assumption, the miracle operation will also not raise the cardinality. Hence the output of a program $P$ with a $C_{2}$-miracle will (for infinite input) always have at most the cardinality of the input and thus cannot in any case witness $C_{1}$.

Remark: In particular, the construction problem of taking a valued field to its linear compactifications (see [Ho2], Theorem 4.10) is not reducible to any of the following construction problems: Field to algebraic closure, formally real field to its real closure, field of characteristic $p$ to its separable algebraic closure.

The above captures the idea that one construction 'helps' carrying out another. There is also a much more restrictive intuitive notion of reducibility between problems, namely that instances of one (construction) problem can be effectively 'translated' to particular instances of another: Given an instance of a problem $C_{1}$, we can first effectively turn it into an instance of a problem $C_{2}$ and then effectively turn the solution to $C_{2}$ into a solution to $C_{1}$. Another way to view this is that $C_{2}$ may only be used once in solving $C_{1}$. Thus, we define:

Definition 6. Let $C_{1}, C_{2}$ be construction problems. Then $C_{1}$ is generalized Weihrauch reducible to $C_{2}$, written $C_{1} \leq_{g W} C_{2}$, if and only if there are OTMprograms $P$ and $Q$ such that the following holds for all sets $x$ in the domain of $C_{1}$, every code c for $x$ and every canonification $F$ of $C_{2}$ :
(1) $Q(c)$ converges to output $c^{\prime}$, where $c^{\prime}$ is a code for a set $y$
(2) For every code $c^{\prime \prime}$ of $F(y), P\left(c^{\prime \prime}\right)$ converges to output $c^{\prime \prime \prime}$, where $c^{\prime \prime \prime}$ is a code for a set $z$
(3) We have $C_{1}(x, z)$

If these clauses hold, we say that $(P, Q)$ witnesses the $g W$-reducibility of $C_{1}$ to $C_{2}$. Also, when $F$ is a canonification, $P$ and $Q$ are OTM-programs and $x$ is a set, we write $[P, F, Q](x)$ for the $z$ obtained by the procedure just described.

If $C_{1} \leq_{g W} C_{2}$ and $C_{2} \leq_{g W} C_{1}$, we write $C_{1} \equiv_{g W} C_{2}$.
Remark: The name of the notion is due to its obvious resemblance with Weihrauch reducibility, which is an analogous notion for classical computability. For some results on classical Weihrauch reducibility, see e.g. [BGM]. Another notion of Weihrauch reducibility on higher cardinalities based on continuous transformations on generalized Baire spaces is introduced in [Ga]. We do not know how this approach connects to ours. ${ }^{5}$
${ }^{5}$ The generalization of uniform reducibility introduced in [HJ] based on (Turing-) computable strategies goes into a different direction, though it may be compatible with our setting. We thank one of our anonymous referees for pointing out this paper as well as the work of Galeotti to us.

We note that reducibility notions satisfy the general order-theoretic properties of reducibility relations:

Lemma 4. Both $\leq$ and $\leq_{g W}$ are transitive and reflexive. Consequently, $\equiv_{g W}$ and $\equiv$ are equivalence relations.

Definition 7. Let $C$ be a construction problem. Then $[C]$ denotes the $\equiv$ equivalence class of $C$ and $[C]_{g W}$ denotes the $\equiv_{g W^{-}}$equivalence class of $C$.

## 3 A Method for Negative Results

We develop a method for showing that a construction problem is not gWreducible to another. We will work with class-sized models of $\mathrm{ZF}^{-}$, which denotes Zermelo-Fraenkel set theory without the axiom of powerset; more precisely, we take the formulation of $\mathrm{ZF}^{-}$given in [GH].

Remark: Note that the following theorem is not trivial even when $\mathrm{ZF}^{-}$is strengthened to full ZF, since a ZF model $M$ may contain a set $x$ without containing a suitable input format for $x$, so that the computation of an OTM cannot be simulated within $M$.

Lemma 5. Let $M \models Z F^{-}$be transitive and suppose that $x \in M$. Then $\mathbb{P}_{x}:=$ $\{f: \omega \rightarrow x:|f|<\omega \wedge f$ injective $\}$ is a set in $M$.

Proof. Let $y:=x \times \omega$. For each $n \in \omega$, we have $y^{n} \in M$ and the function $F: \omega \rightarrow M$ that maps $n$ to $y^{n}$ is definable in $M$. By replacement and union, $A:=\bigcup\left\{y^{n}: n \in \omega\right\} \in M$. Now $P_{x}$ can be obtained from $A$ via separation.

Theorem 8. Let $F$ be a computable class function, $M \models Z F^{-}$transitive such that $O n^{M}=O n^{6}$. Assume moreover that $x \in M$ is such that there are (in $V$ ) two mutually generic $\mathbb{P}_{x}$-generic filters $G_{1}$ and $G_{2}$ over $M$. Then $F(x) \in M$.

Proof. Let $P$ be a program witnessing the computability of $F$. Let $x \in M$ be as in the assumption of the Theorem. By passing to $\operatorname{tc}(x)$ if necessary, we may assume without loss of generality that $x$ is transitive. Let $G_{1}, G_{2}$ be mutually $M$-generic filters over $\mathbb{P}_{x}$ which exist by assumption. In $M\left[G_{1}\right]$ and $M\left[G_{2}\right], x$ is well-ordered in order type $\alpha$ by $\bigcup G_{1}$ and $\bigcup G_{2}$, respectively. Hence both $M\left[G_{1}\right]$ and $M\left[G_{2}\right]$ contain tape contents coding $x$ and thus contain the computations of $P$ on these inputs. As $\mathrm{ZF}^{-}$models, $M\left[G_{1}\right]$ and $M\left[G_{2}\right]$ contain the decoding of every tape content they contain. Thus $F(x) \in M\left[G_{1}\right] \cap M\left[G_{2}\right]$. As $G_{1}$ and $G_{2}$ are mutually generic, we have $M\left[G_{1}\right] \cap M\left[G_{2}\right]=M$, so $F(x) \in M$, as desired.

[^36]This suggests a general method for proving, given constructions $C_{1}$ and $C_{2}$, that $C_{1} Z_{\mathrm{gW}} C_{2}$. In general, find a class $A$ sufficiently closed under OTMcomputability and a canonification $F$ of $C_{2}$ such that there is some $x \in A$ with the property that the closure of $F[A]$ under OTM-computability does not contain a $C_{1}$-solution for $x$. By Theorem 8, we can take for $A$ a transitive class model $M$ of $\mathrm{ZF}^{-}$. We summarize the most important special case of this method in the following lemma:

Lemma 6. Let $C_{1}, C_{2}$ be construction problems. Assume that there are a canonification $F$ of $C_{2}$ and a transitive class-sized $M \models Z F^{-}$and some $x \in M \cap \operatorname{dom}\left(C_{1}\right)$ such that $M$ is closed under $F$, but $\left\{y: C_{1}(x, y)\right\} \cap M=\emptyset$. Assume moreover that $x$ is such that there are (in $V$ ) two mutually generic $\mathbb{P}_{x}$-generic filters $G_{1}$ and $G_{2}$ over $M$. Then $C_{1} \not \not_{g W} C_{2}$.

Proof. Assume otherwise, and let $P$ and $Q$ be OTM-programs such that $(P, Q)$ witnesses the gW-reducibility of $C_{1}$ to $C_{2}$. Pick $F, M$ and $x$ as in the statement of the Lemma. Then $Q$ computes, for every code of $x$ as an input, a code for some (unique) set $y$. By Theorem 8, we have $y \in M$. As $M$ is closed under $F$, we have $F(y) \in M$. Now, for every code of $F(y)$ as an input, $P$ computes a code for some (unique) set $z$. Again by Theorem $8, z \in M$. Also, by the choice of $P$ and $Q$, we have $C_{1}(x, z)$. So $z \in\left\{y: C_{1}(x, y)\right\} \cap M$ and so the latter is not empty, a contradiction.

## 4 Results on Generalized Effective Reducibility

As a sample application of the notions and methods developed above, we consider variants of the axiom of choice with respect to effective reducibility.

Definition 9. Denote by $A C$ the statement that for all sets $x$, there is a function $f$ such that $f(\emptyset)=\emptyset$ and for $y \in x$, if $y \neq \emptyset$, then $f(y) \in y$. Denote by $A C$ the statement that for all sets $x$ whose elements are non-empty and mutually disjoint, there is a set $r$ such that $|r \cap y|=1$ for all $y \in x$. Denote by WO the well-ordering principle, i.e. the statement that for every set $x$, there is an ordinal $\alpha$ and a bijection $f: \alpha \leftrightarrow x$. Denote by ZL Zorn's lemma, i.e. the statement that, for every partially ordered set $(X, \leq)$ in which every ascending chain has an upper bound, there is a $\leq$-maximal element in $X$. Finally, denote by HMP the Hausdorff maximality principle, i.e. the statement that, for every partially ordered set $(X, \leq)$, every totally ordered subset is contained in a $\subseteq$-maximal totally ordered subset.

Remark: HMP can be seen as the combinatorial core behind the proofs of ZL from AC or WO. For effective reducibility, it is the more interesting formulation: For if ZL holds, one can obtain the set of maximal elements of a p.o.-set ( $X, \leq$ ) satisfying the conditions of ZL by merely searching through $X .^{7}$

[^37]It is not hard to see that all these principles are equivalent in the sense of reducibility: The usual equivalence proofs explain, modulo a transfinite version of Church's thesis, how each of these principles can be reduced to any other. However, for gW -reducibility, we can use Lemma 6 to show that (under $0^{\sharp}$ ) the well-ordering principle is not generalized Weihrauch reducible to the axiom of choice:

Theorem 10. If $0^{\sharp}$ exists, then $W O \not \leq_{g W} A C$.
Proof. (Sketch) In Theorem D.-A.C. of [Z], it is shown how to construct a transitive model of $\mathrm{ZF}^{-}+\mathrm{AC}+\neg \mathrm{WO}$ as a union of an ascending chain of symmetric extensions of a transitive ground model $M$ of $\mathrm{ZF}^{-}$. Starting with $M=L$, it is easily checked that, under the assumption that $0^{\sharp}$ exists, the construction leads to a definable transitive class model $N$ of $\mathrm{ZF}^{-}+\mathrm{AC}$ such that some set $A \in N$ that is non-wellorderable in $N$ is countable in $V$ and moreover $\mathbb{P}_{A}$ is countable and thus has two mutually generic filters over $N$. Hence the assumptions of Lemma 6 are satisfied and the non-reducibility follows.

Many other relations between choice principles are effective, however:
Theorem 11. We have (1) $A C \equiv_{g W} A C \leq_{g W} Z L$ and (2) $Z L \leq_{g W} W O$
Proof. The usual equivalence proofs over ZF (see e.g. [Je]) effectivize.

Remark: We do not know how HMP relates to AC in terms of $\leq_{g W}$. We suspect that HMP $\not Z_{\mathrm{gW}}$ AC. Our current state of knowledge (combined with our remark on HMP above) hence gives some meaning to the humorous claim that 'The Axiom of Choice is obviously true, the well-ordering principle obviously false, and who can tell about Zorn's lemma?'s.

## 5 Conclusion and Further Work

We have seen how our framework can be used to distinguish between various versions of set-theoretical principles usually regarded as equivalent.

One may now ask for many statements which are effectively reducible or gW-reducible to which others. This may be viewed as a cardinality-independent version of reverse mathematics (see [Sh]) and the theory of the Weihrauch lattice. Apart from that, it may be interesting to consider variants with OTMcomputability replaced by other models like ITTMs [HL] or parameter-OTMs, along with variations of Weihrauch reducibility, such as strong Weihrauch reducibility.

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[^38]
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# Lightface $\Pi_{3}^{0}$-Completeness of Density Sets Under Effective Wadge Reducibility 

Gemma Carotenuto ${ }^{1(\boxtimes)}$ and André Nies $^{2}$<br>${ }^{1}$ Dipartimento di Matematica, Universita di Salerno, Fisciano, Italy<br>gcarotenuto@unisa.it<br>2 Department of Computer Science, University of Auckland, Auckland, New Zealand


#### Abstract

Let $\mathcal{A} \subseteq{ }^{\omega} 2$ be measurable. The density set $D \mathcal{A}$ is the set of $Z \in{ }^{\omega} 2$ such that the local measure of $\mathcal{A}$ along $Z$ tends to 1 . Suppose that $\mathcal{A}$ is a $\Pi_{1}^{0}$ set with empty interior and the uniform measure of $\mathcal{A}$ is a positive computable real. We show that $D \mathcal{A}$ is lightface $\Pi_{3}^{0}$ complete for effective Wadge reductions. This is an algorithmic version of a result in descriptive set theory by Andretta and Camerlo [1]. They show a completeness result for boldface $\Pi_{3}^{0}$ sets under plain Wadge reductions.


## 1 Introduction

We work in Cantor space ${ }^{\omega} 2$ with the product measure $\mu$. For a finite bit string $s$ and a measurable set $\mathcal{A} \subseteq{ }^{\omega}$, the local measure of $\mathcal{A}$ above $s$ is

$$
\mu_{s}(\mathcal{A})=2^{-|s|} \cdot \mu([s] \cap \mathcal{A})
$$

where $[s]$ is the clopen set in ${ }^{\omega} 2$ consisting of all the extensions of $s$ (this set is often denoted $N_{s}$ in the literature). Let $D \mathcal{A}$ be the points $Z$ of such that $\mathcal{A}$ has density 1 at $Z$, namely

$$
D \mathcal{A}=\left\{Z: \lim _{n} \mu_{Z \upharpoonright n}(\mathcal{A})=1\right\} .
$$

We call density set every set of the form $D \mathcal{A}$, for some measurable $A$. We will be mainly interested in closed $\mathcal{A}$, in which case $D \mathcal{A} \subseteq \mathcal{A}$.

The Lebesgue density theorem for Cantor space states that if $\mathcal{A}$ is measurable, then almost every $Z \in \mathcal{A}$ is in $D \mathcal{A}$. This result has been the seed for recent investigations both in algorithmic randomness and in descriptive set theory.

In algorithmic randomness, density has been instrumental in solving the longopen "covering problem". Combining Bienvenu et al. [3] and Day and Miller [5] yielded the answer; for an overview see [2]. Khan [6] obtained a variety of results, in particular relating density for the reals with density for Cantor space. In a recent article, Myabe et al. [7] define density randomness of a point $Z$ in Cantor space as the combination of Martin-Löf-randomness of $Z$ and the property that $Z \in D \mathcal{P}$ for each $\Pi_{1}^{0}$ set $\mathcal{P}$ containing $Z$; they show the equivalence of density randomness with a number of notions stemming from effective analysis. Martin-Löf-randomness of $Z$ does not necessarily imply that $Z$ satisfies the effective
version of Lebesgues's theorem; for instance, the least element of a non-empty $\Pi_{1}^{0}$ set $\mathcal{P}$ of ML-randoms is not in $D \mathcal{P}$.

In descriptive set theory Andretta and Camerlo [1] have analyzed the complexity of the density sets. It is easily seen that the $\boldsymbol{\Pi}_{3}^{0}$ pointclass is an upper bound for this study. For example, if $A=[s]$, then $D \mathcal{A}=A$, which is closed. In $[1$, Sect. 7$]$ the authors show that $D \mathcal{A}$ is $\boldsymbol{\Pi}_{3}^{0}$-complete with respect to Wadge reducibility in case that $\mathcal{A}$ has empty interior and positive measure. They also prove that density sets can have any complexity within $\boldsymbol{\Pi}_{3}^{0}$. The first author in [4] has conducted a similar study of the difference hierarchy over the closed sets in the setting of the real line with the Lebesgue measure.

The goal of this short paper is to connect the two approaches to Lebesgue's theorem. We give an algorithmic version of the result in [1]. We show that for any (lightface) $\Pi_{1}^{0}$ set $\mathcal{A} \subseteq{ }^{\omega} 2$ with empty interior and measure $\mu \mathcal{A}$ a positive computable real, the density set $D \mathcal{A}$ is lightface $\Pi_{3}^{0}$ complete for algorithmic Wadge reductions.

To say that a real $r$ is computable means that from a number $n \in \omega$ we can compute a rational that is within $2^{-n}$ of $r$. The algorithmic version of Wadge reducibility is as follows: for $\mathcal{C}, \mathcal{D} \subseteq{ }^{\omega} 2$, we write $\mathcal{D} \leq_{m} \mathcal{C}$ if there is a total Turing functional $\Psi$ such that $\mathcal{D}=\Psi^{-1}(\mathcal{C})$. Totality means that $\Psi(Y) \in{ }^{\omega} 2$ for each $Y$; equivalently, $\Psi(Y ; n)$ is obtained by evaluating a truth table computed from $n$ on $Y$. Note that it is easy to construct such a set $\mathcal{A}$, for instance by a Cantor space version of the construction of a Cantor set of positive measure in the unit interval.

In our proof, while we import the basic combinatorics of the approach in [1], the details are more complicated because we need to build an algorithmic Wadge reduction. This is where we use the hypothesis that the measure of $\mathcal{A}$ is computable. It is not known at present whether this hypothesis is actually necessary. The other hypothesis (that the interior be empty) is of course necessary, as for instance shown by taking $\mathcal{A}$ to be the whole Cantor space.

## 2 Completeness for Lightface Point Classes

For the basics about arithmetical hierarchy see for example $[9,10]$.
Definition 1. Consider a lightface point class $\Gamma$ in Cantor space, such as $\Pi_{2}^{0}$ or $\Pi_{3}^{0}$. We say that $\mathcal{C}$ is $\Gamma$-complete if $\mathcal{C}$ is in $\Gamma$, and for each $\mathcal{D} \in \Gamma$ we have $\mathcal{D} \leq{ }_{m} \mathcal{C}$.

The next result is folklore (see e.g. [8]). For the reader's benefit we provide the short proof.

Proposition 2. The class $\mathcal{C}$ of all sequences with infinitely many 1's is $\Pi_{2}^{0}$ complete.

Proof. The class $\mathcal{C}$ is clearly $\Pi_{2}^{0}$. Now suppose $\mathcal{D}$ is $\Pi_{2}^{0}$, so $\mathcal{D}=\bigcap_{n} \mathcal{G}_{n}$ with $\mathcal{G}_{n+1} \subseteq \mathcal{G}_{n}$ for a uniformly $\Sigma_{1}^{0}$ sequence $\left\langle\mathcal{G}_{n}\right\rangle_{n \in \omega}$. Define a total Turing functional
$\Psi$ as follows. Given a sequence of bits $Z$, at stage $t$ we use the first $t$ bits of $Z$, and append one bit at the end of the output $\Psi^{Z}$. We append 0 unless we see at stage $t$ that $\left[Z \upharpoonright_{t}\right] \subseteq \mathcal{G}_{n}$ for the next $n$; in that case we append 1 .

In the following we effectively identify ${ }^{\omega \times \omega} 2$ and ${ }^{\omega} 2$ via the standard computable pairing function $\omega \times \omega \rightarrow \omega$. The following is presumably folklore.

Proposition 3. The set $\mathcal{E}=\left\{Z \in{ }^{\omega} 2: \forall n \forall^{\infty} k Z(n, k)=0\right\}$ is $\Pi_{3}^{0}$-complete.
Proof. Clearly $\mathcal{E}$ is $\Pi_{3}^{0}$. Now suppose a given set $\mathcal{F}$ is $\Pi_{3}^{0}$, so $\mathcal{G}=\bigcap_{n} \mathcal{G}_{n}$ where $\mathcal{G}_{n}$ is uniformly $\Sigma_{2}^{0}$. By the previous proposition and the uniformity in $\mathcal{D}$ of its completeness part, for each $n$ we effectively have a Turing functional $\Psi_{n}$ such that $\mathcal{G}_{n}=\Psi_{n}^{-1}\left({ }^{\omega} 2 \backslash \mathcal{C}\right)$. Define a total Turing functional $\Psi:{ }^{\omega} 2 \rightarrow{ }^{\omega \times \omega} 2$ by

$$
\Psi^{Z}(n, r)=\Psi_{n}^{Z}(r)
$$

Clearly $Z \in \mathcal{F} \Leftrightarrow \forall n\left[Z \in \mathcal{G}_{n}\right] \Leftrightarrow \forall n\left[\Psi_{n}^{Z} \in{ }^{\omega} 2 \backslash \mathcal{C}\right] \Leftrightarrow \Psi^{Z} \in \mathcal{E}$.

## 3 Reaching the Maximal Complexity

Theorem 4. Let $\mathcal{A} \subseteq{ }^{\omega} 2$ be a $\Pi_{1}^{0}$ set with empty interior such that $\mu \mathcal{A}>0$ and $\mu \mathcal{A}$ is a computable real. Then $D \mathcal{A}$ is $\Pi_{3}^{0}$-complete.

We begin with some preliminaries. In the following $s, t, u$ denote strings of bits. Note that if $\mu \mathcal{A}$ is a computable real then $\mu_{s}(\mathcal{A})$ is a computable real uniformly in $s$ (see e.g. [9, 1.9.18]). Also, $\mu_{s}(\mathcal{A})<1$ for each $s$ because $\mathcal{A}$ has empty interior.

Let $L(t)=\mu_{t}(\mathcal{A})$. We note that $L$ is a computable martingale in the sense of algorithmic randomness. That is, $L \geq 0$, the "martingale equality" $L(s 0)+$ $L(s 1)=2 L(s)$ holds for each string $s$, and $L(s)$ is a computable real uniformly in $s$; see e.g. [9, Chap. 7]. By hypothesis that $\mathcal{A}$ has empty interior, we have $L(s)<1$ for each $s$.

We will show that the oscillation behaviour of $L$ can be controlled sufficiently well in order to code the $\Pi_{3}^{0}$-complete set $\mathcal{E}$ from Proposition 3 into $D \mathcal{A}$. For $p \in \mathbb{N}$ let $\delta_{p}=1-3^{-p}$. We write $\theta(s)=p$ if $\delta_{p-1}<L(s)<\delta_{p}$. We leave $\theta(s)$ undefined in case $L(s)=\delta_{k}$ for some $k$. In the following, when we write $\theta(s)$ we imply that $\theta(s)$ is defined. Observe that, if $\theta\left(X \upharpoonright_{n}\right)$ is defined for each $n$ and $\lim _{n} \theta(X \upharpoonright n)=\infty$, then $X \in D(+A$. Note that the binary relations $\{\langle s, p\rangle: \theta(s)=p\}$ and $\{\langle s, p\rangle: \theta(s) \geq p\}$ are $\Sigma_{1}^{0}$ because the martingale $L$ is computable.

## Lemma 5.

(i) (Increasing the value of $L$ ). Let $p<k$ and $\theta(s)=p$. There is $t \supset s$ such that $\theta(t) \geq k$ and $L(u)>\delta_{p-1}$ for each $u$ with $s \subseteq u \subseteq t$.
(ii) (Decreasing the value of $L$ ). Let $p>q$ and $\theta(s)=p$. There is $t \supset s$ such that $\theta(t)=q$ and $L(u)>\delta_{q-1}$ for each $u$ with $s \subseteq u \subseteq t$.

Proof. (i) By an application of the Lebesgue density theorem [1, Proposition 3.5] (where $r$ there is $\delta_{k}$ ), there exists a prefix minimal string $v \supset s$ such that $L(v) \geq \delta_{k}$ and $L(u) \geq L(s)$ for each $u$ with $s \subseteq u \subseteq v$. If $\theta(v)$ is defined, the string $t=v$ is as required. Otherwise we have $L(v)=\delta_{m}$ for some $m \geq k$. By the Lebesgue density theorem, $L$ is not constant on the set of extensions of $v$, so one can choose a prefix minimal string $w \supseteq v$ such that $L(w 0) \neq L(w 1)$. By the minimality of $w$ we have $L\left(v^{\prime}\right)=\delta_{m}$ for each $v^{\prime}$ with $v \subseteq v^{\prime} \subseteq w$.

First suppose that $L(w 0)>L(w)$. Since $L(w 0)<1$, we have $L(w 0)-L(w)<$ $3^{-m}$. Hence $L(w)-L(w 1)=L(w 0)-L(w)<3^{-m}$ by the martingale equality, and thus $\delta_{m-1}<L(w 1)<\delta_{m}$ because $\delta_{m-1}=\delta_{m}-2 \cdot 3^{-m}$. Then $\theta(w 1)=m \geq k$, so the string $t=w 1$ is as required. If $L(w 1)>L(w)$ instead, then $t=w 0$ is as required by an analogous argument.
(ii) As $L(s)<1$, by the Lebesgue density theorem there exists a prefix minimal $t \supset s$ such that $L(t)<\delta_{q}$. Let $t=v b$ where $b \in\{0,1\}$. Since $L \leq 1$ the martingale equality for $1-L$ implies that $2(1-L(v)) \geq 1-L(t)$. So $L(t) \leq \delta_{q-1}$ would imply $2(1-L(v)) \geq 1-L(t) \geq 3 \cdot 3^{-q}>2 \cdot 3^{-q}$, and hence $L(v)<\delta_{q}$ contrary to the minimality of $t$. Hence $\theta(t)=q$ and $L(u)>\delta_{q-1}$ for each $u$ with $s \subseteq u \subseteq t$, as required. This establishes the lemma.

Remark 6. By the remarks on $\theta$ above and since $L$ is computable, all the conditions in the Lemma are $\Sigma_{1}^{0}$. Thus in (i) given a string $s$ and numbers $p<k$, if $\theta(s)=p$ we can run a search for the string $t$. So the function $s, p, k \mapsto t$ is partial computable and defined whenever $\theta(s)=p<k$. A similar remark applies to (ii).

Proof (of Theorem 4). By Proposition 3, the set

$$
\mathcal{E}=\left\{Z \in^{\omega \times \omega} 2 \mid \forall q \forall^{\infty} n[Z(q, n)=0]\right\}
$$

is $\Pi_{3}^{0}$-complete; thus, it will be enough to show that $\mathcal{E} \leq_{m} D(\mathcal{A})$. In the following let $a, b$ denote bit-valued square matrices. Given such a matrix $a$, we will compute a string $s=\psi(a) \in{ }^{<\omega_{2}}$ in such a way that $\theta(s)$ is defined. We ensure that $a \subseteq b \rightarrow \psi(a) \subseteq \psi(b)$. Then the function

$$
\Psi:{ }^{\omega \times \omega} 2 \rightarrow{ }^{\omega} 2, \quad \Psi(Z)=\bigcup_{n} \psi(Z \upharpoonright n \times n)
$$

is a total Turing functional.
Defining $\psi$. The definition of $\psi$ is by recursion, using Lemma 5 and Remark 6. If $\mu \mathcal{A}>1 / 2$, after removing from $\mathcal{A}$ either the elements of Cantor space starting with 0 or the elements starting with 1 , we may assume that $\mu \mathcal{A} \leq 1 / 2$. So we may assume that $\theta(\emptyset)$ is defined. We set $\psi(\emptyset)=\emptyset$.

Now suppose that $\phi(a)$ has been defined for each $n \times n$ bit-valued matrix $a$ in accordance with the conditions above. Given a matrix $b=\langle a(i, j) \mid i, j<n+1\rangle$, let $s=\psi(a)$ where $a=b \upharpoonright n \times n$. Then $p=\theta(s)$ is defined.

If there is a $q \leq n$ such that $b(q, n)=1$, choose $q$ least.

- If $q<p$, via (ii) of Lemma 5 compute a string $t \supset s$ such that $\theta(t)=q$ and $L(u)>\delta_{q-1}$ for each $u$ with $s \subseteq u \subseteq t$, and define $\psi(b)=t$.
- If $q \geq p$, or there is no such $q \leq n$ at all, let $k=\max (p+1, n+1)$. Via (i) of Lemma 5 compute a string $t \supset s$ such that $\theta(t) \geq k$ and $L(u)>\delta_{p-1}$ for each $u$ with $s \subseteq u \subseteq t$, and define $\psi(b)=t$. This completes the recursion step.

We verify that $\mathcal{E} \leq_{m} D \mathcal{A}$ via $\Psi$. First suppose that $Z \notin \mathcal{E}$. Let $q$ be least such that $\exists^{\infty} n[Z(q, n)=1]$. Then for arbitrarily large $n$, we have $\theta(\psi(Z \upharpoonright n \times n))=q$, hence $\theta(\Psi(Z) \upharpoonright r)=q$ for infinitely many $r$. Therefore $\Psi(Z) \notin D \mathcal{A}$.

Now suppose that $Z \in \mathcal{E}$. Then for every $q \in \omega$, there is $m_{q}$ such that $\forall m \geq m_{q}[Z(q, m)=0]$. For $r \in \omega$, let $n_{r}=\max \left\{m_{0}, \ldots, m_{r}, r\right\}$. Note that for each $n \geq n_{r}$, if $Z(q, n)=1$ then $q>r$.

The following claim will show that $\Psi(Z) \in D \mathcal{A}$.
Claim. Given $n>n_{r}$, let $s=\psi(Z \upharpoonright n \times n)$ and $t=\psi(Z \upharpoonright(n+1) \times(n+1))$.
For each string $u$ with $s \subseteq u \subseteq t$, we have $L(u) \geq \delta_{r}$.
To see this, we prove inductively that $\theta(\psi(Z \upharpoonright n \times n))>r$ for each $n>n_{r}$. For the start of the induction at $n_{r}+1$, suppose that $s=\psi\left(Z \upharpoonright n_{r} \times n_{r}\right)$ has just been defined and consider the next step of the definition of $\psi$ along $Z$. The least possible value of $q$ is $r+1$. Since $n_{r} \geq r$, no matter whether we apply (ii) or (i) of the lemma we ensure that $\theta(t)>r$ and hence $L(t)>\delta_{r}$, where $t=\psi\left(Z \upharpoonright\left(n_{r}+1\right) \times\left(n_{r}+1\right)\right)$.

For the inductive step, suppose that $n>n_{r}$ and for $s=\psi(Z \upharpoonright n \times n)$ we have $\theta(s)>r$. Let $t=\psi(Z \upharpoonright(n+1) \times(n+1))$. Again the least possible value of $q$ is $r+1$. If we apply (ii) of the lemma then $\theta(t)>r$ and $L(u)>\delta_{r}$ for each $u$ with $s \subseteq u \subseteq t$. If we apply (i) then, where $\theta(s)=p>r$, we have $\theta(t) \geq p+1$ and $L(u)>\delta_{r}$ for each $u$ with $s \subseteq u \subseteq t$. This completes the claim.

We conclude that for each $r$, we have $L\left(\Psi(Z) \upharpoonright_{m}\right) \geq \delta_{r}$ for sufficiently large $m$. Hence $\Psi(Z) \in D \mathcal{A}$.

Theorem 4 leaves open the following question. Is there a $\Pi_{1}^{0}$ class with empty interior and non-computable measure such that the density set is not $\Pi_{3}^{0}$-complete?

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# Program Size Complexity of Correction Grammars in the Ershov Hierarchy 

John Case ${ }^{1}$ and James S. Royer ${ }^{2(\boxtimes)}$<br>${ }^{1}$ Department of Computer and Information Sciences, University of Delaware, Newark, DE 19716-2586, USA case@udel.edu<br>${ }^{2}$ Department of Electrical Engineering and Computer Science, Syracuse University, Syracuse, NY 13244, USA<br>jsroyer@syr.edu


#### Abstract

A general correction grammar for a language $L$ is a program $g$ that, for each $(x, t) \in \mathbb{N}^{2}$, issues a yes or no (where when $t=0$, the answer is always no) which is $g$ 's $t$-th approximation in answering " $x \in L$ ?"; moreover, $g$ 's sequence of approximations for a given $x$ is required to converge after finitely many mind-changes. The set of correction grammars can be transfinitely stratified based on $O$, Kleene's system of notation for constructive ordinals. For $u \in O$, a $u$-correction grammar's mind changes have to fit a count-down process from ordinal notation $u$; these $u$-correction grammars capture precisely the $\Sigma_{u}^{-1}$ sets in Ershov's hierarchy of sets for $\Delta_{2}^{0}$. Herein we study the relative succinctness between these classes of correction grammars. Example: Given $u$ and $v$, transfinite elements of $O$ with $u<_{o} v$ (Kleene's ordering on $O$ ), for each $\emptyset^{(2)}$-computable $H: \mathbb{N} \rightarrow \mathbb{N}$, there is a $v$-correction grammar $i_{v}$ for a finite (alternatively, a co-finite) set $A$ such that the smallest $u$-correction grammar for $A$ is $>H\left(i_{v}\right)$. We also exhibit relative succinctness progressions in these systems of grammars and study the "information-theoretic" underpinnings of relative succinctness. Along the way, we verify and improve slightly a 1972 conjecture of Meyer and Bagchi.


## 1 Correction Grammars and Relative Succinctness

Burgin [5] suggested that a person's knowledge of a language $L$ may involve his/her storing a representation of $L$ in terms of two grammars, say $g_{1}$ and $g_{2}$, where $g_{1}$ 's language over-generalizes $L$ and $g_{2}$ is used to "edit" errors of (make corrections to) $g_{1}$. Thus, $L=\left(L_{1}-L_{2}\right)=\left\{x \in L_{1} ; x \notin L_{2}\right\}$, where $L_{i}$ is the language generated by the grammar $g_{i}$. The pair $\left\langle g_{1}, g_{2}\right\rangle$ can thus be seen as a single description of (or "grammar" for) the language $L$. Burgin called such pairs grammars with prohibition; we prefer the term correction grammars. These correction grammars may be seen as modeling self-correcting human behavior. ${ }^{1}$

[^39]In the computability-theoretic context of the present paper, we first think of $p$ as being a correction grammar for $L$ if and only if $p=\langle i, j\rangle$ and $L=\left(W_{i}-W_{j}\right) .{ }^{2}$ Hence, such a correction grammar $\langle i, j\rangle$ is an index from [18] for the difference of c.e. sets, i.e., for the d.c.e. set, $\left(W_{i}-W_{j}\right)$. Correction grammars are more powerful than ordinary c.e. grammars in the sense that there are languages (e.g., $L=\left\{x ; x \notin W_{x}\right\}$ ) that correction grammars can describe that ordinary c.e. grammars cannot.

This paper's focus is on a different advantage correction grammars have over c.e. grammars: relative succinctness. That is, correction grammars can provide vastly shorter descriptions of certain c.e. languages (in our cases, finite and cofinite languages) than can any c.e. grammar. Here is a sample result.

Theorem 1. Suppose $H: \mathbb{N} \rightarrow \mathbb{N}$ is $\emptyset^{(2)}$-computable and $W_{i_{1}}=\mathbb{N}$. Then there is a c.e. grammar $i_{2}$ with $W_{i_{2}}$ co-finite (hence, $\left(W_{i_{1}}-W_{i_{2}}\right)$ is finite, and, thus, c.e.) such that $\min \left\{i: W_{i}=\left(W_{i_{1}}-W_{i_{2}}\right)\right\}>H\left(\left\langle i_{1}, i_{2}\right\rangle\right)$; that is, when one magnifies the numerical value of the correction grammar $\left\langle i_{1}, i_{2}\right\rangle$ by $H$, the result is still smaller than the numerical value of any c.e. grammar for the finite language named by $\left\langle i_{1}, i_{2}\right\rangle$.

## Scholium 2 (On Theorem 1).

(a) It is especially interesting to consider the cases where Theorem 1's function $H$ is large valued and fast growing; then the value of $H\left(\left\langle i_{1}, i_{2}\right\rangle\right)$ is a vast magnification of the value of $\left\langle i_{1}, i_{2}\right\rangle$. Hence, thanks to the theorem, the minimum c.e. grammar for $\left(W_{i_{1}}-W_{i_{2}}\right)$ is huge compared its correction grammar $\left\langle i_{1}, i_{2}\right\rangle$.
(b) For the magnification function $H$, one can concretely take $H=$ $\lambda x \cdot 10^{20^{30^{x}}}$ or $H=\lambda x \cdot \operatorname{Ack}(x+100, x+100)$ where Ack is Ackermann's function [20]. However, these are tame choices as there are $\emptyset^{(2)}$-computable functions that dominate every $\emptyset^{(1)}$-computable function. ( $f$ dominates $g \Leftrightarrow_{\text {def }}$ $\{x: g(x)<f(x)\}$ is co-finite.)
(c) In effect, the theorem identifies the size of the c.e. grammar $i$ with the number $i$ itself and the size of the correction grammar $\left\langle i_{1}, i_{2}\right\rangle$ with the number $\left\langle i_{1}, i_{2}\right\rangle$. As explained in Sect. 2, a consequence of Theorem 1 is that the analogous result holds for a wide choice of size-measurement schemes for grammars.
(d) Theorem 1's $\left\langle i_{1}, i_{2}\right\rangle$ can be computed from a relativized program for $H$. Most of our relative succinctness results are uniformly algorithmic in this sense.
(e) For simplicity, Theorem 1 asserts that there is a grammar (e.g., $\left\langle i_{1}, i_{2}\right\rangle$ ) witnessing the relative succinctness. In fact we can construct an infinite c.e. set of such witnesses for Theorem 1 (and analogously for Corollary 4 and Theorems 6 and 7 below).

[^40](f) Theorem 1 turns out to fail if one attempts to strengthen it by allowing $H$ to be an arbitrary $\emptyset^{(3)}$-computable function.

## 2 A Digression on Grammar Size Measures

Before discussing extensions of Theorem 1 we first consider how versions of these theorems still hold under a wide range of size-measurement schemes.

Definition 3. For a given oracle $A$, an $A$-computable Blum program size measure (or simply, $A$-computable size measure) is a finite-to-one, $A$-computable $s: \mathbb{N} \rightarrow \mathbb{N}$ such that $b_{s}=\lambda n . \max \{i: s(i) \leq n\}$ is also $A$-computable.

The $\emptyset$-computable size measures match Blum's original notion of program size measure [2]. These include $s(i)=i, s(i)=|i|=$ the length of the reduced binary representation of $i$ (i.e., no redundant, leading 0 s ), and most standard ways of measuring program (or grammar) size. For some less-standard size measurements, we need a bit of terminology: For each $A \subseteq \mathbb{N}$, let $U^{A}$ be an additively optimal universal partial $A$-computable function [22]. (Such an $U^{A}$ exists by a straightforward relativization of Lemma 2.1.1 in [22].) Define $C^{A}=\lambda x \cdot \min \left\{|p| ; U^{A}(p)=x\right\} . C^{A}(x)$ is the $A$-Kolmogorov complexity of $x$. Now, for each $A, C^{A}$ is an $A^{(1)}$-computable size measure. E.g., $C^{\emptyset}$, the standard notion of (plain) Kolmogorov complexity, is an $\emptyset^{(1)}$-computable size measure.

It turns out that all of our $\emptyset^{(k)}$-relative-succinctness results in this paper (e.g., Theorem 1 for $k=2$ ) still hold under any choice of $\emptyset^{(k)}$-computable size measures. So for example, Theorem 1 generalizes to:

Corollary 4. Suppose $H: \mathbb{N} \rightarrow \mathbb{N}$ is $\emptyset^{(2)}$-computable, $s_{c e}$ and $s_{d c e}$ are any $\emptyset^{(2)}$ _ computable size measures, and $W_{i_{1}}=\mathbb{N}$. Then there is a c.e. grammar $i_{2}$ with $W_{i_{2}}$ co-finite and $\min \left\{s_{c e}(i) ; W_{i}=\left(W_{i_{1}}-W_{i_{2}}\right)\right\}>H\left(s_{d c e}\left(\left\langle i_{1}, i_{2}\right\rangle\right)\right)$.

Scholium 5 (On Corollary 4). Taking $s_{c e}=s_{d c e}=C^{\emptyset^{(1)}}$ and $H=$ a large, fast-growing $\emptyset^{(2)}$-computable function in Corollary 4 provides some insight into the trade-offs behind relative succinctness. As measured by $C^{\emptyset^{(1)}}, \min \left\{C^{\emptyset(1)}(i)\right.$ : $\left.W_{i}=\left(W_{i_{1}}-W_{i_{2}}\right)\right\}$ is the least amount of $\emptyset^{(1)}$-algorithmic information in any c.e. grammar for ( $W_{i_{1}}-W_{i_{2}}$ ) which, by Corollary 4 , is $H$-enormous compared to $s_{d c e}\left(\left\langle i_{1}, i_{2}\right\rangle\right)$.
(a) Thus, intuitively, a necessary price a then witnessing correction grammar $\left\langle i_{1}, i_{2}\right\rangle$ pays for its small size is that $\left\langle i_{1}, i_{2}\right\rangle$ is enormously deficient in $C^{\emptyset^{(1)}}$ information compared to that of any c.e. grammar for $\left(W_{i_{1}}-W_{i_{2}}\right)$.
(b) Such a witness $\left\langle i_{1}, i_{2}\right\rangle$ would have its own information minimality if $C^{\emptyset^{(1)}}\left(\left\langle i_{1}, i_{2}\right\rangle\right)$ just happened to $=\min \left\{C^{\emptyset^{(1)}}\left(\left\langle j_{1}, j_{2}\right\rangle\right):\left(W_{j_{1}}-W_{j_{2}}\right)=\left(W_{i_{1}}-\right.\right.$ $\left.\left.W_{i_{2}}\right)\right\}$. However, as mentioned in Scholium 2(e), we can produce an infinite c.e. set, $A$, of witnesses, $\left\langle i_{1}, i_{2}\right\rangle$, to Corollary 4, and, by straightforward relativization of well-known immunity properties of sets of minimal indices [25], all but finitely many elements of this $A$ are not such $C^{\emptyset^{(1)}}$-minimal size d.c.e. grammars.

Section 5 below further discusses some information-theoretic aspects of relative succinctness.

## 3 The Ershov Hierarchy and Relative Succinctness

Versions of Theorem 1 hold for generalizations of the notion of correction grammar. One natural extension is to consider finite differences of c.e. languages. That is, for $k>0$, define a $k$-fold correction grammar as being a number of the form $\left\langle i_{1}, \ldots, i_{k}\right\rangle$ that names the language ( $W_{i_{1}}-\left(W_{i_{2}}-\cdots\left(W_{i_{k-1}}-W_{i_{k}}\right) \cdots\right)$ ). The languages named by $k$-fold correction grammars are called the $k$-c.e. sets [36]; they formalize the notion of a finite, fixed number of successive edits for errors. The Ershov hierarchy [12-14] includes and extends this basic idea into the transfinite by means of constructive ordinals as discussed below.

Ordinals are representations of well-orderings. A constructive ordinal is, roughly, an ordinal, $\alpha$, for which there is a program, called a notation, that specifies how to build $\alpha$ algorithmically or lay $\alpha$ out end-to-end. Herein we use Kleene's well-known general ordinal notation system $O$ (coded as a proper subset of $\mathbb{N})[21,30]$. Each constructive ordinal has at least one $O$-notation. For each $k \in \mathbb{N}$, let $\underline{k}=$ the unique $O$-notation for the ordinal $k$. Kleene also introduced an order relation $<_{o}$ on $O$-notations that naturally embeds into the ordering of the corresponding constructive ordinals [21,30], e.g., $\underline{0}<_{o} \underline{1}<_{o} \underline{2}<_{o} \ldots<_{o} w$, where $w$ is any $O$-notation for $\omega$ (the ordinal for $\mathbb{N}$ under the ordering: $0<1<2<\ldots$ ).

Our most general formulation of correction grammars is based on the idea of using Kleene's $O$-notations to bound the number of corrections that such a grammar can make. To do this rigorously, we make use of the technique of algorithmic counting-down from $O$-notations. For a notation $u \in O$, our $u$ correction grammars are each algorithms for counting-down corrections from $u$.

For $u \in O$, here is roughly how level- $u$ correction grammars work. Choose uniform programs $c, d$ where $c$ is for counting-down from $u$ (along $<_{o}$ ), and where $d$, for each time $t$ and input $x$, decides whether to exclude or include $x$ at time $t$-but, initially, i.e., at $t=0, d$ excludes (independently of $x$ ). With each mind-change of $d$-about excluding or including a given $x, c$ must walk/leap strictly further down $u$ (along $<_{o}$ ). The correction grammar $p=\langle c, d\rangle$ accepts the set, $W_{p}^{u}$, of all $x$ that $d$ eventually includes-once $c$ counts down from $u$ no more. By the well-ordering property of ordinals, such a count-down sequence must be finite and, hence, such a grammar makes only finitely many corrections on excluding/including any given $x$.

Examples: For $k \in \mathbb{N}$, counting-down corrections from $\underline{k}$ is equivalent to an algorithm that initially excludes each item $x$ and, then, the algorithm can change its mind about $x$ 's inclusion or exclusion up to $k$ times on the way to giving its final, correct answer as to whether $x$ is included or excluded. When $k=1$, such algorithms are grammars for the c.e. sets; when $k=2$, such algorithms are grammars for the d.c.e. sets; and in general, for a fixed $k$, they provide
grammars for the $k$-c.e. sets. ${ }^{3}$ It can be shown that counting-down corrections allowed from any notation for $\omega$ is equivalent to declaring algorithmically, at the time a first correction is made, a finite numerical bound on the number of further corrections to be allowed. This is more powerful than just initially setting a fixed, finite number of corrections allowed. As another example consider using a notation for the ordinal $\omega+\omega$ (two copies of $\omega$ laid end to end), also constructive and transfinite: in this case, when the first correction is made, the algorithm declares a finite bound on the number of further corrections it will make; this bound is, however, allowed to be changed once, at a later time. For a notation for the constructive ordinal $\omega+\omega+\omega$, the algorithm is allowed to update the bound twice. For at least natural notations for the constructive ordinal $\omega^{2}$, the algorithm is allowed to make a finite number of changes to the bound, where the maximum number of changes allowed to the bound is announced when the algorithm makes the first correction!

Each $W_{p}^{u}$ as sketched above is manifestly a $u$-c.e. set from the $\Sigma_{u}^{-1}$ level in the general Ershov Hierarchy [12-14]. Hence, $u$-correction grammars provide a motivation for studying the $u$-c.e. sets. ${ }^{4}$ It turns out that each $u$-c.e. set has a correction grammar as above, and, moreover, that $\left\langle W_{p}^{u}\right\rangle_{p \in \mathbb{N}}$ is an acceptable indexing of $\Sigma_{u}^{-1} .{ }^{5}$ The details of the $W_{p}^{u}$ 's definition will appear in the full paper.

Here is our lift of Theorem 1 to our general formulation of correction grammars.

Theorem 6. Suppose $\underline{0}<_{o} u<_{o} v \in O, H: \mathbb{N} \rightarrow \mathbb{N}$ is $\emptyset^{(2)}$-computable, and $s_{u}$ and $s_{v}$ are $\emptyset^{(2)}$-computable size measures. Then (a) and (b) just below each hold.
(a) If $v$ is transfinite or else finite and even, then there is an $e$ with $W_{e}^{v}$ finite such that $\min \left\{s_{u}(p): W_{p}^{u}=W_{e}^{v}\right.$ or $\left.W_{p}^{u}=\overline{W_{e}^{v}}\right\}>H\left(s_{v}(e)\right)$.
(b) If $v$ is transfinite or else finite and odd, then there is an e with $W_{e}^{v}$ co-finite such that $\min \left\{s_{v}(p) ; W_{p}^{u}=W_{e}^{v}\right.$ or $\left.W_{p}^{u}=\overline{W_{e}^{v}}\right\}>H\left(s_{v}(e)\right)$.

It turns out that if $v$ is finite and odd, then the conclusion of Theorem 6(a) is false; and similarly with Theorem 6(b) for $v$ finite and even. We note that these odd/even provisos go away if the theorem's $H$ is restricted to be $\emptyset^{(1)}$-computable.

Our proof of Theorem 6 is a succinct and indirect combination of ErshovHierarchy refinements of index-set degree-calculations of Rogers [30] together with recursion theorems. The next section has a sample proof in this style, but where the refinement of Rogers' index-set degree-calculations is only a relativization.
${ }^{3}$ For each $k \in \mathbb{N}$, the system of $\underline{k}$-correction grammars turns out to be equivalent to the system of $k$-fold correction grammars introduced above.
${ }^{4}$ Algorithmic counting-down from constructive ordinals has been used, for example, in: computability theory $[1,7,11]$, proof theory $[29,38]$, term rewriting [ 4,40$]$, and computational learning theory [15].
${ }^{5}\left\langle W_{p}^{u}\right\rangle_{p \in \mathbb{N}}$ 's acceptability provides s-m-n and recursion theorems that are needed in some of our proofs.

Let $w$ be an $O$-notation for $\omega$. While our $w$-c.e. sets are precisely $\Sigma_{w}^{-1}$, they are distinct from the well known omega-c.e. sets [37] (each of which has a computable approximation analogous to an $w$-c.e. set except the number of mind changes is bounded by some computable function). The omega-c.e. sets are in $\Delta_{w}^{-1} \subsetneq \Sigma_{w}^{-1}$.

## 4 Confirmation of the Meyer and Bagchi Conjecture

In 1972 Meyer and Bagchi [25] conjectured a (non-Ershov Hierarchy) succinctness result that, by adapting Theorem 6's proof, we can confirm (and improve ${ }^{6}$ ) in:

Theorem 7. Let $A$ and $B$ be sets with $A^{(1)} \leq_{\mathrm{T}} B$ and let $s_{A}$ and $s_{B}$ be $B^{(2)}{ }_{-}$ computable size measures. Then, given any $B^{(2)}$-computable function $H$, one can compute an $e\left(\right.$ from a $\varphi^{B^{(2)}}$-program for $H$ ) such that $W_{e}^{B}$ is co-finite and $\min \left\{s_{A}(p) ; W_{p}^{A}=W_{e}^{B}\right\}>H\left(s_{B}(e)\right)$.

Proof. ${ }^{7}$ Since $A^{(1)} \leq_{\mathrm{T}} B$, we have $\Sigma_{1}^{0, A} \cup \Pi_{1}^{0, A} \subseteq \Delta_{1}^{0, B}$. By standard results, $\left\{\langle p, q\rangle: W_{p}^{A} \neq W_{q}^{B}\right\} \in \Sigma_{2}^{0, B}$. Theorem 14-XV's proof in [30] relativizes to give: $\left(B^{(3)}, \overline{B^{(3)}}\right) \leq_{m}\left(\left\{i ; W_{i}^{B}\right.\right.$ is co-finite $\left.\},\left\{i ; W_{i}^{B} \notin \Delta_{1}^{0, B}\right\}\right)$. Let $r$ witness this reduction. As $s_{A}$ is a $B^{(2)}$-computable size measure, by Definition 3, $b_{s_{A}}=$ $\lambda n$. $\max \left\{i: s_{A}(i) \leq n\right\}$ is also $B^{(2)}$-computable. By the recursion theorem for $\varphi^{B^{(2)}}$, given a $\varphi^{B^{(2)}}$-program for $H$, we can compute an $i_{0}$ such that

$$
\varphi_{i_{0}}^{B^{(2)}}=\lambda x \cdot \begin{cases}0, & \text { if }\left(\forall p \leq b_{s_{A}}\left(H\left(s_{B}\left(r\left(i_{0}\right)\right)\right)\right)\right)\left[W_{p}^{A} \neq W_{r\left(i_{0}\right)}^{B}\right] \\ \uparrow, & \text { otherwise }\end{cases}
$$

Let $e=r\left(i_{0}\right)$. Case 1: $\left(\exists p \leq b_{s_{A}}\left(H\left(s_{B}(e)\right)\right)\right)\left[W_{p}^{A}=\overline{W_{e}^{B}}\right]$. Then $\varphi_{i_{0}}^{B^{(2)}}\left(i_{0}\right) \uparrow$. Thus, $W_{e}^{B} \notin \Delta_{1}^{0, B}$, and so, $W_{e}^{B} \notin \Pi_{1}^{0, A}$ contradicting Case 1. Hence, we must have Case 2: $\left(\forall p \leq b_{s_{A}}\left(H\left(s_{B}(e)\right)\right)\right)\left[W_{p}^{A} \neq \overline{W_{e}^{B}}\right]$. Then $\varphi_{i_{0}}^{\emptyset^{(2)}}\left(i_{0}\right) \downarrow$. Hence, $W_{e}^{B}\left(=W_{r\left(i_{0}\right)}^{B}\right)$ is co-finite. Also, since $\max \left\{p: s_{A}(p) \leq H\left(s_{B}(e)\right)\right\}=$ $b_{s_{A}}\left(H\left(s_{B}(e)\right)\right)$, we also have $\left(\forall p: s_{A}(p) \leq H\left(s_{B}(e)\right)\right)\left[W_{p}^{A} \neq \overline{W_{e}^{B}}\right]$.

Theorem 7 can be shown by a direct (albeit involved) construction of a suitable $W_{e}^{B}$ thus. By standard results, write $H$ as the double limit of a $B$-computable function (with the inner limit also total). Double limits tend to thrash. To handle thrashing in $W_{e}^{B}$ 's construction, while crucially ensuring that $\overline{W_{e}^{B}}$ is finite, one can use a variant of Sacks' restraint trick [32] that allows a new number to be held out of $W_{e}^{B}$ only when no higher priority concern holds a number out of $W_{e}^{B}$.

[^41]
## 5 Why Are the Programs Succinct?

The succinct programs above are so, in part, because they are missing information. Scholium 5 above noted a link between relative succinctness and "information deficiency." Below is a sample of our results that explore this link. For this sample the notion of information deficiency is based on unprovability. Let $\mathcal{T}$ be a $\emptyset^{(2)}$-computably axiomatized extension of Second Order Arithmetic (SOA) $[30,34]$ such that $\mathcal{T}$ 's theorems expressible in SOA are true. E.g., $\mathcal{T}=$ ZFC + all $\Delta_{3}^{0}$-truths of first-order Arithmetic. $\mathcal{T}$ 's being $\emptyset^{(2)}$-computably axiomatized makes its set of theorems a $\Sigma_{3}^{0}$-set by direct relativization of standard results $[9,23,30] . \mathcal{T}$ can thus be regarded as a $\emptyset^{(2)}$-algorithmic extractor of SOA truth.

Theorem 8. Suppose (i) through (iv) just below.
(i) $\mathcal{T}$ is a $\emptyset^{(2)}$-computability axiomatized extension of SOA such that $\mathcal{T}$ 's theorems expressible in SOA are true.
(ii) $\underline{0}<_{o} u<_{o} v \in O ; \mathcal{A} \subseteq \Sigma_{u}^{-1} ; s_{u}$ and $s_{v}$ are $\emptyset^{(2)}$-computable size measures.
(iii) For each informal sentence $E$ (translatable to $S O A$ ), we write $\langle\langle E\rangle$ for a naturally corresponding, fixed closed well-formed formula of SOA that semantically expresses $E$.
(iv) For each $\emptyset^{(2)}$-computable $H: \mathbb{N} \rightarrow \mathbb{N}$, there is an $e_{H}$ with $W_{e_{H}}^{v} \in \mathcal{A}$ and $\min \left\{s_{u}(p): W_{p}^{u}=W_{e_{H}}^{v}\right\}>H\left(s_{v}\left(e_{H}\right)\right)$.

Then, for each sufficiently large, $\emptyset^{(2)}$-computable $H: \mathbb{N} \rightarrow \mathbb{N}$ and for each $e_{H}$ as in (iv) just above, we have $\mathcal{T} \nvdash\left\langle\left\langle u\right.\right.$ and $v \in O$ and $\left.\left.W_{e_{H}}^{v} \in \Sigma_{u}^{-1}\right\rangle\right\rangle$.

Theorem 8's conclusion just above provides an information deficiency of $e_{H}$, e.g., in many obvious cases where SOA $\vdash\langle\langle u$ and $v \in O\rangle\rangle$.

The base of $\mathcal{T}$ above can be first-order arithmetic [23,30]-provided that, roughly, we change $\langle\langle u$ and $v \in O\rangle\rangle$ above to $\langle\langle u$ and $v$ settled $\rangle\rangle$, where settled is an approximation to $O$ meaning that the limit, tied to our uniform count-down decision process, converges. ${ }^{8}$

## 6 Relative Succinctness Progressions

Fix a $\emptyset^{(2)}$-computable $H: \mathbb{N} \rightarrow \mathbb{N}$ and $\underline{0}<_{o} u_{0}<_{o} u_{1}<_{o} \ldots<_{o} u_{n} \in O$. By Theorem 6 above, we know that, for each of $j=1, \ldots, n$, there is an $e_{j}$ with $W_{e_{j}}^{u_{j}}$ finite or co-finite (subject to the even/odd conditions on the finite $u_{j}$ 's) with $e_{j}$ $H$-more succinct than any $W^{u_{j-1}}$-program for $W_{e_{j}}^{u_{j}}$. It turns out that we can arrange for each of these $W_{e_{j}}^{u_{j}}$ 's to be the same set or its complement. That is:

[^42]Theorem 9 (A sample succinctness progression). Suppose $H$ is $\emptyset^{(2)}$-computable and $\underline{0}<_{o} u_{0}<_{o} u_{1}<_{o} \ldots<_{o} u_{n} \in O$. Let $s_{u_{0}}, \ldots, s_{u_{n}}$ be $\emptyset^{(2)}$-computable size measures. For $j=1, \ldots, n$, let $\mathrm{op}_{j}$ be an element of $\{\lambda A \cdot A, \lambda A \cdot \bar{A}\}$ that is chosen arbitrarily subject to two constraints: (i) $\mathrm{op}_{j}=\lambda A$. $A$ when $u_{j}$ is finite and even, and (ii) $\mathrm{op}_{j}=\lambda A \cdot \bar{A}$ when $u_{j}$ is finite and odd. Then there are $e_{1}, \ldots, e_{n}$ and $A$, a finite set, such that (a) $\mathrm{op}_{1}\left(W_{e_{1}}^{u_{1}}\right)=\cdots=\mathrm{op}_{n}\left(W_{e_{n}}^{u_{n}}\right)=A$, and (b) for $j=1, \ldots, n: \min \left\{s_{u_{j-1}}(p): W_{p}^{u_{j-1}}=W_{e_{j}}^{u_{j}}\right.$ or $\left.W_{p}^{u_{j-1}}=\overline{W_{e_{j}}^{u_{j}}}\right\}>$ $H\left(s_{u_{j}}\left(e_{j}\right)\right)$.

It is notable that a single finite set can witness the succinctness between each of the adjacent pairs in $\left\langle W_{i}^{u_{0}}\right\rangle_{i \in \mathbb{N}}, \ldots,\left\langle W_{i}^{u_{n}}\right\rangle_{i \in \mathbb{N}}$ : conceivably, the sets witnessing the relative succinctness between distinct pairs must be very different-perhaps because the basis for relative succinctness between adjacent pairs are "orthogonal." Theorem 9 and related (but here not included) succinctness progression results show that this is not the case. We note that, as with Theorem 6, the theorem's odd/even provisos go away if the theorem's $H$ is restricted to be $\emptyset^{(1)}$ computable. Theorem 9 is proved by a parameterized extension of the proof of Theorem 6.

## 7 Related Results

Blum in his seminal paper [2] provided a computability-theoretic, axiomatic treatment of program size and established the remarkable result that: for any computable function $h$, for some primitive recursive functions $f$, there are general recursive procedures for $f$ which are $h$ more succinct than any primitive recursive procedures for $f$. Hence, although one need not go beyond primitive recursions to compute such $f$, judicious use of more general recursion leads to programs for $f$ which are considerably, i.e., $h$, smaller than any of those employing primitive recursions only! Drumm [10] and Constable [8] independently refined Blum's techniques to exhibit relative-succinctness results analogous to Blum's, but between successive levels of the LOOP-hierarchy [27].

The above results left open the possibility that the functions admitting more succinct programs might be pathological examples that no one would ever want to compute anyway. However, Meyer and Fischer [26] showed that push-down automata (PDAs) are computably more succinct than finite state automata for some co-finite sets, and Meyer [24] provided very general techniques for obtaining relative succinctness results for characteristic functions of finite sets! In particular, Meyer showed that double recursive procedures [28] are $\emptyset^{(1)}$-computably more succinct than primitive recursive procedures for some characteristic functions of finite sets. He was the first to notice that, in many cases, one can obtain programs that are more succinct by a $\emptyset^{(1)}$-computable amount.

Computable relative-succinctness results were shown by: Borodin [3] for context-sensitive grammars over context-free grammars (CFGs); Valiant [39] for ambiguous CFGs over deterministic CFGs; and Schmidt and Szymanski [33] for ambiguous CFGs over unambiguous CFGs. Hartmanis and Baker showed
that, if $\mathrm{P} \neq \mathrm{NP}$, then nondeterministic poly-time procedures are computably more succinct than deterministic poly-time procedures for some finite initial segments of a particular NP-complete set. By way of contrast, Hartmanis and Baker also showed that deterministic Turing Machines (TMs) that provably run in poly-time are computably more succinct than nondeterministic TMs that are explicitly clocked to run in poly-time. Hartmanis [16] proved that nondeterministic PDAs are computably more succinct than deterministic PDAs. Hay [19] showed how to improve from computable, to $\emptyset^{(1)}$-computable, the relative succinctness results mentioned above of $[16,26,33,39]$. She also showed that TMs are computably more succinct than finite state automata for singleton sets. Hartmanis [17] obtained a strong computability-theoretic, sufficient condition for computable relative succinctness between two programming systems. Royer and Case [31] constructed a systematic framework for relative succinctness between subrecursive programming systems, characterized when relativesuccinctness phenomena occur between programming systems, and introduced the notion of relative-succinctness progressions (as in Theorem 9) and established a number of such progressions through a variety of subrecursive and complexitybounded hierarchies.

Beyond the subrecursive context of the above results, Parikh [30, p. 216] showed that, for each computable $h: \mathbb{N} \rightarrow \mathbb{N}$, there is: (i) an $i_{1}$ with $W_{i_{1}} \in \Delta_{1}^{0}$ such that $\min \left\{p: W_{p}=\overline{W_{i_{1}}}\right\}>h\left(i_{1}\right)$, and (ii) an $i_{2}$ with $W_{i_{2}}^{\emptyset^{(1)}} \in \Sigma_{1}^{0}$ such that $\min \left\{p: W_{p}=W_{i_{2}}^{\emptyset^{(1)}}\right\}>h\left(i_{2}\right)$. Meyer and Bagchi [25] showed that, for sets $A$ and $B$ with $A^{(1)} \leq_{\mathrm{T}} B$, for each $B^{(2)}$-computable $H$, there is an $e$ such that $W_{e}^{B} \in \Sigma_{1}^{A}$, but $\min \left\{s_{A}(p): W_{p}^{A}=W_{e}^{B}\right\}>H\left(s_{B}(e)\right)$, where $s_{A}$ and $s_{B}$ are computable size measures; they conjectured that this $W_{e}^{B}$ could be made co-finite. (Theorem 7 above confirms and improves slightly upon this conjecture.)

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# Automorphism Groups of Substructure Lattices of Vector Spaces in Computable Algebra 

Rumen Dimitrov ${ }^{1(\boxtimes)}$, Valentina Harizanov ${ }^{2}$, and Andrei Morozov ${ }^{3}$<br>${ }^{1}$ Department of Mathematics, Western Illinois University, Macomb, IL 61455, USA<br>rd-dimitrov@wiu.edu<br>${ }^{2}$ Department of Mathematics, George Washington University, Washington, D.C. 20052, USA<br>harizanv@gwu.edu<br>${ }^{3}$ Sobolev Institute of Mathmematics, Novosibirsk 630090, Russia<br>morozov@math.nsc.ru


#### Abstract

For a Turing degree $\mathbf{x}$, we investigate the automorphisms of the lattice of $\mathbf{x}$-c.e. vector spaces. We establish the equivalence of the embedding relation for these automorphism groups with the order relation on the corresponding Turing degrees. By a result of Guichard the automorphisms of the lattice of $\mathbf{x}$-c.e. vector spaces are induced by $\mathbf{x}$-computable invertible semilinear transformations, GSL $_{\mathbf{x}}$. We prove that the Turing degree spectrum of the group GSL $\mathbf{x}_{\mathbf{x}}$ is the upper cone of Turing degrees $\geq \mathbf{x}^{\prime \prime}$.


## 1 Automorphisms of Effective Structures

The study of automorphisms on computable or computably enumerable structures connects computability theory and classical group theory. The set of all automorphisms of a computable structure forms a group under composition, and it is natural to ask questions about its complexity as well as the complexity of its subgroups. It is also interesting to connect the embedability of the subgroups with Turing reducibility.

The following notion is the focus of our investigation. Let $\mathbf{d}$ be a Turing degree. For an infinite computable structure $\mathcal{M}$, we define $A u t_{\mathbf{d}}(\mathcal{M})$ to be the set of all automorphisms of $\mathcal{M}$ computable in $\mathbf{d}$. The set $A u t_{\mathbf{d}}(\mathcal{M})$ under composition is a subgroup of $\operatorname{Aut}(\mathcal{M})$. When the structure $\mathcal{M}$ is $\omega$ with equality, then its automorphism group $\operatorname{Aut}(\mathcal{M})$ is usually denoted by $\operatorname{Sym}(\omega)$, the symmetric group of $\omega$. Hence we have

$$
\operatorname{Sym}_{\mathbf{d}}(\omega)=\{f \in \operatorname{Sym}(\omega): \operatorname{deg}(f) \leq \mathbf{d}\}
$$

where $\operatorname{deg}(f)$ is the Turing degree of $f$. Our other computability theoretic notation is also standard and as in [18].

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The Turing degree spectrum of a countable structure $\mathcal{A}$ is

$$
D g S p(\mathcal{A})=\{\operatorname{deg}(\mathcal{B}): \mathcal{B} \cong \mathcal{A}\}
$$

where $\operatorname{deg}(\mathcal{B})$ is the Turing degree of the atomic diagram of $\mathcal{B}$. Knight [8] proved that the degree spectrum of any structure is either a singleton or is upward closed. Only the degree spectrum of a so-called automorphically trivial structure is a singleton, and if the language is finite, that degree must be $\mathbf{0}$ (see [7]). Automorphically trivial structures include all finite structures, and also some special infinite structures, such as the complete graph on countably many vertices. Jockusch and Richter (see [15]) defined the degree of the isomorphism type of a structure, if it exists, to be the least Turing degree in its Turing degree spectrum. Richter [15,16] was first to systematically study such degrees. For these and more recent results about these degrees see [4]. In this paper we are especially interested in the following result by Morozov.

Theorem 1 [10]. The degree of the isomorphism type of the group $\operatorname{Sym}_{\mathbf{d}}(\omega)$ is $\mathbf{d}^{\prime \prime}$.

We will establish a similar result in the context of effective vector spaces.
Let $V_{\infty}$ be a canonical fully effective $\aleph_{0}$-dimensional vector space over a computable field $F$. We can think of the vectors in $V_{\infty}$ as (the codes of) the finitely non-zero $\omega$-sequences of elements of $F$. By $\mathcal{L}$ we denote the lattice of all subspaces of $V_{\infty}$. For a Turing degree $\mathbf{d}$, by $\mathcal{L}_{\mathbf{d}}\left(V_{\infty}\right)$ we denote the following sublattice of $\mathcal{L}$ :

$$
\mathcal{L}_{\mathbf{d}}\left(V_{\infty}\right)=\{V \in \mathcal{L}: V \text { is } \mathbf{d} \text {-computably enumerable }\} .
$$

Note that in the literature $\mathcal{L}_{\mathbf{0}}\left(V_{\infty}\right)$ is usually denoted by $\mathcal{L}\left(V_{\infty}\right)$. Guichard [6] established that there are countably many automorphisms of $\mathcal{L}_{\mathbf{0}}\left(V_{\infty}\right)$ by showing that each computable automorphism is generated by a $1-1$ and onto computable semilinear transformation of $V_{\infty}$. Recall that a map $\mu: V_{\infty} \rightarrow V_{\infty}$ is called a semilinear transformation of $V_{\infty}$ if there is an automorphism $\sigma$ of $F$ such that

$$
\mu(\alpha u+\beta v)=\sigma(\alpha) \mu(u)+\sigma(\beta) \mu(v)
$$

for every $u, v \in V_{\infty}$ and every $\alpha, \beta \in F$.
Notation 2. By $G S L_{\mathbf{d}}$ we denote the group of $1-1$ and onto semilinear transformations $\langle\mu, \sigma\rangle$ such that $\operatorname{deg}(\mu) \leq \mathbf{d}$ and $\operatorname{deg}(\sigma) \leq \mathbf{d}$.

Hence Guichard proved that every element of $\operatorname{Aut}\left(\mathcal{L}_{\mathbf{0}}\left(V_{\infty}\right)\right)$ is generated by an element of $G S L_{\mathbf{0}}$. This result can be relativized to an arbitrary Turing degree d. The proof of Theorem 3 below is essentially identical to the proof in [6].

Theorem 3 [6]. Every $\Phi \in \operatorname{Aut}\left(\mathcal{L}_{\mathbf{d}}\left(V_{\infty}\right)\right)$ is generated by some $\langle\mu, \sigma\rangle \in G S L_{\mathbf{d}}$. Moreover if $\Phi$ is also generated by some other $\left\langle\mu_{1}, \sigma_{1}\right\rangle \in G S L_{\mathbf{d}}$, then there is $\gamma \in F$ such that

$$
\left(\forall v \in V_{\infty}\right)\left[\mu(v)=\gamma \mu_{1}(v)\right] .
$$

There are two main results in this paper. The first is Theorem 4 in Sect. 2, which establishes that for every pair $\mathbf{a}, \mathbf{b}$ of Turing degrees, we have $\operatorname{Aut}\left(\mathcal{L}_{\mathbf{a}}\left(V_{\infty}\right)\right) \hookrightarrow$ $\operatorname{Aut}\left(\mathcal{L}_{\mathbf{b}}\left(V_{\infty}\right)\right)$ if and only if $\mathbf{a} \leq \mathbf{b}$. The second main result is Theorem 7 in Sect. 3, which establishes that the isomorphism degree type of the group $G S L_{\mathbf{d}}$ is $\mathbf{d}^{\prime \prime}$.

## 2 Group Embeddings and Turing Reducibility

Morozov showed that the correspondence $\mathbf{a} \rightarrow \operatorname{Sym}_{\mathbf{a}}(\omega)$ can be used to substitute Turing reducibility with group-theoretic embedding. More precisely, Morozov [11] established that

$$
\left(\operatorname{Sym}_{\mathbf{a}}(\omega) \hookrightarrow \operatorname{Sym}_{\mathbf{b}}(\omega)\right) \Leftrightarrow(\mathbf{a} \leq \mathbf{b})
$$

for every pair $\mathbf{a}, \mathbf{b}$ of Turing degrees. It follows from this result that $\mathbf{a}=\mathbf{b}$ if and only if $\operatorname{Sym}_{\mathbf{a}}(\omega) \cong \operatorname{Sym}_{\mathbf{b}}(\omega)$. Here, we establish an analogous result for the subgroups of the group of automorphisms of the corresponding sublattices of $\mathcal{L}$. In the proof of the next, main theorem we will use the standard notation: $[x, y]=x^{-1} y^{-1} x y$ and $x^{y}=y^{-1} x y$.

Theorem 4. For any pair of Turing degrees a, b we have

$$
\left(\operatorname{Aut}\left(\mathcal{L}_{\mathbf{a}}\left(V_{\infty}\right)\right) \hookrightarrow \operatorname{Aut}\left(\mathcal{L}_{\mathbf{b}}\left(V_{\infty}\right)\right)\right) \Leftrightarrow \mathbf{a} \leq \mathbf{b} .
$$

Proof. Obviously, if $\mathbf{a} \leq \mathbf{b}$, then $\operatorname{Aut}\left(\mathcal{L}_{\mathbf{a}}\left(V_{\infty}\right)\right) \hookrightarrow \operatorname{Aut}\left(\mathcal{L}_{\mathbf{b}}\left(V_{\infty}\right)\right)$.
Now, assume that $\operatorname{Aut}\left(\mathcal{L}_{\mathbf{a}}\left(V_{\infty}\right)\right) \hookrightarrow \operatorname{Aut}\left(\mathcal{L}_{\mathbf{b}}\left(V_{\infty}\right)\right)$. Let $\left\{e_{0}, e_{1}, \ldots\right\}$ be a fixed computable basis of $V_{\infty}$. For $\left\langle\mu_{1}, \sigma_{1}\right\rangle,\left\langle\mu_{2}, \sigma_{2}\right\rangle \in G S L_{\mathbf{a}}$, we define $\left\langle\mu_{1}, \sigma_{1}\right\rangle \sim$ $\left\langle\mu_{2}, \sigma_{2}\right\rangle$ iff:
(1) $\sigma_{1}=\sigma_{2}$, and
(2) there is $\alpha \in F$ such that $\alpha \neq 0$ and $\left(\forall v \in V_{\infty}\right)\left[\mu_{1}(v)=\alpha \mu_{2}(v)\right]$.

Note that $\operatorname{Aut}\left(\mathcal{L}_{\mathbf{a}}\left(V_{\infty}\right)\right) \cong G S L_{\mathbf{a}} / \sim$. We can define a group embedding $\delta$ : $\operatorname{Sym}_{\mathbf{a}}(\omega) \hookrightarrow G S L_{\mathbf{a}} / \sim$ as follows. For any $f \in \operatorname{Sym}_{\mathbf{a}}(\omega)$, we let $\delta(f)$ be the $\sim$-equivalence class of a linear transformation $\langle\widetilde{f}, i d\rangle$ such that

$$
\widetilde{f}\left(e_{i}\right)=e_{f(i)}
$$

Note that if $\delta\left(f_{1}\right)=\delta\left(f_{2}\right)$, then $\widetilde{f}_{1}=c \widetilde{f}_{2}$ for some $c \in F$, and thus

$$
(\forall i \in \omega)\left[e_{f_{1}(i)}=\widetilde{f}_{1}\left(e_{i}\right)=c \widetilde{f}_{2}\left(e_{i}\right)=c e_{f_{2}(i)}\right] .
$$

Since the vectors $e_{i}, i \in \omega$, are independent, we must have

$$
(\forall i \in \omega)\left[f_{1}(i)=f_{2}(i)\right] .
$$

Therefore, there exists a map

$$
K: \operatorname{Sym}_{\mathbf{a}}(\omega) \hookrightarrow G S L_{\mathbf{b}} / \sim
$$

such that if $f \in \operatorname{Sym}_{\mathbf{a}}(\omega)$, then $K(f)$ is a b-computable linear transformation of $V_{\infty}$ modulo scalar multiplication.

We claim that if a set $A$ is c.e. in $\mathbf{a}$, then $A$ is c.e. in $\mathbf{b}$. Fix $A \subseteq \omega$ such that $A$ is c.e. in a , and let $h: \omega \rightarrow \omega$ be an a-computable enumeration of $A$. Hence $r n g(h)=A$. Fix a partition of the natural numbers into uniformly computable infinite sets $R_{i}$ for $i \in \mathbb{Z}$ with enumerations $R_{i}=\left\{c_{i}^{0}<c_{i}^{1}<\cdots\right\}$. Let the permutations $g_{0}, g_{1}, w, b \in \operatorname{Sym}_{\mathbf{a}}(\omega)$ be defined as follows:
$w\left(c_{i}^{j}\right)=c_{i+1}^{j}$ for each $i \in Z$ and $j \in \omega$,
$g_{0}=\prod_{j \in \omega}\left(c_{0}^{2 j}, c_{0}^{2 j+1}\right)$,
$g_{1}=\prod_{j \in \omega}\left(c_{0}^{2 j+1}, c_{0}^{2 j+2}\right)$, and
$\left.b=\prod_{n, t \in \omega} \wedge h(t)=n=1 c_{n}^{t}, c_{n}^{t+1}\right)$.
We will also use the following abbreviation: $w^{n}=\underbrace{w \cdots w}_{n \text { times }}$. Then we have

$$
n \notin A \Leftrightarrow\left(\left[g_{0}, b^{w^{n}}\right]=1 \wedge\left[g_{1}, b^{w^{n}}\right]=1\right) .
$$

This is because $g_{0}$ and $b^{w^{n}}$ commute iff $n$ is not enumerated into $A$ at an odd stage $t$, and, similarly, $g_{1}$ and $b^{w^{n}}$ commute iff $n$ is not enumerated into $A$ at an even stage $t$. Hence for $\widetilde{g_{0}}=K\left(g_{0}\right), \widetilde{g_{1}}=K\left(g_{1}\right), \widetilde{w}=K(w)$, and $\widetilde{b}=K(b)$, we have

$$
\begin{gathered}
n \notin A \Leftrightarrow\left(\left[K\left(g_{0}\right), K(b)^{(K(w))^{n}}\right]=1 \wedge\left[K\left(g_{1}\right), K(b)^{(K(w))^{n}}\right]=1\right) \\
\Leftrightarrow\left(\left[\widetilde{g_{0}}, \widetilde{b}^{\widetilde{w}^{n}}\right] / \sim=1 \wedge\left[\widetilde{g}_{1}, \widetilde{b}^{\widetilde{w}^{n}}\right] / \sim=1\right)
\end{gathered}
$$

We will now show that $\left[\widetilde{g_{0}}, \widetilde{b^{w}} \widetilde{w}^{n}\right] \nsim 1$ is c.e. relative to $\mathbf{b}$. Let $\tau_{n}={ }_{\text {def }}\left[\widetilde{g_{0}}, \widetilde{b}^{w^{n}}\right]$. Then

$$
\begin{gathered}
\tau_{n} \nsim 1 \Leftrightarrow \tau_{n}\left(e_{0}\right) \text { and } e_{0} \text { are linearly independent, or } \\
(\exists m \in \omega)(\exists \alpha \neq 0)\left[\tau_{n}\left(e_{0}\right)=\alpha e_{0} \wedge \tau_{n}\left(e_{m}\right) \neq \alpha e_{m}\right] .
\end{gathered}
$$

Let $A \in \mathbf{a}$. Then $A$ and $\bar{A}$ are both c.e. in $\mathbf{b}$, and, therefore, $A$ is computable in $\mathbf{b}$. Hence $\mathbf{a} \leq \mathbf{b}$.

## 3 Complexity of $G S L_{\mathrm{d}}$

In this section we will determine the Turing degree spectrum of $G S L_{\mathbf{d}}$. For the statement of the main theorem we will use terminology and notation from the following definition.

Definition 1. A permutation $p$ on a set $M$ is:
(i) $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ on $M$ if it is a product of infinitely many 1 -cycles and infinitely many 2-cycles;
(ii) $1_{\mathrm{inf}} 2_{\text {fin }}$ on $M$ if it is a product of infinitely many 1-cycles and finitely many 2-cycles.

The main theorem about the degree spectrum of $G S L_{\mathbf{d}}$ will be derived from the following embeddability theorem.

Theorem 5. Let $G$ be an $X$-computable group, and let $H: \operatorname{Sym}_{\mathbf{0}}(\omega) \hookrightarrow G$ be an embedding. Suppose that for every $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ permutation $p \in \operatorname{Sym}_{\mathbf{0}}(\omega)$, the image $H(p)$ is not a conjugate of the image of any $1_{\mathrm{inf}} 2_{\text {fin }}$ permutation in $\operatorname{Sym}_{\mathbf{0}}(\omega)$.

Then $\mathbf{0}^{\prime \prime} \leq \operatorname{deg}(X)$.
Proof. Let $A$ be a $\Pi_{2}^{0}$-complete set and let $R(x, t)$ be a computable predicate such that

$$
n \in A \Leftrightarrow\left(\exists^{\infty} t\right) R(n, t) .
$$

We will prove that $A \leq_{T} X$. Fix a partition of the natural numbers into uniformly computable infinite sets $S_{i, j}$ for $i \in \mathbb{Z}$ and $j \in\{1,2\}$ with enumerations $S_{i, j}=$ $\left\{c_{i, j}^{0}<c_{i, j}^{1}<\cdots\right\}$. The sets $S_{i, 1}$ and $S_{i, 2}$ will be referred to as the left and the right parts of the $i$-th column, $S_{i}=S_{i, 1} \cup S_{i, 2}$. This reference will be useful when we define certain maps below. We can graphically present this partition as follows:


We will now define the following maps.
(i) $w\left(c_{i+1, j}^{k}\right)={ }_{\text {def }} c_{i, j}^{k}$ for each $i \in \mathbb{Z}, k \in \omega$ and $j=1,2$.

Clearly, the map $w$ is such that $w\left(S_{i+1,1}\right)=S_{i, 1}$ and $w\left(S_{i+1,2}\right)=S_{i, 2}$. It maps the left (right) part of the ( $i+1$ )-st column to the left (right) part of the $i$-th column for each $i$.
(ii) $p_{0}=\operatorname{def} \prod_{k \in \omega}\left(c_{0,1}^{k}, c_{0,2}^{k}\right)$.

It is easy to see that the map $p_{0}$ switches the left and right parts of the 0-th column (i.e., $p_{0}\left(S_{0,1}\right)=S_{0,2}$ and $\left.p_{0}\left(S_{0,2}\right)=S_{0,1}\right)$, and is identity on all other elements of $\omega$.
(iii) $p_{n}={ }_{\text {def }} p_{0}^{w^{n}}=w^{-n} p_{0} w^{n}$.

Note that the map $p_{n}$ switches the left and right parts of the $n$-th column (i.e., $p_{n}\left(S_{n, 1}\right)=S_{n, 2}$ and $\left.p_{n}\left(S_{n, 2}\right)=S_{n, 1}\right)$, and is identity on all other elements of $\omega$.
(iv) $z(k)={ }_{\text {def }}\left\{\begin{array}{c}0 \quad \text { if } \quad k=0, \\ 1 \quad \text { if } \quad k=2, \\ k-2 \text { if } k=2 t \geq 4, \\ k+2 \text { if } k=2 t+1 .\end{array}\right.$

Note that the map $z$ is a permutation of $\omega$, which contains only one infinite cycle and (0).
(v) $\tau={ }_{d e f}(0,1)$.

For $k \in \mathbb{Z}$ we have

$$
\tau^{z^{k}}=\left\{\begin{array}{cc}
(0,2 k), & \text { if } k \geq 1 \\
(0,2|k|+1), & \text { if } k \leq 0
\end{array}\right.
$$

so

$$
\begin{equation*}
(\forall n, m \in \omega)\left(\exists n_{1}, m_{1} \in \mathbb{Z}\right)\left[\left(\tau^{z^{n_{1}}}\right)^{\tau^{z^{m_{1}}}}=(n, m)\right] . \tag{1}
\end{equation*}
$$

Note that property (1) guarantees that any $1_{\text {inf }} 2_{\text {fin }}$ permutation on $\omega$ can be represented as a finite product of the permutations $\tau$ and $z$.
(vi) We will now construct a permutation $b$ on $\omega$ with the following properties:

$$
\begin{aligned}
& b \upharpoonright_{S_{n, 1}}=i d \upharpoonright_{S_{n, 1}} \\
& b \upharpoonright_{S_{n, 2}} \text { is }\left\{\begin{array}{l}
1_{\mathrm{inf}} 2_{\mathrm{inf}}, \text { if } n \in A, \\
1_{\inf } 2_{\mathrm{fin}},
\end{array} \text { if } n \notin A .\right.
\end{aligned}
$$

We will define $b$ in stages. At each stage $s$ we will have $E^{s}=\operatorname{def} \operatorname{dom}\left(b^{s}\right)=$ $r n g\left(b^{s}\right)$.
Construction
Stage 0.
Let $b^{0} \upharpoonright S_{i}={ }_{\text {def }} i d$ for $i \leq-1$, and $E^{0}=\bigcup_{i \leq-1} S_{i}$.
Stage $s+1=\langle n, t\rangle$.
Case 1. If $R(n, t)$, then find the least elements $p, q, r \in S_{n, 2}$ such that $p, q, r \notin E^{s}$. Let $b^{s+1}=b^{s} \cdot(p, q)$ and assume that $b^{s+1}(r)=r$. Thus, we have $E^{s+1}=$ $E^{s} \cup\{p, q, r\}$ and $b^{s+1} \upharpoonright E^{s}=b^{s}$.
Case 2. If $\neg R(n, t)$, then find the least elements $p, q, r \in S_{n, 2}$ such that $p, q, r \notin$ $E^{s}$. Let $b^{s+1} \upharpoonright E^{s}=b^{s}$ and $b^{s+1}(p)=p, b^{s+1}(q)=q, b^{s+1}(r)=r$. Then $E^{s+1}=E^{s} \cup\{p, q, r\}$.

End of construction.
By construction, $\operatorname{dom}(b)=\operatorname{rng}(b)=\omega$.
It follows that if $n \in A$, then $\left(\exists^{\infty} t\right) R(n, t)$, so Case 1 applies infinitely often for this $n$, and hence the map $b$ switches infinitely many pairs in the right part of the $n$-th column. Therefore, $b \upharpoonright S_{n, 2}$ is $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ and $b \upharpoonright S_{n, 1}=i d$.

If $n \notin A$, then $(\exists<\infty t) R(n, t)$, so Case 1 applies finitely often for this $n$, and hence the map $b$ switches only finitely many pairs in the right part of the $n$-th column. Therefore, $b \upharpoonright S_{n, 2}$ is $1_{\text {inf }} 2_{\text {fin }}$ and $b \upharpoonright S_{n, 1}=i d$.

In both cases, the map $b^{p_{n}}$ reverses the action of $b$ on the left and right parts of the $n$-th column $S_{n}$, while for $k \neq n$, we have $b^{p_{n}} \upharpoonright S_{k}=b \upharpoonright S_{k}$.

Then $b \cdot b^{p_{n}}$ is $\begin{cases}1_{\text {inf }} 2_{\text {inf }} \text { on } S_{n}, & \text { if } n \in A, \\ 1_{\inf } 2_{\text {fin }} \text { on } S_{n}, & \text { if } n \notin A, \\ i d \text { on } S_{k}, & \text { if } n \neq k .\end{cases}$

Therefore, $b \cdot b^{p_{n}}$ is $\left\{\begin{array}{l}1_{\mathrm{inf}} 2_{\mathrm{inf}} \text { on } \omega, \text { if } n \in A, \\ 1_{\mathrm{inf}} 2_{\mathrm{fin}} \text { on } \omega, \text { if } n \notin A .\end{array}\right.$
Finally, note that on $\omega$, every computable $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ permutation is the conjugate of a fixed computable $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ permutation and some other computable permutation. Therefore, assume that $f$ is a fixed computable $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ permutation such that:

$$
\left(\forall z_{1} \in \operatorname{Sym}_{\mathbf{0}}(\omega)\right)\left(\exists h \in \operatorname{Sym}_{\mathbf{0}}(\omega)\right)\left[z_{1}=f^{h}\right] .
$$

Hence for every $n$, we have

$$
\begin{align*}
n \in A & \Leftrightarrow b \cdot b^{p_{n}} \text { is a } 1_{\mathrm{inf}} 2_{\mathrm{inf}} \text { permutation on } \omega \\
& \Leftrightarrow\left(\exists h \in \operatorname{Sym}_{\mathbf{0}}(\omega)\right)\left[b \cdot b^{p_{n}}=f^{h}\right]  \tag{2}\\
& \Leftrightarrow\left(\exists u \in H\left(\operatorname{Sym}_{\mathbf{0}}(\omega)\right)\right)\left[H(b) \cdot H(b)^{H\left(p_{n}\right)}=H(f)^{u}\right], \text { and }
\end{align*}
$$

$$
\begin{align*}
n \notin A & \Leftrightarrow b \cdot b^{p_{n}} \text { is a } 1_{\text {inf }} 2_{\text {fin }} \text { permutation on } \omega \\
& \Leftrightarrow b \cdot b^{p_{n}}=\prod_{(i, j) \in F}\left(\tau^{z^{i}}\right)^{\tau^{z^{j}}}  \tag{3}\\
& \Leftrightarrow H(b) \cdot H(b)^{H\left(p_{n}\right)}=\prod_{(i, j) \in F}\left(H(\tau)^{H(z)^{i}}\right)^{H(\tau)^{H(z)^{j}}} .
\end{align*}
$$

The set $F$ in the last line of (3) denotes some finite set of pairwise disjoint cycles and the maps referenced in (2) and (3) are those that we defined in (i)-(vi) above. For the map $H: \operatorname{Sym}_{\mathbf{0}}(\omega) \hookrightarrow G$ note that $H\left(p_{n}\right)=H(w)^{-n} \cdot H\left(p_{0}\right)$. $H(w)^{n}$.

We claim that the last equivalence in (2) can be strengthened so that we have:

$$
\begin{equation*}
n \in A \Leftrightarrow(\exists u \in G)\left[H(z)=H(f)^{u}\right] \tag{4}
\end{equation*}
$$

That is, if $n \in A$, then

$$
\begin{aligned}
\left(\exists h \in \operatorname{Sym}_{\mathbf{0}}(\omega)\right)\left[b \cdot b^{p_{n}}\right. & \left.=f^{h}\right] \text { and, therefore, } \\
\left(\exists u \in H\left(\operatorname{Sym}_{0}(\omega)\right)\right)\left[H(b) \cdot H(b)^{H\left(p_{n}\right)}\right. & \left.=H(f)^{u}\right], \text { and } \\
(\exists u \in G)\left[H(b) \cdot H(b)^{H\left(p_{n}\right)}\right. & \left.=H(f)^{u}\right] .
\end{aligned}
$$

For the proof of the other direction of (4) suppose that for some fixed $h \in G$ we have

$$
H(b) \cdot H(b)^{H\left(p_{n}\right)}=H(f)^{h}, \text { but } n \notin A
$$

Then, because of (3), we have the following:
(i) $b \cdot b^{p_{n}}$ is a $1_{\text {inf }} 2_{\text {fin }}$ permutation on $\omega$,
(ii) $H(b) \cdot H(b)^{H\left(p_{n}\right)}$ is the image of the $1_{\text {inf }} 2_{\text {fin }}$ permutation $b \cdot b^{p_{n}}$, while
(iii) $H(f)$ is the image of the $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ permutation $f$.

This contradicts our assumption that the image under $H$ of the $1_{\mathrm{inf}} 2_{\mathrm{fin}}$ permutation $b \cdot b^{p_{n}}$ cannot be the conjugate of the image of the $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ permutation $f$.

Theorem 6. The degree of the isomorphisms type of the group $G S L_{\mathbf{0}}$ is $\mathbf{0}^{\prime \prime}$.
Proof. Let $V=\left\{v_{0}, v_{1}, \ldots\right\}$ be a computable basis of $V_{\infty}$. Define

$$
H: \operatorname{Sym}_{\mathbf{0}}(\omega) \hookrightarrow G S L_{\mathbf{0}}
$$

so that for any $p \in \operatorname{Sym}_{\mathbf{0}}(\omega)$ the image $H(p)=\langle L, i d\rangle$ is a semilinear map such that

$$
L\left(v_{i}\right)=v_{p(i)} \text { for every } i \in \omega
$$

We claim that under $H$ the image of a $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ permutation from $\operatorname{Sym}_{\mathbf{0}}(\omega)$ cannot be a conjugate of the image of a $1_{\text {inf }} 2_{\text {fin }}$ permutation from $S y m_{0}(\omega)$. To establish this fact, suppose that $\langle f, i d\rangle,\left\langle f_{1}, i d\right\rangle \in G S L_{0}$ are the images of some $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ and $1_{\mathrm{inf}} 2_{\mathrm{fin}}$ computable permutations on $\omega$, respectively. Suppose that $\langle f, i d\rangle$ and $\left\langle f_{1}, i d\right\rangle$ are conjugates, and let $\langle h, \sigma\rangle \in G S L_{\mathbf{0}}$ be such that $\langle f, i d\rangle{ }^{\langle h, \sigma\rangle}=$ $\left\langle f_{1}, i d\right\rangle$. Note that the map $h: V_{\infty} \rightarrow V_{\infty}$ is $1-1$ and onto. The associated field automorphism $\sigma: F \rightarrow F$ from $\langle h, \sigma\rangle$ is used to indicate that $h(a v+b w)=$ $\sigma(a) h(v)+\sigma(b) h(w)$. To simplify the notation, we will refer to the semilinear maps $\langle f, i d\rangle,\left\langle f_{1}, i d\right\rangle$, and $\langle h, \sigma\rangle$ simply as $f, f_{1}$, and $h$, respectively.

Note that the definition of the map $H$ allows us to view $f \upharpoonright V$ and $f_{1} \upharpoonright V$ as $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ and $1_{\mathrm{inf}} 2_{\text {fin }}$ permutations on $V$, respectively. We now claim that $f_{1}$ satisfies the following property:

$$
\begin{equation*}
\left(\exists W \subset_{f i n} V_{\infty}\right)\left(\forall v \in V_{\infty}\right)\left[\left(v-f_{1}(v)\right) \in W\right] . \tag{5}
\end{equation*}
$$

Here, $W \subset_{f i n} V_{\infty}$ stands for $W$ being a finite-dimensional subspace of $V_{\infty}$. For a set $U \subseteq V_{\infty}$, by $\operatorname{cl}(U)$ we will denote the closure of $U$, which is the set of all linear combinations of the vectors in $U$. To prove (5), assume that $B=\left\{x_{1}, \ldots, x_{k}, y_{1}, \ldots, y_{k}\right\} \subseteq V$ is such that $f_{1} \upharpoonright V=\prod_{1 \leq i \leq k}\left(x_{i}, y_{i}\right)$. Note that for every $v \in V_{\infty}$, there are $v_{1} \in \operatorname{cl}(V-B)$ and $v_{2} \in \operatorname{cl}(B)$ such that $v=v_{1}+v_{2}$. Then

$$
f_{1}(v)=f_{1}\left(v_{1}\right)+f_{1}\left(v_{2}\right)=v_{1}+f_{1}\left(v_{2}\right),
$$

and so

$$
v-f_{1}(v)=v_{1}+v_{2}-v_{1}-f_{1}\left(v_{2}\right)=v_{2}-f_{1}\left(v_{2}\right) \in \operatorname{cl}(B),
$$

because $f_{1}\left(v_{2}\right) \in \operatorname{cl}(B)$. Therefore, $W=\operatorname{cl}(B)$ is a finite-dimensional subspace of $V_{\infty}$ for which property $(D)$ holds.

We will now prove that $f^{h}$ does not satisfy property (5), which will contradict the assumption that $f^{h}=f_{1}$. Thus, assume that $W$ is a finite-dimensional subspace of $V_{\infty}$ such that

$$
\begin{equation*}
\left(\forall x \in V_{\infty}\right)\left[\left(x-f^{h}(x)\right) \in W\right] \tag{6}
\end{equation*}
$$

The support of a vector $x$ with respect to a basis $Z=\left\{z_{j}: j \in J\right\}$, denoted by $\operatorname{supp}_{Z}(x)$, is the set $\left\{z_{j_{l}}: l \in\{0, \ldots, t\}\right\}$ such that

$$
x=\sum_{l=0}^{t} \lambda_{l} z_{j_{l}}
$$

and $\left(\forall l \in\{0, \ldots, t)\left[\lambda_{l} \neq 0\right]\right.$.
Let $W_{1}=h(W)$ and note that $W_{1}$ is finite-dimensional. Let $B_{1}$ be a finite subset of the basis $V$ such that

$$
\left(\forall x \in W_{1}\right)\left[\operatorname{supp}_{V}(x) \subseteq B_{1}\right]
$$

We will now find $u_{1} \in V_{\infty}$ such that $u_{1}-f\left(u_{1}\right) \notin W_{1}$. Since $f \upharpoonright V$ is a $1_{\inf } 2_{\inf }$ permutation on $V$, there are infinitely many pairs $(u, v) \in V \times V$ such that

$$
\begin{equation*}
u \neq v, f(u)=v \text { and } f(v)=u \tag{7}
\end{equation*}
$$

Since $B_{1}$ is finite, we can also find $u_{1}, v_{1} \in V-B_{1}$, which have property (7). Then:
(i) $u_{1}-f\left(u_{1}\right)=u_{1}-v_{1} \neq 0$, and
(ii) $u_{1}-f\left(u_{1}\right)=\left(u_{1}-v_{1}\right) \notin \operatorname{cl}\left(B_{1}\right)$ because $B_{1} \cup\left\{u_{1}, v_{1}\right\} \subseteq V$.

Since $W_{1} \subseteq \operatorname{cl}\left(B_{1}\right)$, we have that $u_{1}-f\left(u_{1}\right) \notin W_{1}$. Therefore,

$$
\begin{aligned}
\left(h^{-1}\left(u_{1}\right)-h^{-1}\left(f\left(u_{1}\right)\right)\right) & \notin h^{-1}\left(W_{1}\right), \text { and so } \\
\left(h^{-1}\left(u_{1}\right)-h^{-1} f h h^{-1}\left(u_{1}\right)\right) & \notin W .
\end{aligned}
$$

If we let $x_{1}=h^{-1}\left(u_{1}\right)$, we obtain

$$
x_{1}-f^{h}\left(x_{1}\right) \notin W,
$$

which contradicts that $f^{h}$ satisfies (6).
We constructed an embedding $H: \operatorname{Sym}_{\mathbf{0}}(\omega) \hookrightarrow G S L_{\mathbf{0}}$ such that the images of any $1_{\mathrm{inf}} 2_{\mathrm{inf}}$ and $1_{\mathrm{inf}} 2_{\mathrm{fin}}$ permutations from $\operatorname{Sym}_{\mathbf{0}}(\omega)$ cannot be conjugates in $G S L_{\mathbf{0}}$. We use Theorem 5 to conclude that $\mathbf{0}^{\prime \prime}$ is computable in any copy of $G S L_{\mathbf{0}}$. We can construct a copy of $G S L_{\mathbf{0}}$, which is computable in $\mathbf{0}^{\prime \prime}$. Therefore, the degree of the isomorphisms type of $G S L_{\mathbf{0}}$ is $\mathbf{0}^{\prime \prime}$.

Note that the result of the previous theorem can be easily relativized to any Turing degree $\mathbf{d}$.

Theorem 7. The degree of the isomorphisms type of the group $G S L_{\mathbf{d}}$ is $\mathbf{d}^{\prime \prime}$.

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# Parameterized Complexity and Approximation Issues for the Colorful Components Problems 

Riccardo Dondi ${ }^{1}$ and Florian Sikora ${ }^{2(\boxtimes)}$<br>${ }^{1}$ Dipartimento di Scienze umane e sociali, Università degli Studi di Bergamo, Bergamo, Italy<br>riccardo.dondi@unibg.it<br>${ }^{2}$ Université Paris-Dauphine, PSL, CNRS UMR 7243, LAMSADE, Paris, France<br>florian.sikora@dauphine.fr


#### Abstract

The quest for colorful components (connected components where each color is associated with at most one vertex) inside a vertexcolored graph has been widely considered in the last ten years. Here we consider two variants, Minimum Colorful Components (MCC) and Maximum Edges in transitive Closure (MEC), introduced in the context of orthology gene identification in bioinformatics. The input of both MCC and MEC is a vertex-colored graph. MCC asks for the removal of a subset of edges, so that the resulting graph is partitioned in the minimum number of colorful connected components; MEC asks for the removal of a subset of edges, so that the resulting graph is partitioned in colorful connected components and the number of edges in the transitive closure of such a graph is maximized. We study the parameterized and approximation complexity of MCC and MEC, for general and restricted instances.


## 1 Introduction

The quest for colorful components inside a vertex colored graph has been a widely investigated problem in the last years, with application for example in bioinformatics $[5,8,12]$. Roughly speaking, given a vertex-colored graph, the problem asks to find the colorful components of the graph, that is connected components that contain at most one vertex of each color. While most of the approaches have focused on the identification of a single connected colorful component, the identification of the minimum number of colorful connected components that match a given motif has been considered in [4, 7].

Here we consider a similar framework, where instead of looking for a single colorful component inside a vertex-colored graph, we ask for a partition of the graph vertices in colorful components. This approach stems from a problem in bioinformatics, and more specifically in comparative genomics. In this context, a fundamental task is to infer the relations between genes in different genomes and, more precisely, to infer which genes are orthologous, that is those genes that originate via a speciation event from a gene of an ancestral genome.

A graph approach has been proposed aiming to identify disjoint orthology sets, where each of such sets corresponds to colorful disjoint component in the given graph [13].

Different combinatorial problem formulations, based on different objective functions, have been proposed and studied in this direction [2,13]. Here, we considered two such approaches, Minimum Colorful Components (MCC) and Maximum Edges in transitive Closure (MEC). Given a vertex-colored graph, both problems ask for the removal of some edges so that the resulting graph is partitioned in colorful components but with different objective functions. The former aims to minimize the number of connected colorful components, while the latter aims to maximize the transitive closure of the resulting graph. A related but different problem has been considered in [5], where the objective function is the minimization of edge removal, so that the computed graph consists only of colorful components.

Previous Results. Given a graph on $n$ vertices, MCC is known not only to be NPhard, but also not approximable within factor $O\left(n^{1 / 14-\varepsilon}\right)$ unless $\mathrm{P}=\mathrm{NP}$ [2]. It is easy to see that the reduction leading to this inapproximability result implies also that MCC cannot be solved in time $n^{f(k)}$ for any function $f$, where $k$ is the number of colorful components.

MEC is known to be APX-hard even when colored by at most three colors (while it is solvable in polynomial time for two colors), and, unless $\mathrm{P}=\mathrm{NP}$, it is not approximable within factor $O\left(n^{1 / 3-\varepsilon}\right)$ when the number of colors is arbitrary, even when the input graph is a tree where each color appears at most twice [1]. A heuristic to solve MEC is presented in [13], while in [1], the authors present a polynomial-time $\sqrt{2 \cdot O P T}$ approximation algorithm.

Contributions and Organization of the Paper. In this paper we investigate more deeply the complexity of MCC and MEC. More precisely, we show in Sect. 3 that MCC on trees is essentially equivalent to Minimum MultiCut on Trees, thus MCC is not approximable within factor $1.36-\varepsilon$ unless $\mathrm{P}=\mathrm{NP}$ for any $\varepsilon>0$, but 2-approximable, it is fixed-parameter tractable and it admits a poly-kernel (when the parameter is the number of colorful components). Moreover, in Sect. 4 we show that MCC is easily solvable in polynomial time on paths, while it is not in XP class when parameterized by the structural parameter Distance to Disjoint Paths.

Then we consider the parameterized complexity of MEC with respect to the number $k$ of edges in the transitive closure of a solution. For this parameter we give in Sect. 5 a parameterized algorithm, by reducing the problem to an exponential kernel. We use a similar idea in Sect. 6, to improve it to a polykernel for MEC when the input graph is a tree. Finally, we show in Sect. 7 that results similar to those of Sect.4, hold also for MEC. Due to space constraints, some proofs (marked with a $\star$ ) are deferred to the full version of the paper.

## 2 Definitions

In this section we introduce some preliminary definitions. Consider a set of colors $C=\left\{c_{1}, \ldots, c_{q}\right\}$. A $C$-colored graph $G=(V, E, C)$ is a graph where every vertex
in $V$ is associated with a color in $C$; the color associated with a vertex $v \in V$ is denoted by $c(v)$. A connected component induced by a vertex set $V^{\prime} \subseteq V$ is called a colorful component, if it does not contain two vertices having the same color. If a graph has $t$ connected components where each component $i \in[t]$ has exactly $n_{i}$ vertices, the number of edges in its transitive closure is defined by $\sum_{i=1}^{t} \frac{n_{i}\left(n_{i}-1\right)}{2}$.

Next, we introduce the formal definitions of the optimization problems we deal with.

## Minimum Colorful Components (MCC)

- Input: a $C$-colored graph $G=(V, E, C)$.
- Output: remove a set of edges $E^{\prime} \subseteq E$ such that each connected component in $G^{\prime}=\left(V, E \backslash E^{\prime}, C\right)$ is colorful, and the number of connected components of $G^{\prime}$ is minimized.

Maximum Edges in transitive Closure (MEC)

- Input: a $C$-colored graph $G=(V, E, C)$.
- Output: remove a set of edges $E^{\prime} \subseteq E$ such that each connected component in $G^{\prime}=\left(V, E \backslash E^{\prime}, C\right)$ is colorful, and the number of edges in the transitive closure of $G^{\prime}$ is maximum.

The parameterized versions of MCC and MEC are defined analogously (and abusively denoted with the same names), with the addition in the input of an integer $k$, that denotes the number of connected components in $G^{\prime}$ for MCC and the number of edges in the transitive closure of $G^{\prime}$ for MEC.

Notice that, when considering an instance of MCC and MEC, we assume that $E$ contains no edge $\{u, v\}$ with $c(u)=c(v)$, otherwise such an edge can be deleted from $E$ as $u$ and $v$ will not be part of the same colorful component in any feasible solution of MCC or MEC.

Complexity. A parameterized problem $(I, k)$ is said fixed-parameter tractable (or in the class FPT) with respect to a parameter $k$ if it can be solved in $f(k) \cdot|I|^{c}$ time (in fpt-time), where $f$ is any computable function and $c$ is a constant (see [9] for more details about fixed-parameter tractability). The class XP contains problems solvable in time $|I|^{f(k)}$, where $f$ is an unrestricted function.

A powerful technique to design parameterized algorithms is kernelization. In short, kernelization is a polynomial-time self-reduction algorithm that takes an instance $(I, k)$ of a parameterized problem $P$ as input and computes an equivalent instance ( $I^{\prime}, k^{\prime}$ ) of $P$ such that $\left|I^{\prime}\right| \leqslant h(k)$ for some computable function $h$ and $k^{\prime} \leqslant k$. The instance ( $I^{\prime}, k^{\prime}$ ) is called a kernel in this case. If the function $h$ is polynomial, we say that $\left(I^{\prime}, k^{\prime}\right)$ is a polynomial kernel.

Concerning approximation definitions, we refer the reader to some reference textbook like [3].

## 3 MCC for Trees: Parameterized Complexity and Approximability

In this section, we show that MCC on trees is essentially equivalent to the Minimum Multi-CUT problem on Trees (M-CUT-T), thus the positive and negative results of (M-CUT-T) for parameterized complexity and approximability transfer to MCC. We recall that M-CUT-T, given a tree $T_{M}$ and a set $S_{M}$ of pairs of terminals, asks if there exist a minimum cut (that is a set of removed edges) such that, for each pair $(x, y) \in S_{M}, x$ and $y$ are disconnected through that cut.

### 3.1 Positive Results

We show that MCC on trees admits an FPT algorithm (and a polynomial kernel) and a 2-approximation algorithm by reducing MCC to M-CUT-T. We first describe the reduction. Given a colored tree $G_{T}=(V, E, C)$ as an instance of MCC, we define an instance ( $T_{M}, S_{M}$ ) of M-CUT-T as follows: $T_{M}$ is exactly $G_{T}$ (except for the colors of the vertices); for each pair $(x, y)$ of vertices in $G_{T}$ such that $c(x)=c(y)$, we define a pair $(x, y)$ in $S_{M}$.

Now, we prove the main lemma of this section.
Lemma 1. Consider an instance $G_{T}$ of $M C C$ and the corresponding instance $\left(T_{M}, S_{M}\right)$ of M-CUT-T. Then: (1) given a solution of $M C C$ on $G_{T}$ consisting of $k+1$ connected components, a solution of $M-C U T-T$ on $\left(T_{M}, S_{M}\right)$ consisting of $k$ edges cut can be computed in polynomial time; (2) given a solution of $M-C U T-T$ on $\left(T_{M}, S_{M}\right)$ consisting of $k$ edges, a solution of MCC on $G_{T}$ consisting of $k+1$ connected components can be computed in polynomial time.

Proof. Consider a solution of MCC consisting of $k+1$ components obtained by removing a set $E^{\prime}$ of $k$ edges. Then, $E^{\prime}$ is a solution of M-CUT-T over instance $\left(T_{M}, S_{M}\right)$. Indeed, for each pair $(x, y) \in S_{M}, c(x)=c(y)$, hence the two vertices belong to different connected components after the removal of edges in $E^{\prime}$.

Conversely, consider a solution $E^{\prime}$ of M-CUT-T over instance ( $T_{M}, S_{M}$ ), with $\left|E^{\prime}\right|=k$. Then, remove the edges in $E^{\prime}$ from $G_{T}$ and consider the $k+1$ connected components induced by this removal in $G_{T}$. Since each pair $(x, y) \in S_{M}$ is disconnected after the removal of $E^{\prime}$, it follows that each connected component of $G_{T}$ after the removal of $E^{\prime}$ is colorful.

We can now easily give the main result of this section:
Theorem 2 ( $\star$ ). If the input graph of $M C C$ is a tree, $M C C$ can be solved in time $O^{*}\left(1.554^{k}\right)^{1}$ where $k$ is the natural parameter and also admits a 2-approximation algorithm.

Lemma 1 implies also a poly-kernel for MCC on trees.
Theorem 3 ( $\star$ ). If the input graph of $M C C$ is a tree, it is possible to compute in polynomial time a kernel of size $O\left(k^{3}\right)$ where $k$ is the natural parameter.

[^43]
### 3.2 Approximation Lower Bound of MCC on Trees

Let us now prove a lower bound for the approximation of MCC on trees, by giving a reduction from M-CUT-T. Starting from an instance ( $T_{M}, S_{M}$ ) of M-CUT-T, we compute a colored tree $G_{T}=(V, E, C)$, input of MCC, as follows. First, $G_{T}$ is isomorphic to $T_{M}$, and we color each vertex $v$ of $G_{T}$ as $c_{v}$. Denote by $E_{1}$ the edge set of such a tree. Then for each pair $(u, v) \in S_{M}$, we define a leaf $u_{v}$ adjacent to $v$ and colored $c_{u, v}$ and a leaf $v_{u}$ adjacent to $u$ and colored $c_{u, v}$ (see Fig. 1). Denote by $E_{2}$ the edge set introduced by adding these edges.

$T_{M}$

$G_{T}$

Fig. 1. Sample construction of $G_{T}$ from $T_{M}$ with $S_{M}=\{(2,8),(5,6),(4,5)\}$. Edge set $E_{2}$ of $T$ is drawn thick. For ease, colors of $G_{T}$ are drawn inside the nodes. On possible solution for this instance of M-CUT-T cuts edges $\{\{2,6\},\{1,4\}\}$ and implies 3 colorful connected components in the corresponding instance of MCC.

Lemma 4. Given a solution of $M C C$ on $G_{T}=(V, E, C)$ consisting of $k$ colorful components, we can compute in polynomial time a solution of MCC on $G_{T}=$ $(V, E, C)$ consisting of at most $k$ colorful components such that the edges cut belong only to set $E_{1}$.

Proof. Consider the case that an edge $\{u, v\}$ has been deleted, where $v$ is a leaf introduced in $G_{T}$. Then, notice that the removal of edge $\{u, v\}$ makes $v$ an isolated vertex. By construction $u$ and $v$ (and each leaf adjacent to $u$ ) have different colors. Hence there are two possible cases: either the colorful component $H$ that contains $u$ does not include vertices colored by $c_{v}$, hence we can add $v$ to $H$, thus we can avoid removing edge $\{u, v\}$, or there is a vertex $w$ colored by $c_{v}$ in $H$. In this case we can remove an edge of $E_{1}$, which separates $w$ from $u$ without removing edge $\{u, v\}$; such an edge must exist, since $v$ and $w$ are leaves incident in different internal vertices.

Lemma $5(\star)$. Consider an instance $\left(T_{M}, S_{M}\right)$ of $M-C U T-T$ and the corresponding instance $G_{T}=(V, E, C)$ of $M C C$. Then: (1) given a solution of M-CUT-T over instance $\left(T_{M}, S_{M}\right)$ that cuts $k$ edges, we can compute in polynomial time a solution of MCC over instance $G_{T}=(V, E, C)$ consisting of at most $k+1$ colorful components; (2) given a solution of MCC over instance $G_{T}=(V, E, C)$ consisting of at most $k+1$ colorful components, we can compute in polynomial time a solution of M-CUT-T over instance $\left(T_{M}, S_{M}\right)$ that cuts at most $k$ edges.

Since M-CUT-T cannot be approximated within factor 1.36 (since it is as hard as Minimum Vertex Cover to approximate [10]), Lemmas 4 and 5 allow to extend the result to MCC.

Theorem $6(\star) . M C C$ on trees cannot be approximated within factor $1.36-\varepsilon$, for any constant $\varepsilon>0$ unless $P=N P$.

## 4 Structural Parameterization of MCC

Since the MCC problem is already hard on trees, we consider in this section the complexity of MCC when the input graph is a path or is close to a set of disjoint paths. We show that MCC can be easily solved in polynomial time, while, as a sharp contrast, MCC is not in the class XP for parameter distance to disjoint paths.

Theorem 7. MCC on paths can be solved in $O\left(n^{3}\right)$-time.
Proof (Sketch). Assume that the input graph is a path $G_{P}=(V, E, C)$, and assume that the vertices on the path are ordered from $v_{1}$ to $v_{n}$. Define $M[j]$ as the minimum number of colorful components of a solution of MCC over instance $G_{P}$ restricted to vertices $\left\{v_{1}, \ldots, v_{j}\right\} . M[j]$, with $j>1$, can be computed as follows:
$M[j]=\min _{0 \leqslant t<j} M[t]+1$, such that $v_{t+1}, \ldots, v_{j}$ induce a colorful component.
In the base cases, it holds $M[1]=1$, and $M[0]=0$. Next, we prove the correctness of the dynamic programming recurrence.

We claim that given a path $G_{P}=(V, E, C)$ instance of MCC, there exists a solution of MCC on instance $G_{P}$ restricted to vertices $\left\{v_{1}, \ldots, v_{j}\right\}$ consisting of $h$ colorful components if and only if $M[j]=h$.

It is then easy to see that the value of an optimal solution of MCC on path $G_{P}=(V, E, C)$ is stored in $M[n]$. Since the table $M[j]$ consists of $n$ entries and each entry can be computed in time $O\left(n^{2}\right)$, it follows that MCC on paths can be computed in time $O\left(n^{3}\right)$.

Let us now prove that MCC is not in XP when parameterized by the Distance to Disjoint Paths number $d$ (the minimum number of vertices to remove from the input graph to have disjoint paths), even when the input graph is a tree. We prove this result by giving a reduction from Minimum Vertex Cover (MinVC) to MCC on trees.

Consider an instance $G=(V, E)$ of MinVC, and let $G_{C}=\left(V_{C}, E_{C}\right)$ be the corresponding instance of MCC. $G_{C}$ is a rooted tree, defined as follows. First, we define $|V|$ paths, one for each vertex in $G$. Path $P_{i}$ contains vertex $v_{c, i}$, colored by $c_{i}$, and vertices $e_{c, i, j}$, for each $\left\{v_{i}, v_{j}\right\} \in E$, colored by $c_{i j}$. Notice that vertices $e_{c, i, j}$ appears in $P_{i}$ based on the lexicographic order of the corresponding edges. Moreover, there exist two vertices associated with edge $\left\{v_{i}, v_{j}\right\} \in E$, namely


Fig. 2. Sample construction of an instance of MCC from an instance of MinVC. A possible solution for MinVC is given in thick while edges to be cut for the instance of MCC are also in thick.
$e_{c, i, j}$ (in $P_{i}$ ) and $e_{c, j, i}\left(\right.$ in $\left.P_{j}\right)$, which are both colored by $c_{i j}$. The tree $G_{C}$ is obtained by connecting the paths $P_{1}, \ldots P_{|V|}$ to a root $r$, which is colored by $c_{r}$, where $c_{r}$ is a fresh new color (see Fig. 2).

Lemma 8 (*). Let $G=(V, E)$ be an instance of MinVC, and let $G_{C}=$ $\left(V_{C}, E_{C}\right)$ be the corresponding instance of $M C C$. Then: (1) given a vertex cover of $G$ of size $k$, we can compute in polynomial time a solution of $M C C$ over instance $G_{C}$ consisting of $k+1$ colorful components; (2) given a solution of $M C C$ over instance $G_{C}$ consisting of $k+1$ colorful components, we can compute in polynomial time a vertex cover of $G$ of size $k$.

By the previous lemma, the following result holds.
Theorem $9(\star) . M C C$ is $N P$-hard even when the input graph is at distance 1 to Disjoint Paths.

It is worth noticing that this result extends to parameter pathwidth or distance to interval graph, as these last parameters are "stronger" than distance to disjoint path in the sense of [11].

## 5 An FPT Algorithm for MEC Parameterized by $\boldsymbol{k}$

We present a parameterized algorithm for MEC with respect to the natural parameter $k$. To do so, we will show that the problem admits an exponential size kernel.

Given a colored graph $G$, we first compute a Depth-First-Search (DFS) $D=$ $\left(V, E_{D}, E_{B}\right)$ of $G$. Recall that $D$ consists of a tree induced by $D^{\prime}=\left(V, E_{D}\right)$ (hence not considering edges in $E_{B}$ ), while $E_{B}=E \backslash E_{D}$ are called backward edges and have the following well-known property (see [6] for details).

Lemma 10. Consider a graph $G$ and the corresponding $D F S D=\left(V, E_{D}, E_{B}\right)$. Let $\{u, v\}$ be a backward edge. Then there exists a path $p$ in $D^{\prime}=\left(V, E_{D}\right)$ that starts in the root of $D^{\prime}$ and contains both $u$ and $v$.

We will first show some easy cases where there is a solution of MEC of size at least $k$. Let $V_{A}$ be the set of vertices of $V$ which are parent of a leaf in $D^{\prime}$.

Lemma 11 ( $\star$ ). If there exists a path in $D^{\prime}$ from the root $r\left(D^{\prime}\right)$ to a leaf of $D^{\prime}$ of length at least $2 k$, then there exists a solution of MEC of size at least $k$.

Lemma 12 ( $\star$ ). There exists a solution of MEC of size at least $k$ if $\left|V_{A}\right| \geqslant k$.
Now, for each vertex $v \in V_{A}$ we consider the leaves adjacent to $v$ and their colors. Define the set $C_{x}(v)$ as the set of leaves colored by $c_{x}$ and adjacent to $v \in V_{A}$ in $D$. Then the following property holds.
Lemma 13 ( $\star$ ). Given a vertex $v \in V_{A}$, if there exist $\sqrt{2 k}$ non-empty sets $C_{x}(v)$ associated with distinct colors $c_{x}$, then there exists a solution of MEC of size at least $k$.

Given a vertex $v \in V_{A}$ and a set $C_{x}(v)$, consider the sets of vertices connected with backward edges to a vertex $u \in C_{x}(v)$. Define $\operatorname{Adj}\left(C_{x}(v)\right)=\left\{V_{A}^{\prime} \subseteq V_{A}\right.$ : $\left.\{u, w\} \in E, u \in V_{A}^{\prime}, w \in C_{x}(v)\right\}$.

The following property holds.
Lemma $14(\star)$. Given a vertex-colored graph $G$ such that the hypothesis of Lemma 11 does not hold, consider a vertex $v$ in $V_{A}$ and a set $C_{x}(v)$. of possible sets of adjacent vertices to a node $u$ of $C_{x}(v)$.

Based on Lemma 14, we can partition the vertices of each $C_{x}(v)$ into sets $C_{x, 1}(v), \ldots, C_{x, p}(v)$, with $p \leqslant 2^{2 k+1}$, depending on their set of adjacent vertices (that is two vertices of $C_{x}(v)$ belong to the same set $C_{x, t}(v)$ if they have the same set of adjacent vertices).

Now, assume that the hypotheses of Lemmas 11-13 do not hold. Consider an algorithm that, for each set $C_{x, i}(v)$, computes a set $C_{x, i}^{\prime}(v)$ by picking at most $k$ vertices of $C_{x, i}(v)$ and removing the other vertices of $C_{x, i}(v)$. Let $G^{\prime}$ be the resulting graph. We claim that $G^{\prime}$ contains at most $O\left(k^{2} 2^{2 k+1}\right)$ vertices. First, notice that each $C_{x, t}^{\prime}(v)$ contains at most $k$ vertices and that, for each vertex $v$, there exists at most $2^{2 k+1}$ sets $C_{x, t}^{\prime}(v)$. Since, there exist at most $O(k \sqrt{k})$ sets $C_{x}(v)$ (at most $\sqrt{2 k}$ colors $c_{x}$ and at most $k$ vertices $v \in V_{A}$ ), we can conclude that $G^{\prime}$ contains at most $O\left(k^{2} \sqrt{k} 2^{2 k+1}\right)$ vertices in sets $C_{x, i}^{\prime}(v)$.

Now, consider the vertices $G^{\prime}$ which are not contained in some set $C_{x, i}^{\prime}(v)$. These vertices correspond to internal vertices of $D^{\prime}$. Since the hypothesis of Lemma 11 does not hold, $D^{\prime}$ is a tree of depth at most $2 k$, and there exist at most $k$ vertices adjacent to leaves, as $\left|V_{A}\right|<k$. Hence there exist at most $k$ paths of length $2 k$ in $D^{\prime}$ from the root to vertices adjacent to leaves, thus we can conclude that there exist at most $2 k^{2}$ internal vertices in $D^{\prime}$. Hence there exists at most $2 k^{2}$ vertices in $G^{\prime}$ which are not contained in some set $C_{x, i}^{\prime}(v)$.

Now, we prove that $\left(G^{\prime}, k\right)$ is a kernel for MEC.
Lemma $15(\star)$. There exists a collection of disjoint colorful components $V_{1}, \ldots, V_{h}$ of size at least 2 in $G$ if and only if there exists a collection of disjoint colorful components $V_{1}^{\prime}, \ldots, V_{h}^{\prime}$ in $G^{\prime}$, with $\left|V_{i}\right|=\left|V_{i}^{\prime}\right|, 1 \leqslant i \leqslant h$.

Hence we have the following result.
Theorem $16(\star)$. There exists a kernel of size $O\left(k^{2} \sqrt{k} 2^{2 k+1}\right)$ for MEC.

## 6 A Poly-Kernel for MEC on Trees

In this section, we show that in the special case of MEC where the input graph is a tree, the kernel size can be quadratic. The algorithm is similar to the one of Sect. 5. Consider a colored tree $G_{T}=(V, E, C)$, and let $r\left(G_{T}\right)$ denote the root of $G_{T}$. Lemmata 11-13 hold for $G_{T}$. Hence, we focus only on the leaves of $G_{T}$.

Since $G_{T}$ is a tree, it follows that a leaf $u$ having ancestor $v$ belongs to a component of size at least 2 only if $u$ and $v$ belongs to the same component. It follows that among the leaves having color $c_{x}$ and adjacent to a vertex $u$, only one can belong to a colorful component of size at least 2 . Hence, given $v \in V_{A}$, let $C_{x}(v)$ be the set of leaves adjacent to $v$ and colored by $c_{x}$. We remove all but one vertex from $C_{x}(v)$. Let $G_{T}^{\prime}$ be the resulting tree. We have the following property for $G_{T}^{\prime}$.

Lemma 17 ( $\boldsymbol{\star}$ ). There exists a collection of disjoint colorful components $V_{1}, \ldots, V_{h}$ of size at least 2 in $G_{T}$ if and only if there exists a collection of disjoint colorful components $V_{1}^{\prime}, \ldots, V_{h}^{\prime}$ in $G_{T}^{\prime}$, with $\left|V_{i}\right|=\left|V_{i}^{\prime}\right|, 1 \leqslant i \leqslant h$.

Theorem 18. There exists a kernel of size $O\left(k^{2}\right)$ for MEC on trees.
Proof. The result follows from Lemma 17 and from the fact that tree $G_{T}^{\prime}$ contains at most $k^{2}$ internal vertices (by Lemma 11 and by Lemma 12) and there exist at most $O(k \sqrt{k})$ sets $C_{x, i}(v)$ (by Lemma 13), each of size bounded by 1 .

## $7 \quad$ Structural Parameterization of MEC

It is easy to see that the results on structural parameterization for MCC hold also for MEC (after appropriate modifications).

Theorem $19(\star)$. MEC on paths can be solved in $O\left(n^{3}\right)$-time.
Similarly to MCC, MEC is NP-hard even if we restrict the instance to graphs having distance 1 to Disjoint Paths. As for MCC, it is worth noticing that this hardness result extends to other stronger parameters like pathwidth [11].

Theorem 20 ( $\star$ ). MEC is NP-hard even when the input graph has distance 1 to Disjoint Paths.

## 8 Conclusion

In the future, we aim at refining the parameterized complexity analysis, for example deepen the structural results for MCC and MEC. Moreover, it would be interesting to study the parameterized complexity of the two problems under other meaningful parameters in the direction of parameterizing above a guaranteed value.

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# A Candidate for the Generalised Real Line 

Lorenzo Galeotti ${ }^{(\boxtimes)}$<br>Universität Hamburg, Hamburg, Germany<br>lorenzo.galeotti@gmail.com


#### Abstract

Let $\kappa$ be an uncountable cardinal with $\kappa^{<\kappa}=\kappa$. In this paper we introduce $\mathbb{R}_{\kappa}$, a Cauchy-complete real closed field of cardinality $2^{\kappa}$. We will prove that $\mathbb{R}_{\kappa}$ shares many features with $\mathbb{R}$ which have a key role in real analysis and computable analysis. In particular, we will prove that the Intermediate Value Theorem holds for a non-trivial subclass of continuous functions over $\mathbb{R}_{\kappa}$. We propose $\mathbb{R}_{\kappa}$ as a candidate for extending computable analysis to generalised Baire spaces.


## 1 Introduction

Computable analysis is the study of the computational properties of real analysis. We refer the reader to [21] for an introduction to computable analysis. In classical computability theory one studies the computational properties of functions over natural numbers and transfers these properties to arbitrary countable spaces via coding. The same approach is taken in computable analysis. By using coding, in fact, one can transfer topological and computational results from the Baire space $\omega^{\omega}$ to sets of cardinality $2^{\aleph_{0}}$. In particular, by encoding the real numbers, one can use the Baire space to study computability in the context of real analysis.

Of particular interest in computable analysis is the study of the computational content of theorems from classical analysis. The idea is that of formalizing the complexity of theorems by means similar to those used in computability theory to classify functions over the natural numbers. In this context, the Weihrauch theory of reducibility plays an important role. For an introduction to the theory of Weihrauch reductions, see [4]. Weihrauch reductions can be used to classify functions over the Baire space $\omega^{\omega}$. By using this concept it is possible to arrange many theorems from classical real analysis in a complexity hierarchy called the Weihrauch hierarchy. A study of the Weihrauch degrees of some of the most important theorems from real analysis can be found in [3,4].

Recently, the study of the descriptive set theory of the generalised Baire spaces $\kappa^{\kappa}$ for cardinals $\kappa>\omega$ has been catching the interest of set theorists (see [12] for an overview on the subject). This fact is also witnessed by the increasing
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number of workshops dedicated to generalised Baire spaces organized in the last two years (AST 2014 in Amsterdam and a satellite workshop to DMV 2015 in Hamburg). Even though generalised Baire spaces are not a new concept in set theory, many aspects of this theory are still unknown. In particular there has been no attempt to generalise computable analysis to spaces of cardinality $2^{\kappa}$.

This paper provides the foundational basis for the study of generalised computable analysis, namely the generalisation of computable analysis to generalised Baire spaces. Since in classical computable analysis and classical Weihrauch theory the field of real numbers has a central role, a question arises naturally in this context: what is the right generalisation of $\mathbb{R}$ in the context of generalised computable analysis?


In this paper we answer this question. In particular, we will introduce $\mathbb{R}_{\kappa}$, a generalisation of the real line, which provides a well behaved environment for generalising classical results from real analysis to uncountable cardinals. We propose $\mathbb{R}_{\kappa}$ as the starting point for the study of generalised computable analysis.

The problem of generalising the real line is not new in mathematics. Different approaches have been tried for very different purposes. A good introduction to these number systems can be found in [9]. Among the most influential contributions to this field particularly important are the works of Sikorski [20] and Klaua [15] on the real ordinal numbers and that of Conway [6] on the surreal numbers. Sikorski's idea was to repeat the classical Dedekind construction of the real numbers starting from an ordinal equipped with the Hessenberg operations (i.e., commutative operations over the ordinal numbers). Unfortunately, one can prove that these fields do not have the density properties that, as we will see, will have a central role in the context of real analysis. The surreal numbers were introduced by Conway in order to generalise both the Dedekind construction of real numbers and the Cantor construction of ordinal numbers. In his introduction to surreal numbers, Conway proved that they form a (class) real closed field. Later, Ehrlich [15] proved that every real closed field is isomorphic to a subfield of the surreal numbers, showing that they behave like a universal (class) model for real closed fields. It is then natural for us to use this framework in the developing of $\mathbb{R}_{\kappa}$.

As we will see, doing analysis over field extensions of $\mathbb{R}$ is not an easy task. This is due to the fact that no proper ordered field extension of $\mathbb{R}$ is connected. However, many of the basic theorems of real analysis are linked to the fact that $\mathbb{R}$ is a connected space. To overcome this problem, instead of using standard topological tools, we will use a different mathematical framework which, under specific conditions over the density of $\mathbb{R}_{\kappa}$, will allow us to see our field extension of $\mathbb{R}$ as a linear continuum.

In this paper, we shall give necessary requirements for a space $\mathbb{R}_{\kappa}$ to be the generalisation of the real numbers. Considering the Intermediate Value Theorem (IVT) as one of the pillars of real analysis, we place particular emphasis on its validity in the generalised case, and we develop the requirements in such a way that they will allow us to prove it.

## 2 The Surreal Numbers

In this paper $\kappa$ will refer to a fixed cardinal larger than $\omega$. As usual in generalised descriptive set theory, let $\kappa$ be an uncountable cardinal with $\kappa^{<\kappa}=\kappa$. Note in particular that this assumption implies that $\kappa$ is a regular cardinal.

We will assume basic knowledge of topology, field theory and computable analysis. A good introduction to these subjects can be found in [5, 18] and [21], respectively.

The following definition as well as most of the results in this section, are due to Conway [6] and have also been deeply studied by Gonshor in [14].

Definition 1 (Surreal numbers). A surreal number is a function from an ordinal $\alpha \in$ On to $\{+,-\}$, i.e., a sequence of pluses and minuses of ordinal length. We will denote the class of surreal numbers by No. The length of a surreal number $x \in$ No is the smallest ordinal $\ell(x) \in$ On for which $x$ is not defined. Moreover, we will use $\mathrm{No}_{<\alpha}$ and $\mathrm{No}_{\leq \alpha}$ to denote the set of $x \in$ No with $\ell(x)<\alpha$ and the set of $x \in$ No with $\ell(x) \leq \alpha$ respectively.

We can define a total order over No as follows:
Definition 2. Let $x, y \in$ No be two surreal numbers. We say that $x$ is smaller than $y$ in symbol $x<y$ iff $x(\alpha)<y(\alpha)$, where $\alpha$ is the smallest ordinal s.t. $x(\alpha) \neq y(\alpha)$ using the order $-<0<+$ where $x(\alpha)=0$ if $x$ is not defined at $\alpha$.

According to Conway's original idea, every surreal number is generated by filling some cut between shorter numbers. The following theorem gives us a connection between this intuition and the surreal numbers as we have defined them.

Theorem 3 (Simplicity theorem). Let $L$ and $R$ be two sets of surreal numbers such that $L<R$ (i.e. $\forall \ell \in L \forall r \in R \ell<r$ ). Then there is a unique surreal $z$, denoted by $[L \mid R]$, of minimal length such that $L<\{z\}<R$. We will call $[L \mid R]$ $a$ representation of $z$.

By using the Simplicity Theorem Conway defined the field operations over No and proved that these operations satisfy the axioms of real closed fields. Later Ehrlich proved that the class field No behaves like a universal model for the theory of real closed fields, this means that every set-like model of the theory of real closed fields is isomorphic to a subfield of No. In particular Conway proved that the real numbers are a subfield of $\mathrm{No}_{\leq \omega}$.

The last theorem we want to mention in this section is due to van der Dries and Ehrlich [8]:

Theorem 4 (van der Dries and Ehrlich). The set of surreal numbers $\mathrm{No}_{<\kappa}$ is a real closed field.

## 3 Super Dense $\kappa$-real Extensions of $\mathbb{R}$

In this paper we will have a quasi-axiomatic approach. In particular, we will first determine the properties that we need on $\mathbb{R}_{\kappa}$ in order to prove some basic facts from classical analysis. Then we will show how it is possible to define $\mathbb{R}_{\kappa}$ as a subfield of the surreal numbers.

Let us consider some of the basic properties that we expect from $\mathbb{R}_{\kappa}$. First of all we want $\mathbb{R}_{\kappa}$ to be a generalisation of $\mathbb{R}$ to the uncountable cardinal $\kappa$. Therefore we require that $\mathbb{R}_{\kappa}$ is a proper ordered field extension of $\mathbb{R}$. As we said, we want to use $\mathbb{R}_{\kappa}$ to do analysis. For this reason, we expect $\mathbb{R}_{\kappa}$ to behave as much as possible like $\mathbb{R}$. Formally we will require that $\mathbb{R}_{\kappa}$ is a real closed field extending $\mathbb{R}$. Since the theory of real closed fields is complete [17, Corollary 3.3.16], this implies that $\mathbb{R}_{\kappa}$ has the same first order properties as $\mathbb{R} .^{1}$

## REQUIREMENT R1: $\mathbb{R}_{\kappa}$ is a real closed field extending $\mathbb{R}$.

Now, since we want to use $\mathbb{R}_{\kappa}$ to do computable analysis over sets of cardinality $2^{\kappa}$, we require that $\left|\mathbb{R}_{\kappa}\right|=2^{\kappa}$.

## REQUIREMENT R2: $\mathbb{R}_{\kappa}$ has cardinality $2^{\kappa}$.

Finally, since the set of rational numbers $\mathbb{Q}$ has a central role in the representation theory of $\mathbb{R}$ (the interested reader is referred to [21]), we want $\mathbb{R}_{\kappa}$ to have a dense subset which can play the same role as $\mathbb{Q}$.

$$
\text { REQUIREMENT R3: } \mathbb{R}_{\kappa} \text { has a dense subset of cardinality } \kappa \text {. }
$$

In general we define:
Definition 5 ( $\kappa$-real extension of $\mathbb{R}$ ). Let $K$ be a field satisfying R1, R2, R3. Then we will call $K$ a $\kappa$-real extension of $\mathbb{R}$.

Now we focus on those properties that are needed to extend theorems from classical analysis to $\mathbb{R}_{\kappa}$. Many of these classical results depend on the order over $\mathbb{R}$ and on its interval topology. So we will start considering interval topologies over $\kappa$-real extensions of $\mathbb{R}$ and their properties. First we recall few facts from field theory and classical analysis. It is a well known result from classical analysis that $\mathbb{R}$ has no Dedekind complete ordered field extensions (see [5, Corollary 8.7.4]). Therefore $\mathbb{R}_{\kappa}$ will not be Dedekind complete. More generally we have:

[^44]Corollary 6. Let $K$ be a $\kappa$-real extension of $\mathbb{R}$. Then $K$ is not Dedekind complete.
As usual, given an ordered field $K$, one can define Cauchy sequences over $K$.
Definition 7 (Cauchy sequences). Let $\alpha$ be an ordinal and $K^{+}$the positive part of $K$. Then a sequence $\left(x_{i}\right)_{i \in \alpha}$ of elements of $K$ is Cauchy iff

$$
\forall \varepsilon \in K^{+} \exists \beta<\alpha \forall \gamma, \gamma^{\prime} \geq \beta\left|x_{\gamma^{\prime}}-x_{\gamma}\right|<\varepsilon .
$$

The sequence is convergent if there is $x \in K$ such that

$$
\forall \varepsilon \in K^{+} \exists \beta<\alpha \forall \gamma \geq \beta\left|x_{\gamma}-x\right|<\varepsilon
$$

We will call $x$ the limit of $\left(x_{i}\right)_{i \in \alpha}$. Given an ordered field $K$ it is said to be Cauchy complete iff every Cauchy sequence whose length is equal to the degree $\operatorname{Deg}(K)=\operatorname{Coi}\left(K^{+}\right)$of $K$ has a limit in $K$.

It is an easy exercise to see that Cauchy and Dedekind completeness coincide on Archimedean fields, while on non-Archimedean fields Cauchy completeness is weaker than Dedekind completeness.

Another property which is central in mathematical analysis is connectedness. It turns out that connectedness and Dedekind completeness are equivalent properties. Therefore, it is easy to see that we will not be able to define $\mathbb{R}_{\kappa}$ in such a way that its interval topology is connected.

As we said, our main purpose is that of proving basic facts from analysis over $\mathbb{R}_{\kappa}$. In particular we want to be able to prove the Intermediate Value Theorem (IVT). It turns out that if the IVT holds on an ordered field $K$ then $K$ is connected. This means that we cannot aim to prove the IVT over $\kappa$-real extensions of $\mathbb{R}$ in all its strength.

## $3.1 \kappa$-topologies

Given what we have proved in the previous section, it is hard to do analysis over $\kappa$-real extensions of $\mathbb{R}$ by using standard topological tools. To overcome this problem we will use a tool introduced by Alling called $\kappa$-topology. A similar approach to do analysis over the surreal numbers was taken in [19].

Definition 8 ( $\kappa$-topology). A $\kappa$-topology $\tau$ over a set $X$ is a collection of subsets of $X$ such that:

1. $\emptyset, X \in \tau$.
2. $\forall \alpha<\kappa$ if $\left\{A_{i}\right\}_{i \in \alpha}$ is a collection of sets in $\tau$, then $\bigcup_{i<\alpha} A_{i} \in \tau$.
3. $\forall A, B \in \tau A \cap B \in \tau$.

The elements of $\tau$ are called $\kappa$-open sets.
Intuitively, the reason why we use $\kappa$-topologies is that, as we have seen in the previous section, interval topologies over $\kappa$-real extensions of $\mathbb{R}$ are too fine. As we will see $\kappa$-topologies will be coarser than topologies and will allow us to prove a weaker version of the Intermediate Value Theorem over particularly well-behaved $\kappa$-real extensions of $\mathbb{R}$.

Theorem 9 (Alling). Let $X$ be a set and $B$ be a topological base over $X$. Then the set $\tau_{\kappa}$ defined as follows: $\emptyset, X \in \tau_{\kappa}$ and union of less than $\kappa$ elements of $B$ is in $\tau_{\kappa}$, is a $\kappa$-topology. We will call $\tau_{\kappa}$ the $\kappa$-topology generated by B. Moreover we will call $B$ a base for the $\kappa$-topology.

Obviously many topological definitions can be relativized to $\kappa$-topologies. In particular we have the following:
Definition 10 ( $\kappa$-continuity). Let $X$ and $Y$ be two sets and $\tau$, $\tau^{\prime}$ be two $\kappa$-topologies respectively on $X$ and on $Y$. Then $f: X \rightarrow Y$ is a $\kappa$-continuous function iff $\forall U \in \tau^{\prime} f^{-1}[U] \in \tau$.
Definition 11 ( $\kappa$-connectedness). Let $X$ be a set and $\tau$ be a $\kappa$-topology over $X$. Then $X$ is $\kappa$-connected iff $\forall U, V \in \tau X=U \cup V \wedge U \cap V=\emptyset \Rightarrow U=\emptyset \vee V=\emptyset$.

All these definitions behave quite well with respect to their topological counterparts. Indeed, many classical theorems from topology hold for $\kappa$-topologies (see [1]). However, there are theorems from topology that are not valid on $\kappa$-topologies. Typically for $\kappa$-topologies local properties do not transfer to global properties (e.g. in $\kappa$-topologies openness is not implied by local openness).

Now we will introduce a $\kappa$-topological analogue of the interval topology over an ordered set.

Definition 12 (Interval $\kappa$-topology). Let $X$ be an ordered set and $B$ be the set of open intervals with end points in $X \cup\{+\infty,-\infty\}$. We will call interval $\kappa$-topology over $X$ the $\kappa$-topology generated by $B$.

From now on we will consider the interval $\kappa$-topology as the standard $\kappa$-topology over $\kappa$-real extensions of $\mathbb{R}$.

As we have seen, in order to be able to prove some basic theorems from analysis we need to work within a connected space. However, as we have already pointed out, we can not aim for connectedness of $\kappa$-real extensions of $\mathbb{R}$. The next result is due to Alling [1] and it makes precise the connection between the density of an ordered set and the connectedness of its interval $\kappa$-topology.
Definition 13 (Hausdorff $\eta_{\kappa}$-set). Let $X$ be an ordered set and $\kappa$ be a cardinal. We say that $X$ is an $\eta_{\kappa}$-set iff given $L, R \subseteq X$, such that $L<R$ and $|L|+|R|<\kappa$ then there is $x \in X$ such that $L<\{x\}<R$.
Theorem 14 (Alling). Let $X$ be an $\eta_{\kappa}$-set endowed with the interval $\kappa$-topology and $X^{\prime}$ a subset of $X$. Then $X^{\prime}$ is $\kappa$-connected iff $X^{\prime}$ is an interval in $X$.

In view of Theorem 14, it is natural to require:

## REQUIREMENT R4: $\mathbb{R}_{\kappa}$ is an $\eta_{\kappa}$-set.

Definition 15. A field $K \supseteq \mathbb{R}$ is called $a$ super dense $\kappa$-real extension of $\mathbb{R}$ if it satisfies requirements R1, R2, R3, and R4.

As in classical topology $\kappa$-continuous functions preserve $\kappa$-connectedness.
Theorem 16. Let $f: X \rightarrow Y$ be a $\kappa$-continuous function. If $X$ is $\kappa$-connected then $f(X)$ is $\kappa$-connected.

### 3.2 Analysis over Super Dense $\kappa$-real Extensions of $\mathbb{R}$

Using the results from the previous section we can modify the standard topological proof of the IVT to show that its restriction to $\kappa$-continuous functions holds over super dense $\kappa$-real extensions of $\mathbb{R}$.

Theorem $17\left(\mathrm{IVT}_{\kappa}^{K}\right)$. Let $K$ be a super dense $\kappa$-real extension of $\mathbb{R}$, the set $[a, b] \subset K$ be a closed subinterval of $K$ and $f:[a, b] \rightarrow K$ be a $\kappa$-continuous function. Then for every $r \in K$ such that $r$ is in between $f(a)$ and $f(b)$, there is $c \in[a, b]$ such that $f(c)=r$.

It is a well-known fact that in every real closed field the IVT holds for polynomials in one variable (see [17, Theorem 3.3.9]), therefore it is natural to ask if polynomials over super dense $\kappa$-real extensions of $\mathbb{R}$ are $\kappa$-continuous.

Theorem 18. Let $K$ be a super dense $\kappa$-real extension of $\mathbb{R}$ and $p$ be a polynomial in one variable with coefficients in $K$. Then $p$ is $\kappa$-continuous.

## 4 The Generalised Real Line $\mathbb{R}_{\kappa}$

We are now ready to define $\mathbb{R}_{\kappa}$. A naïve attempt to define such extension would be that of starting from $\kappa$ endowed with the surreal operations (i.e., the Hessenberg operations) and try to repeat the standard construction of $\mathbb{Z}^{\kappa}$ and $\mathbb{Q}^{\kappa}$. Then, we could define $\mathbb{R}^{\kappa}$ as the Cauchy completion of $\mathbb{Q}^{\kappa}$ obtaining a Cauchy complete field. Unfortunately this approach does not work. This is due to the fact that, as Sikorski proved, the field $\mathbb{Q}^{\kappa}$ is Cauchy complete and then $\mathbb{R}^{\kappa}=\mathbb{Q}^{\kappa}$. Recall that $\mathbb{Q}^{\kappa}$ is a set of equivalence classes of pairs of elements in $\mathbb{Z}^{\kappa}$, hence it has cardinality at most $\kappa$. Therefore $\mathbb{R}^{\kappa}$ violates $R 2$ and is not a good candidate for our purposes. This construction appeared for the first time in a paper from Sikorski in 1948 [20] (see also [2,15] for a complete study of this approach). For this reason we have to take a different approach in defining $\mathbb{R}_{\kappa}$.

We will follow the work done by Dales and Woodin in [7] for $\kappa=\aleph_{1}{ }^{2}$. By Theorem 4 we know that $\mathrm{No}_{<\kappa}$ is a real closed field. Moreover, since $\kappa>\omega$, it is easy to see that $\mathbb{R} \subset \mathrm{No}_{<\kappa}$. In particular this means that R 1 holds for $\mathrm{No}_{<\kappa}$. Furthermore, it is not hard to prove that $\mathrm{No}_{<\kappa}$ also satisfies R 4 . Then we have:

Proposition 19. The field $\mathrm{No}_{<\kappa}$ has the following properties:

1. $\left|\mathrm{No}_{<\kappa}\right|=\kappa$ and $\operatorname{Deg}\left(\mathrm{No}_{<\kappa}\right)=\kappa$.
2. $\operatorname{Cof}\left(\mathrm{No}_{<\kappa}\right)=\operatorname{Coi}\left(\mathrm{No}_{<\kappa}\right)=\kappa$ and $\mathrm{No}_{<\kappa}$ has weight $\kappa$.

Proposition 19 tells us that $\mathrm{No}_{<\kappa}$ has almost all the properties that we want from $\mathbb{R}_{\kappa}$ but is still too small. Moreover, it is not hard to see that $\mathrm{No}_{<\kappa}$ is not Cauchy complete in the sense of Definition 7 this fact is particularly problematic

[^45]in the context of computable analysis, where most of the classical representations of $\mathbb{R}$ rely on the fact that $\mathbb{R}$ is the Cauchy completion of $\mathbb{Q}$.

It is therefore natural to consider $\mathrm{No}_{<\kappa}$ as generalised rational numbers, and to define $\mathbb{R}_{\kappa}$ as the Cauchy completion of $\mathrm{No}_{<\kappa}$ as in classical analysis ${ }^{3}$. Since we are working within the surreal numbers, this can be done in a natural way.

Definition 20 (Veronese cuts). Let $K$ be an ordered field. We call $\langle L, R\rangle$ a cut over $K$ iff $L, R \subseteq K$ and $L<R$. Moreover we will say that $\langle L, R\rangle$ is a Veronese cut iff it is a cut such that, $L$ has no maximum, $R$ has no minimum and for each $\varepsilon \in K^{+}$there are $\ell \in L$ and $r \in R$ such that $r<\ell+\varepsilon$. We will say that $K$ is Veronese complete iff for each Veronese cut $\langle L, R\rangle$, there is $x \in K$ such that $L<\{x\}<R$.

It is a well known fact that Cauchy and Veronese completeness are equivalent notions (see $[7,10]$ ). For this reason we can define the Cauchy completion of $\mathrm{No}_{<\kappa}$ by using the Simplicity Theorem as follows:

Definition $21\left(\mathbb{R}_{\kappa}\right)$. We define $\mathbb{R}_{\kappa}$ as follows:

$$
\mathbb{R}_{\kappa}=\mathrm{No}_{<\kappa} \cup\left\{x \mid x=[L \mid R] \text { where }\langle L, R\rangle \text { is a Veronese cut on } \mathrm{No}_{<\kappa}\right\} .
$$

Now we will show that $\mathbb{R}_{\kappa}$ is a super dense $\kappa$-real extension of $\mathbb{R}$. First of all we have that $\mathrm{No}_{<\kappa}$ is a dense subfield of $\mathbb{R}_{\kappa}$ and that $\mathbb{R}_{\kappa}$ is Cauchy complete (i.e., $\mathbb{R}_{\kappa}$ the Cauchy completion of $\mathrm{No}_{<\kappa}$ ).

Lemma 22. The field $\mathrm{No}_{<\kappa}$ is dense in $\mathbb{R}_{\kappa}$. Moreover the set $\mathbb{R}_{\kappa}$ is Cauchy complete.

In view of the previous lemma from now on we will call $\mathrm{No}_{<\kappa}$ the $\kappa$-rational numbers and we will use the symbol $\mathbb{Q}_{\kappa}$ instead of $\mathrm{No}_{<\kappa}$.

Since we have showed that $\mathbb{R}_{\kappa}$ is the Cauchy completion of a real closed field, by a standard model theoretical argument we have:

Corollary 23. The set $\mathbb{R}_{\kappa}$ is a real closed field extending $\mathbb{R}$.
Now that we have shown that $\mathbb{R}_{\kappa}$ is a real closed field extending $\mathbb{R}$ we want to check that all the other properties of super dense $\kappa$-real extensions of $\mathbb{R}$ hold for $\mathbb{R}_{\kappa}$.

Theorem 24. The field $\mathbb{R}_{\kappa}$ is the unique Cauchy complete real closed field of cardinality $2^{\kappa}$, with degree and weight $\kappa$ which is an $\eta_{\kappa}$-set.

Proof. We will only prove $\left|\mathbb{R}_{\kappa}\right|=2^{\kappa}$ the rest follows from the fact that $\mathbb{Q}_{\kappa}$ is dense in $\mathbb{R}_{\kappa}$. We want to prove $2^{\kappa} \leq\left|\mathbb{R}_{\kappa}\right| \leq 2^{\kappa}$. On the one hand we have that $\mathbb{R}_{\kappa} \subset \mathrm{No}_{\leq \kappa}$. Indeed, $\mathrm{No}_{\leq \kappa}$ contains the Dedekind completion of $\mathrm{No}_{<\kappa}$, hence also its Cauchy completion $\mathbb{R}_{\kappa}$. Then, since $\left|\mathrm{No}_{\leq_{\kappa}}\right|=2^{\kappa}$, we have that $\left|\mathbb{R}_{\kappa}\right| \leq 2^{\kappa}$.

[^46]On the other hand let $\{0,1\}^{<\kappa}$ be the full binary tree of height $\kappa$, we define a tree $T$ which is in bijection with $\{0,1\}^{<\kappa}$ and whose nodes are subintervals of $\mathbb{R}_{\kappa}$ and whose branches corresponds to different elements of $\mathbb{R}_{\kappa}$. We define the tree by recursion as follows: set $T_{\lambda}=(0,1)$ as the root of the tree. Now assume that for $p \in 2^{<\kappa}$ and that the element $T_{p} \neq \emptyset$ is already defined. We define $T_{p 0}$ and $T_{p 1}$ as two non-empty disjoint subintervals of $T_{p}$ such that $T_{p 0}=\left(a_{p 0}, b_{p 0}\right)$ and $T_{p 1}=\left(a_{p 1}, b_{p 1}\right)$, where $a_{p 0}, b_{p 0}, a_{p 1}, b_{p 1} \in \mathbb{Q}_{\kappa}$ with $\left|a_{p 0}-b_{p 0}\right| \leq \frac{1}{\ell(p)+1}$ and $\left|a_{p 1}-b_{p 1}\right| \leq \frac{1}{\ell(p)+1}$, where $\ell(p)$ is the length of $p$. Finally if $p \in 2^{<\kappa}$ is of limit length $\gamma$ and $T_{p \upharpoonright \alpha}$ has already been defined for every $\alpha<\gamma$, we define $T_{p}^{\prime}=\bigcap_{\alpha<\gamma} T_{p \upharpoonright \alpha}$. Note that by the fact that $\mathbb{R}_{\kappa}$ is an $\eta_{\kappa}$-set, the set $T_{p}^{\prime}$ nonempty, moreover $T_{p}^{\prime}$ is trivially convex (i.e., if $x, y \in T_{p}^{\prime}$ and $x \leq z \leq y$, then $z \in T_{p}^{\prime}$ ). Therefore we can define $T_{p}$ as we have done for the successor stage starting from $T_{p}^{\prime}$. It follows trivially by the way in which we have defined the tree that for $p \in 2^{\kappa}$ the set $\bigcap_{\alpha \in \kappa} T_{p \upharpoonright \alpha}$ contains a single element of $\mathbb{R}_{\kappa}$. Indeed by the properties of the tree we have that $\left[\left\{a_{p \upharpoonright \alpha} \mid \alpha \in \kappa\right\} \mid\left\{b_{p \upharpoonright \alpha} \mid \alpha \in \kappa\right\}\right]$ is a Veronese cut in $\mathbb{Q}_{\kappa}$. Therefore $\bigcap_{\alpha \in \kappa} T_{p \upharpoonright \alpha}$ is a singleton in $\mathbb{R}_{\kappa}$ as desired. Moreover given $p, p^{\prime} \in 2^{\kappa}$ such that $p \neq p^{\prime}$ we have $\bigcap_{\alpha \in \kappa} T_{p \upharpoonright \alpha} \neq \bigcap_{\alpha \in \kappa} T_{p^{\prime} \upharpoonright \alpha}$. Therefore we trivially have $2^{\kappa} \leq\left|\mathbb{R}_{\kappa}\right|$ as desired.

## 5 Conclusions and Future Work

In this paper we have introduced a real closed field extending $\mathbb{R}$ suitable for doing real analysis over the generalised Baire space $\kappa^{\kappa}$. We have showed that, although it has some limitations intrinsic to the problem, $\mathbb{R}_{\kappa}$ preserves many interesting properties of the real numbers. In particular we showed:

1. $\mathbb{R}_{\kappa}$ is a Cauchy complete super dense $\kappa$-real extension of $\mathbb{R}$ of cardinality $2^{\kappa}$.
2. $\mathbb{R}_{\kappa}$ has a dense subset of cardinality $\kappa$ and $\operatorname{Coi}\left(\mathbb{R}_{\kappa}^{+}\right)=\kappa$.
3. The IVT holds for $\kappa$-continuous functions.

As we have seen, most of these properties are motivated by computable analysis. For this reason we propose $\mathbb{R}_{\kappa}$ as the generalised real line in the context of computable analysis. An example of how $\mathbb{R}_{\kappa}$ can be used to study the topological Weihrauch complexity of theorems of analysis can be found in [13].

There are two natural continuations of this paper. On one hand it is natural to ask for a study of the computational strength of generalisations of theorems from real analysis. To accomplish this, a theory of generalised type two computability is needed. As we shall show in a paper soon to appear, it is possible to modify the notions of Ordinal Turing Machine introduced by Koepke in [16] to define a generalised version of Type Two Turing Machine (T2TM). The intuition behind this notion is that generalised T2TMs should run classical programs for Turing machines for $\kappa$ steps instead of just $\omega$. These machines lead to a very natural notion of computability, in which, because of the properties of $\mathbb{R}_{\kappa}$ and $\mathbb{Q}_{\kappa}$, the field operations restricted to $\mathbb{Q}_{\kappa}$ are computable in less than $\kappa$ steps, while one may need to run forever (i.e., up to $\kappa$ ) to compute the same operations over $\mathbb{R}_{\kappa}$.

Moreover, this notion of computability preserves the correspondence between continuous functions and functions which are computable with an oracle.

A second natural continuation of this paper is the systematic study of the real analysis of $\mathbb{R}_{\kappa}$. Particularly interesting would be the study of a notion of integral. This problem is not new in the theory of surreal numbers and partial solutions have been proposed in the last decades (see [11, pp. 2-3]). Recently a solution to the problem of integration over the surreal numbers has been proposed in [11]. We are currently working on the problem of integration over $\mathbb{R}_{\kappa}$.

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# Finitely Generated Semiautomatic Groups 

Sanjay Jain ${ }^{1}$, Bakhadyr Khoussainov ${ }^{2}$, and Frank Stephan ${ }^{1,3(\boxtimes)}$<br>${ }^{1}$ Department of Computer Science, National University of Singapore, 13 Computing Drive, COM1, Singapore 117417, Republic of Singapore sanjay@comp.nus.edu.sg<br>${ }^{2}$ Department of Computer Science, University of Auckland, Private Bag 92019, Auckland, New Zealand<br>bmk@cs.auckland.ac.nz<br>${ }^{3}$ Department of Mathematics, National University of Singapore, 10 Lower Kent Ridge Road, S17, Singapore 119076, Republic of Singapore fstephan@comp.nus.edu.sg


#### Abstract

The present work shows that Cayley automatic groups are semiautomatic and exhibits some further constructions of semiautomatic groups and in particular shows that every finitely generated group of nilpotency class 3 is semiautomatic.


## 1 Introduction

Hodgson [4,5] as well as Khoussainov and Nerode [8] initiated the study of automatic structures, including that of groups. In their approach, such a group is given by a regular set $A$ as the domain (denoting the representatives of the group) such that both, the group operation $\circ$ and the equality $=$, are automatic; that is, there is an automaton which reads the convoluted tuples $(x, y, z)$ or $(x, y)$ and decides whether such a tuple satisfies $x \circ y=z$ or $x=y$, respectively. Here, for $x=x_{0} x_{1} \ldots x_{m}$ and $y=y_{0} y_{1} \ldots y_{n}$ with $x_{i}, y_{i} \in \Sigma$, the convolution of the pair $(x, y)$ is the string $z_{0} z_{1} \ldots z_{\max \{m, n\}}$, over the new alphabet $(\Sigma \cup\{\#\})^{2}$, where $z_{i}=\left(x_{i}, y_{i}\right)$ and $x_{i}$ (respectively $\left.y_{i}\right)$ is taken to be $\#$ in case of $i>$ $m$ (respectively, $i>n$ ). The convolution over triples or tuples in general is defined similarly. The advantage of this setting is that every function and relation definable in the language of group theory using parameters from the group is again automatic. Furthermore, automata providing the mappings can be found algorithmically. This also leads to the conclusion that for every fixed automatic group, the first-order theory is decidable [8]. Furthermore, automatic functions are precisely those which can be computed in linear time by a position-faithful one-tape Turing machine [2], thus the automatic functions coincide with the smallest reasonable time complexity class for functions.
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Epstein et al. [3] argued that in the above formalisation, automaticity is, at least from the viewpoint of finitely generated groups, too restrictive. They furthermore wanted that the representatives of the group elements are given as words over the generators, leading to more meaningful representatives than arbitrary strings. Their concept of automatic groups led, for finitely generated groups, to a larger class of groups, though, by definition, it of course does not include groups which require infinitely many generators; groups with infinitely many generators, to some extent, were covered in the notion of automaticity by Hodgson, Khoussainov and Nerode. Nies and Thomas [10,11] provide results which contrast and compare these two notions of automaticity and give an overview on results for groups which are automatic in the sense of Hodgson, Khoussainov and Nerode.

Kharlampovich et al. [7] generalised the notion further to Cayley automatic groups. Here a finitely generated group $(A, \circ)$ is Cayley automatic iff the domain $A$ is a regular set, for every group element there is a unique representative in $A$ and, for every $a \in A$, the mapping $x \mapsto x \circ a$ is automatic. Note that the above requires multiplication by constants to be automatic only from one side; when multiplication by a constant from both sides are automatic, then the group is called Cayley biautomatic.

Finitely generated Cayley automatic groups have word problem decidable in quadratic time, carrying over the corresponding result from the two previous versions of automaticity. As opposed to the case of automatic groups (in the original sense of Hodgson), Miasnikov and Šunić [9] showed that several natural problems like the conjugacy problem can be undecidable for some Cayley automatic groups.

Jain et al. [6] investigated the general approach where, in a structure for some relations and functions, it is only required that the versions of the functions or relations with all but one variable fixed to constants is automatic. Here the convention is to put the automatic domains, functions and relations before a semicolon and the semiautomatic relations after the semicolon. For a group, the semiautomatic group $(A, \circ ;=)$ would be a structure where the domain $A$ is regular, the group operation (with both inputs) is automatic and for each fixed element $a \in A$ the set $\{b \in A: b=a\}$ is regular - note that group elements might have several representatives in semiautomatic groups.

In the present work, for any group, $\varepsilon$ represents the neutral element. One of the basic results obtained is that the notion $(A, \circ ;=)$ collapses to an automatic group (in the sense of Hodgson, Khoussainov and Nerode), as

$$
a=b \Leftrightarrow \exists c[a \circ c=\varepsilon \text { and } b \circ c=\varepsilon] .
$$

For semiautomatic groups, the two interesting group structures are $(A,=; \circ)$ and $(A ; \circ,=)$. In the first one, the equality is automatic, while in the second one, both the group operation and the equality are only semiautomatic. If a group is finitely generated, then the definition of being Cayley biautomatic is the same as having a presentation of the form $(A,=; \circ)$.

Finitely generated semiautomatic groups share with the other types of automatic groups one important property: The word problem can be decided in
quadratic time and the algorithm is the same as known for the Cayley automatic groups [7]. Thus finitely generated groups with an undecidable or very complex word problem are not semiautomatic.

## 2 Constructions of Semiautomatic Groups

Recall that Cayley automatic groups are finitely generated group $(A, \circ)$ iff the domain $A$ is regular, every group element has a unique representative in $A$ and for every $a \in A$, the mapping $x \mapsto x \circ a$ is automatic. This notion is equivalent to allowing multiple representatives in $A$ for the group elements, but additionally requiring that equality is automatic.

Miasnikov and Šunić [9] showed that there are Cayley automatic groups which are not Cayley biautomatic, that is, which have no semiautomatic representation of the form $(A,=; \circ)$; furthermore, there are Cayley automatic groups for which the conjugacy problem is undecidable and that the isomorphism problem is also undecidable for the class of Cayley automatic groups. Berdinsky and Khoussainov [1] have shown that every Baumslag Solitar group is Cayley automatic and Jain et al. [6] announced that every Baumslag Solitar group is semiautomatic. This and other results follow now from a general transfer theorem which shows that every Cayley automatic group is semiautomatic.

Theorem 1. If $(A,=;\{x \mapsto x \circ a: a \in A\})$ is a Cayley automatic group, then the group has a semiautomatic presentation $\left(B, x \mapsto x^{-1} ; \circ,=\right)$; in this presentation the domain is regular and the inversion is an automatic function, whereas the equality and the group operation are semiautomatic.

Proof. Given the Cayley automatic group $A$ as in the statement of the theorem, let $B=\{(v, w): v, w \in A\}$ be the set of convoluted pairs $(v, w)$, where the pair $(v, w)$ stands for the element $v^{-1} \circ w$ of $A$. Now $(v, w) \circ(\varepsilon, u)=v^{-1} \circ w \circ \varepsilon^{-1} \circ u=$ $v^{-1} \circ w \circ u=(v, w \circ u),(u, \varepsilon) \circ(v, w)=u^{-1} \circ \varepsilon \circ v^{-1} \circ w=(v \circ u)^{-1} \circ w=(v \circ u, w)$ and $(v, w) \circ(w, v)=v^{-1} \circ w \circ w^{-1} \circ v=\varepsilon$. As every fixed element $a \in A$ can be represented by either $(\varepsilon, a)$ or $\left(a^{-1}, \varepsilon\right)$, multiplication with a fixed group element from either side is automatic. Furthermore, the mapping $(v, w) \mapsto(w, v)$ is automatic and computes the inverse. The set of representations of a fixed element $a \in A$ is the set $\{(v, w):(a, \varepsilon) \circ(v, w)=(\varepsilon, \varepsilon)\}=\{(v, w): v \circ a=w\}$, where the latter set is easily seen to be regular.

The above result shows that the undecidability results for Cayley automatic groups by Miasnikov et al. [9,12] generalise to finitely generated semiautomatic groups.

Corollary 2. There is a semiautomatic group $(A ; \circ,=)$ for which the conjugacy problem is undecidable. Furthermore, the isomorphism problem for semiautomatic groups is undecidable.

The next result shows that all semiautomatic groups can have an automatic inversion.

Proposition 3. Every semiautomatic group $(A ; \circ,=)$ has a further semiautomatic presentation $\left(B, x \mapsto x^{-1} ; \circ,=\right)$.

Proof. The proposition is proven by introducing two new symbols, + and - , such that $B=\{+,-\} \cdot A$ consists of all $+x$ representing $x \in A$ and $-x$ representing $x^{-1}$ for $x \in A$. The inverse is now computed by the function interchanging + and - . For fixed $a \in A, x \mapsto x \circ a$ becomes $+x \mapsto+(x \circ a)$ and $-x \mapsto-\left(a^{-1} \circ x\right)$, $x \mapsto a \circ x$ is implemented similarly and $\{+x: x=a\} \cup\left\{-x: x=a^{-1}\right\}$ is the regular set of representatives of $a$ in $B$.

Theorem 4. Assume that $(A, \circ,=)$ is an automatic group (in the sense of Hodgson, Khoussainov and Nerode), ( $B ; \circ,=$ ) is a semiautomatic group and $\left\{\varphi_{b}: b \in B\right\}$ is a family of homomorphisms from $A$ to $A$ such that for each $b, b^{\prime} \in B, \varphi_{b \circ b^{\prime}}(a)=\varphi_{b}\left(\varphi_{b^{\prime}}(a)\right)$ and each $\varphi_{b}$ is an automatic mapping, then the semidirect product $A \rtimes_{\varphi} B$ is also a semiautomatic group. Here the group operation in $A \rtimes_{\varphi} B$ is defined by $(a, b) \circ\left(a^{\prime}, b^{\prime}\right)=\left(a \circ \varphi_{b}\left(a^{\prime}\right), b \circ b^{\prime}\right)$.

Proof. Consider the representation set $C=\{(a, b, \tilde{a}): a, \tilde{a} \in A$ and $b \in B\}$, where $(a, b, \tilde{a}) \in C$ stands for $(a, \varepsilon) \circ(\varepsilon, b) \circ(\tilde{a}, \varepsilon)$ in the group $A \rtimes_{\varphi} B$. Now, for a fixed $a^{\prime} \in A, b^{\prime} \in B$ and arbitrary $(a, b, \tilde{a}) \in C$, the multiplications are defined as follows:

$$
\begin{aligned}
& (a, b, \tilde{a}) \circ\left(a^{\prime}, \varepsilon\right) \mapsto\left(a, b, \tilde{a} \circ a^{\prime}\right) ; \\
& \left(a^{\prime}, \varepsilon\right) \circ(a, b, \tilde{a}) \mapsto\left(a^{\prime} \circ a, b, \tilde{a}\right) ; \\
& (a, b, \tilde{a}) \circ\left(\varepsilon, b^{\prime}\right) \mapsto\left(a, b \circ b^{\prime}, \varphi_{b^{\prime-1}}(\tilde{a})\right) ; \\
& \left(\varepsilon, b^{\prime}\right) \circ(a, b, \tilde{a}) \mapsto\left(\varphi_{b^{\prime}}(a), b^{\prime} \circ b, \tilde{a}\right) .
\end{aligned}
$$

Note that multiplying with $\left(a^{\prime}, b^{\prime}, \tilde{a}^{\prime}\right)$ in $C$ can be defined using the above as $\left(a^{\prime}, b^{\prime}, \tilde{a}^{\prime}\right)=\left(a^{\prime}, \varepsilon\right) \circ\left(\varepsilon, b^{\prime}\right) \circ\left(\tilde{a}^{\prime}, \varepsilon\right)$. Now, all the four mappings above are automatic as they only use homomorphisms from $B$, which are automatic, and multiplication with fixed group elements in the basic groups $A$ and $B$, which are automatic. It follows that $\circ$ is semiautomatic in $C$.

For equality, note that $(a, b, \tilde{a})=\left(a^{\prime}, b^{\prime}, \tilde{a}^{\prime}\right)$ iff $b=b^{\prime}$ (in group $B$ ) and $a \circ \varphi_{b}(\tilde{a})=a^{\prime} \circ \varphi_{b^{\prime}}\left(\tilde{a}^{\prime}\right)$ (in group $A$ ). Thus, for a fixed $\left(a^{\prime}, b^{\prime}, \tilde{a}^{\prime}\right) \in C$ and any $(a, b, \tilde{a}) \in C,(a, b, \tilde{a})=\left(a^{\prime}, b^{\prime}, \tilde{a}^{\prime}\right)$ iff $b=b^{\prime}($ in group $B)$ and $a \circ \varphi_{b^{\prime}}(\tilde{a})=$ $a^{\prime} \circ \varphi_{b^{\prime}}\left(\tilde{a}^{\prime}\right)$. As $\varphi_{b^{\prime}}, \circ$ restricted to $A$ and equality in $A$ are automatic, it follows that equality is semiautomatic in $C$.

It can also be shown that the free product of finitely many semiautomatic groups is semiautomatic. The construction is very much inline with the one of Kharlampovich et al. [7] for Cayley automatic groups with some adjustments for semiautomaticity.

Theorem 5. The free product of finitely many semiautomatic groups is semiautomatic.

Proof. Let $\left(A_{1} ; \circ,=\right), \ldots,\left(A_{n} ; \circ,=\right)$ be semiautomatic groups which all share the empty word $\varepsilon$ as neutral element and use disjoint alphabets to represent the other elements. Note that, for each fixed $a \in A_{k}$, there is an automatic mapping $x \mapsto a \circ x$ (for $x \in A_{k}$ ) such that the length of $a \circ x$ is at most a constant more than the length of $x$. Let \# be a symbol not appearing in the members of $A_{1}, \ldots, A_{k}$. Now each member of the free product $B$ of $A_{1}, \ldots, A_{k}$ is a word of the form $\#^{+} u_{1} \#^{+} u_{2} \#^{+} \ldots \#^{+} u_{m} \#^{+}$with $u_{1}, \ldots, u_{m}$ representing elements different from $\varepsilon$ and no two subsequent $u_{h}, u_{h+1}$ are from the same group $A_{k}$. Any word from $\#^{+}$denotes the neutral element of the group. It is sufficient to show that the multiplication with any fixed element from $A_{1} \cup A_{2} \cup \ldots \cup A_{n}-\{\varepsilon\}$ is automatic, multiplying with $\varepsilon$ can be realised by the identity function. Consider $a \in A_{k}-\{\varepsilon\}$.

Now $x \mapsto a \circ x$ is given as follows: $x$ is mapped to $\# a \#$ in the case that $x \in \#^{+} ; x$ is mapped to $\# a \# x$ in the case that the first component $u_{1}$ from $x$ is not from $A_{k} ; x$ is mapped to the word $x^{\prime}$, where $u_{1}$ is replaced by $\#^{\left|u_{1}\right|}$ in the case that $u_{1}=a^{-1}$; otherwise $x$ is mapped to the word $x^{\prime}$, where $u_{1}$ is replaced by a word from $\left(a \circ u_{1}\right) \#^{*}$. To ensure automaticity of the mapping, in the last two cases above, enough \#'s are filled in to make sure that length of $x$ and $x^{\prime}$ do not differ by more than a constant (independent of $x$ ).

The mapping $x \mapsto x \circ a$ is given as follows: $x$ is mapped to $\# a \#$ in the case that $x \in \#^{+} ; x$ is mapped to $x \# a \#$ in the case that the last component $u_{m}$ of $x$ is not from $A_{k}$; the last part of the form $u_{m} \#^{+}$is erased from $x$ by the mapping in the case that $u_{m}=a^{-1}$ and the last part $u_{m} \#^{+}$is replaced by $\left(u_{m} \circ a\right) \#$ in the case that $u_{m} \in A_{k}-\left\{a^{-1}\right\}$.

Furthermore, for comparing whether $x$ of the form $\#^{+} u_{1} \#^{+} \ldots \#^{+} u_{n} \#^{+}$ represents a fixed element $\# a_{1} \# a_{2} \# \ldots \# a_{m} \#$, consider the automaton consisting of $m$ distinct subautomatons: the $h$-th subautomaton checks whether the component $u_{h}$ of $x$ is from the same $A_{k}$ as $a_{h}$ and has the same value; the automaton accepts iff all these checks succeed and the number of components $n$ of $x$ is exactly $m$.

## 3 Nilpotent Groups

Kharlampovich et al. [7] showed that finitely generated groups of nilpotency class 1 or 2 are Cayley automatic. The next theorem uses semiautomatic groups in place of Cayley automatic groups and pushes the above result one step further. As it is open whether all the finitely generated groups of nilpotency class 3 are Cayley automatic, this result provides a candidate for separating the two notions within the finitely generated groups.

Theorem 6. Every finitely generated group of nilpotency class 3 can be represented as a semiautomatic group $(A ; \circ,=)$.

Proof. Let $a_{1}, \ldots, a_{n}$ be the finitely many generators in the original nilpotent group.

Consider the factor group of the given group over the quotient group generated by all elements of the form $x \circ y \circ x^{-1} \circ y^{-1}$. This group is Abelian and is isomorphic to

$$
\mathbb{Z}^{r} \times\left\{0,1, \ldots, p_{r+1}-1\right\} \times \ldots \times\left\{0,1, \ldots, p_{n}-1\right\}
$$

for some $r \leqslant n$ and natural numbers $p_{r+1}, \ldots, p_{n} \geqslant 2$.
Let $b_{1}, \ldots, b_{n^{\prime}}$ be all the group elements of the form $a_{i}^{-1} \circ a_{i^{\prime}}^{-1} \circ a_{i} \circ a_{i^{\prime}}$ and let $c_{1}, \ldots, c_{n^{\prime \prime}}$ be all the group elements of the form $a_{i}^{-1} \circ b_{j}^{-1} \circ a_{i} \circ b_{j}$ or $b_{j}^{-1} \circ a_{i}^{-1} \circ b_{j} \circ a_{i}$. Note that the $c_{1}, \ldots, c_{n^{\prime \prime}}$ commute with all group elements, that for each $i, j$ there is a $k$ with

$$
a_{i} \circ b_{j}=b_{j} \circ a_{i} \circ c_{k}, a_{i} \circ b_{j}^{-1}=b_{j}^{-1} \circ a_{i} \circ c_{k}^{-1}
$$

and that for each $i, i^{\prime}$ there are $j, k$ with

$$
a_{i} \circ a_{i^{\prime}}=a_{i^{\prime}} \circ a_{i} \circ b_{j}, a_{i} \circ a_{i^{\prime}}^{-1}=a_{i^{\prime}}^{-1} \circ a_{i} \circ b_{j}^{-1} \circ c_{k} .
$$

Similar rules allow to move $a_{i}^{-1}$ over $a_{i^{\prime}}, b_{j}$. Note that the group elements $b_{j}, b_{j^{\prime}}$ also commute with each other, as when, for example, $b_{j^{\prime}}=a_{i}^{-1} \circ a_{i^{\prime}}^{-1} \circ a_{i} \circ a_{i^{\prime}}$ then $b_{j} \circ b_{j^{\prime}}=b_{j} \circ a_{i}^{-1} \circ a_{i^{\prime}}^{-1} \circ a_{i} \circ a_{i^{\prime}}=a_{i}^{-1} \circ a_{i^{\prime}}^{-1} \circ a_{i} \circ a_{i^{\prime}} \circ b_{j}=b_{j^{\prime}} \circ b_{j}$. The reason is that the $c_{k}, c_{k^{\prime}}$ produced by moving $a_{i}^{-1}, a_{i^{\prime}}^{-1}$, respectively, over $b_{j}$, are cancelled out when moving $a_{i}, a_{i^{\prime}}$ over $b_{j}$. Now, each group element is given by a convoluted tuple of integers

$$
\left(m_{1}, \ldots, m_{n}, m_{1}^{\prime}, \ldots, m_{n^{\prime}}^{\prime}, m_{1}^{\prime \prime}, \ldots, m_{n^{\prime \prime}}^{\prime \prime}\right)
$$

where, for $i=r+1, \ldots, n, m_{i} \in\left\{0,1, \ldots, p_{i}-1\right\}$. The above member of $A$ represents

$$
a_{1}^{m_{1}} \circ \ldots \circ a_{n}^{m_{n}} \circ b_{1}^{m_{1}^{\prime}} \circ \ldots \circ b_{n^{\prime}}^{m_{n^{\prime}}^{\prime}} \circ c_{1}^{m_{1}^{\prime \prime}} \circ \ldots \circ c_{n^{\prime \prime}}^{m_{n^{\prime \prime}}^{\prime \prime}} .
$$

Note that several tuples of this type may represent the same group element due to products of some $b_{j}$ and $c_{k}$ being $\varepsilon$.

In the representation set $A$, the integers $m_{i}$ and $m_{j}^{\prime}$ in the above are represented in binary, with the reverse ordering of the bits to allow automatic addition of components. Furthermore, each $m_{k}^{\prime \prime}$ is represented as a convoluted tuple ( $h_{0}, h_{1}, \ldots, h_{n}$ ) of integers (in binary using reverse ordering of the bits) satisfying

$$
m_{k}^{\prime \prime}=h_{0}+h_{1} \cdot m_{1}+\ldots+h_{n} \cdot m_{n}
$$

The number of integers used in the overall representation described above is $n+n^{\prime}+(n+1) \cdot n^{\prime \prime}$, which is a constant independent of the group element; therefore convolution can indeed be used to represent the group element.

Now it will be shown that multiplication with a fixed $a_{i}$ is automatic and that equality is semiautomatic.

First, for automaticity of the multiplication with a fixed element, note that it is sufficient to show that multiplication with a fixed generator from
$a_{1}, a_{1}^{-1}, \ldots, a_{n}, a_{n}^{-1}$ is automatic, as every other group element is a fixed product of these. This is shown in several steps, the number of steps is constant and each step is automatic. For showing that the mapping $x \mapsto a_{i} \circ x$ is automatic, it is now described how, $a_{i} \circ a_{i^{\prime}}^{m_{i^{\prime}}} a_{i^{\prime}+1}^{m_{i^{\prime}+1}} \ldots a_{n}^{m_{n}} b_{1}^{m_{1}^{\prime}} \ldots b_{n^{\prime}}^{m_{n^{\prime}}^{\prime}} c_{1}^{m_{1}^{\prime \prime}} \ldots c_{n^{\prime \prime}}^{m_{n^{\prime \prime}}^{\prime \prime}}$ is updated to $a_{i^{\prime}}^{m_{i^{\prime}}} \circ a_{i} \circ a_{i^{\prime}+1}^{m_{i^{\prime}+1}} \ldots a_{n}^{m_{n}} b_{1}^{s_{1}^{\prime}} \ldots b_{n^{\prime}}^{s_{n^{\prime}}^{\prime}} c_{1}^{s_{1}^{\prime \prime}} \ldots c_{n^{\prime \prime}}^{s_{n^{\prime \prime}}^{\prime \prime}}$, where $i^{\prime}<i$ and $m_{k}^{\prime \prime}=\left(h_{0}^{k}, h_{1}^{k}, \ldots, h_{n}^{k}\right)$. Repeatedly using this mechanism to shift $a_{i}$ over $a_{1}^{m_{1}} a_{2}^{m_{2}} \ldots a_{i-1}^{m_{i-1}}$ and then showing how to handle the increase of $m_{i}$ by 1 , gives the multiplication by $a_{i}$ for any member of the group as represented in $A$. Now suppose $1 \leqslant i^{\prime}<i, 1 \leqslant q \leqslant n$, and $m_{i^{\prime}}>0$. There are $j, k_{1}, \ldots, k_{n}$ such that $a_{i} a_{i^{\prime}}=a_{i^{\prime}} a_{i} b_{j}$ and $b_{j} a_{q}=a_{q} b_{j} c_{k_{q}}$. Then, the following operations are done to update $m_{j}^{\prime}$ and various $m_{k}^{\prime \prime}$ to obtain the corresponding $m_{j}^{\prime}$ and $s_{j}^{\prime \prime}$ (values not updated are unchanged).
(a) $m_{i^{\prime}}$ is added to $m_{j}^{\prime}$ (to handle the increase in the $b_{j}$ ).
(b) $m_{i^{\prime}}\left(m_{i^{\prime}}-1\right) / 2$ is added to $m_{k_{i^{\prime}}}^{\prime \prime}$ ( to handle the increase in $c_{k_{i^{\prime}}}$ due to moving of $b_{j}$ generated in (a) over $a_{i^{\prime}}^{m^{\prime}}$ ). If $m_{i^{\prime}}$ is odd, then the above addition can be achieved by adding $m_{i^{\prime}} / 2$ to $h_{0}^{k_{i^{\prime}}}$ and adding $\left(m_{i^{\prime}}-2\right) / 2$ to $h_{i^{\prime}}^{k_{i^{\prime}}}$. If $m_{i^{\prime}}$ is even, $m_{i^{\prime}}\left(m_{i^{\prime}}-1\right) / 2=m_{i^{\prime}}\left(m_{i^{\prime}}-2\right) / 2+m_{i^{\prime}} / 2$. Thus, the above addition can be achieved by adding $m_{i^{\prime}}\left(m_{i^{\prime}}-2\right) / 2$ to $h_{0}^{k_{i^{\prime}}}$ and adding $m_{i^{\prime}} / 2$ to $h_{i^{\prime}}^{k_{i^{\prime}}}$.
(c) $m_{i^{\prime}} * m_{q}$ is added to $m_{k_{q}}^{\prime \prime}$, for $q=i^{\prime}+1, \ldots, n$ (to handle the increase in $c_{k_{q}}$ due to moving of $b_{j}^{m_{i^{\prime}}}$ over $a_{q}^{m_{q}}$ ). This can be done by adding $m_{i^{\prime}}$ to $h_{q}^{k_{q}}$.

Note that $a_{i} \circ a_{i^{\prime}}^{-1}=a_{i^{\prime}}^{-1} \circ a_{i} \circ b_{j}^{-1} \circ c_{k^{\prime}}$ for some $k^{\prime}$ which permits to handle the case when $m_{i^{\prime}}<0$ in a similar manner. For the multiplication

$$
a_{i} \circ a_{i}^{m_{i}} a_{i+1}^{m_{i+1}} \ldots a_{n}^{m_{n}} b_{1}^{m_{1}^{\prime}} \ldots b_{n^{\prime}}^{m_{n^{\prime}}^{\prime}} c_{1}^{m_{1}^{\prime \prime}} \ldots c_{n^{\prime \prime}}^{m_{n^{\prime \prime}}^{\prime \prime}}
$$

one updates $m_{i}$ to $m_{i}+1$ and, as a chain reaction, for $k=1, \ldots, n^{\prime \prime}$, update $h_{0}^{k}$ to $h_{0}^{k}-h_{i}^{k}$, for the tuple $\left(h_{0}^{k}, h_{1}^{k}, \ldots, h_{n}^{k}\right)$ representing $m_{k}^{\prime \prime}$ (so that the new value of $m_{i}$ is used rather than the older value in the computation of $m_{k}^{\prime \prime}$ ).

Similarly it can be shown that also the mappings $x \mapsto a_{i}^{-1} \circ x, x \mapsto x \circ a_{i}$ and $x \mapsto x \circ a_{i}^{-1}$ are automatic.

The above handles all multiplications by $a_{i}$ on the left except for the case of $i>r$ and $m_{i}+1=p_{i}$. To handle this, an additional multiplication by $a_{i}^{-p_{i}}$ can be done using the above mechanism to bring the power of $a_{i}$ to 0 .

Now, for showing semiautomaticity of equality, for any fixed element

$$
a_{1}^{m_{1}} \circ \ldots \circ a_{n}^{m_{n}} \circ b_{1}^{m_{1}^{\prime}} \circ \ldots \circ b_{n^{\prime}}^{m_{n^{\prime}}^{\prime}} \circ c_{1}^{m_{1}^{\prime \prime}} \circ \ldots \circ c_{n^{\prime \prime}}^{m_{n^{\prime \prime}}^{\prime \prime}} \in A
$$

it is shown that the set of its representatives in $A$ is regular. Note that in the vector of the exponents, for each further representative of the group element, the first $n$ coordinates must also be $m_{1}, m_{2}, \ldots, m_{n}$, which can be easily checked. In the case that these $n$ coordinates are equal, one can tailormake an automaton to check for equality. The automaton can, for each $k$ and for the coordinates $\left(h_{0}, h_{1}, \ldots, h_{n}\right)$ representing $m_{k}^{\prime \prime}$, use the formula

$$
h_{0}+h_{1} \cdot m_{1}+\ldots+h_{n} \cdot m_{n}
$$

to get the explicit value corresponding to $m_{k}^{\prime \prime}$ in the other representative, in binary notation, as multiplication of integers by constants can be done automatically. However, the $m^{\prime}$-coordinates and $m^{\prime \prime}$-coordinates can be different for the two representatives. The difference of these coordinates must, however, give a vector representing $\varepsilon$. Thus, it suffices to give a test for neutrality in the $m^{\prime}$ coordinates and $m^{\prime \prime}$-coordinates in order to be able to decide equality to the fixed given element. Call a set $\left\{v_{1}, \ldots, v_{r}\right\}$ of vectors representing these coordinates to be independent over $\mathbb{Z}$ iff no $v_{h}$ can be obtained from a linear combination of the others using coefficients from $\mathbb{Z}$. If one of the sets $\left\{v_{1}, v_{2}, \ldots, v_{r}\right\}$, $\left\{-v_{1}, v_{2}, \ldots, v_{r}\right\},\left\{v_{1}-v_{2}, v_{2}, \ldots, v_{r}\right\},\left\{v_{2}-v_{1}, v_{2}, \ldots, v_{r}\right\}$ is independent, then all of them are. So Euclid's algorithm can be run on the vectors until all but one vector have a 0 in the first coordinate; then one can run the algorithm until, among all those vectors with a 0 in the first coordinate, all but at most one have a 0 in the second coordinate and so on. This implies that the number of independent vectors is not larger than the number of coordinates. Thus there are fixed vectors $\left\{v_{1}, \ldots, v_{\ell}\right\}$ such that two vectors

$$
\begin{aligned}
& \left(m_{1}, \ldots, m_{n}, m_{1}^{\prime}, \ldots, m_{n^{\prime}}^{\prime}, m_{1}^{\prime \prime}, \ldots, m_{n^{\prime \prime}}^{\prime \prime}\right) \text { and } \\
& \left(m_{1}, \ldots, m_{n}, \tilde{m}_{1}^{\prime}, \ldots, \tilde{m}_{n^{\prime}}^{\prime}, \tilde{m}_{1}^{\prime \prime}, \ldots, \tilde{m}_{n^{\prime \prime}}^{\prime \prime}\right)
\end{aligned}
$$

represent the same element iff the difference

$$
\left(0, \ldots, 0, m_{1}^{\prime}-\tilde{m}_{1}^{\prime}, \ldots, m_{n^{\prime}}^{\prime}-\tilde{m}_{n^{\prime}}^{\prime}, m_{1}^{\prime \prime}-\tilde{m}_{1}^{\prime \prime}, \ldots, m_{n^{\prime \prime}}^{\prime \prime}-\tilde{m}_{n^{\prime \prime}}^{\prime \prime}\right)
$$

is of the form $r_{1} \cdot v_{1}+r_{2} \cdot v_{2}+\ldots r_{\ell} \cdot v_{\ell}$ for some $r_{1}, \ldots, r_{\ell} \in \mathbb{Z}$. This is an existentially quantified formula, where the multiplication with fixed vectors (represented as convoluted tuples) can be done by an automatic function and their adding and comparing with the target as well. Thus the predicate whether the two vectors from above representing the two group elements are the same is automatic. Thus for each group element

$$
a_{1}^{m_{1}} \circ \ldots \circ a_{n}^{m_{n}} \circ b_{1}^{m_{1}^{\prime}} \circ \ldots \circ b_{n^{\prime}}^{m_{n^{\prime}}^{\prime}} \circ c_{1}^{m_{1}^{\prime \prime}} \circ \ldots \circ c_{n^{\prime \prime}}^{m_{n^{\prime \prime}}^{\prime \prime}}
$$

there is a finite automaton which decides whether another group element is equal to it. So the group $(A ; \circ,=)$ is semiautomatic.

For the representation used in the above theorem, by using the natural subgroup $B$ of all elements in $A$ generated by the $b_{j}$ and $c_{k}$, the following Theorem 7 below can be shown. The key idea is to represent each group element in the form $b \circ a \circ \tilde{b}$ where $b, \tilde{b}$ are in $B$ and $a$ is either $a_{1}^{m_{1}} \circ \ldots \circ a_{n}^{m_{n}}$ or $a_{n}^{m_{n}} \circ \ldots \circ a_{1}^{m_{1}}$; these two orderings are used in order to facilitate inversion. Item (b) in the theorem below is proven by coding a computationally difficult problem into the theory of the structure of the group and then conclude that this problem would be solvable in the case that the given structure is semiautomatic.

Theorem 7. In the following, $(A, \circ)$ denotes a finitely generated group of nilpotency class $3, B$ denotes the commutator subgroup generated by all elements of the form $x \circ y \circ x^{-1} \circ y^{-1}$ with $x, y \in A$ and $\bullet$ denotes the restriction of $\circ$ to the domain $(A \times B) \cup(B \times A)$.
(a) For every $A$ as above, the structure $\left(A, B, x \mapsto x^{-1}, \bullet ; \circ,=\right)$ is semiautomatic.
(b) For some $A$ as above, the structure $(A, B, \bullet,=; \circ)$ is not semiautomatic.

Proof. (a): The notation from the proof of Theorem 6 is carried over for this proof. The group elements are represented as products $b \circ a \circ \tilde{b}$ where (i) $b, \tilde{b}$ are products of $b_{j}$ 's and $c_{k}$ 's represented in the same format as they are represented in Theorem 6 and (ii) $a$ is a member of ( $a_{1}^{*} a_{2}^{*} \ldots a_{n}^{*} \cup a_{n}^{*} \ldots a_{2}^{*} a_{1}^{*}$ ) represented as a convoluted tuple $\left(m_{0}, m_{1}, \ldots, m_{n}\right)$, where $m_{0} \in\{-1,1\}$; the tuple $\left(m_{0}, m_{1}, \ldots, m_{n}\right)$ represents $a_{1}^{m_{1}} \circ \ldots \circ a_{n}^{m_{n}}$, if $m_{0}=1$, and $a_{n}^{m_{n}} \circ \ldots \circ a_{1}^{m_{1}}$, if $m_{0}=-1$.

The mappings $b \mapsto b^{-1}, a \mapsto a^{-1}$ and $\tilde{b} \mapsto \tilde{b}^{-1}$ are realised by negating all entries in the corresponding tuples; for mapping $(b \circ a \circ \tilde{b})$ to $(b \circ a \circ \tilde{b})^{-1}$, one has to exchange the entries of $b$ and $\tilde{b}$ as well, as $(b \circ a \circ \tilde{b})^{-1}=\tilde{b}^{-1} \circ a^{-1} \circ b^{-1}$. Thus the mapping $x \mapsto x^{-1}$ (in the chosen representation) is automatic.

Note that, in the representation for $\hat{b} \in B$, all the $m$-coordinates are 0 . Thus for the component $m_{k}^{\prime \prime}$ in the representation of $\hat{b}$, the integers $h_{1}, h_{2}, \ldots$ (as in Theorem 6) can be ignored. Hence, the multiplication $\hat{b} \bullet(b \circ a \circ \tilde{b})$, can be done by adding $m_{j}^{\prime}$ coordinate of the representation of $\hat{b}$ to the corresponding $m_{j}^{\prime}$ coordinate of $b$ and the $h_{0}$-component of the $m_{k}^{\prime \prime}$ coordinate of $\hat{b}$ to the corresponding $h_{0}$-component of the $m_{k}^{\prime \prime}$ coordinates of $b$. Similarly, when computing $(b \circ a \circ \tilde{b}) \bullet \hat{b}$, the coordinates of $\hat{b}$ are added as above to those of $\tilde{b}$.

Note that, in the representation chosen for this proof, $\hat{b} \in B$ is actually a product of the form: $b^{\prime} \circ \varepsilon \circ \tilde{b}^{\prime}$, where $b^{\prime}, \tilde{b}^{\prime}$ are represented as in Theorem 6. The coordinates of $b^{\prime}$ and $\tilde{b}^{\prime}$ as above can be contracted to the coordinates of one member of $B$ by simply adding, component-wise, prior to carrying out the multiplication $\bullet$ as described above. These arguments explain why $\bullet$ is an automatic function.

Multiplication of an element $x$ with generators $a_{i}$ from either side as done in Theorem 6 , can easily be adjusted to the representation chosen here.

Now, assume a fixed group element $x=a_{1}^{m_{1}} a_{2}^{m_{2}} \ldots a_{n}^{m_{n}} \circ b^{\prime}$, where $b^{\prime} \in B$, is given. Let $a=a_{1}^{m_{1}} a_{2}^{m_{2}} \ldots a_{n}^{m_{n}}$. Note that for all representatives $y=b \circ a^{\prime} \circ \tilde{b}$ of $x$ in the representation chosen, the coordinates $m_{1}, m_{2}, \ldots, m_{n}$ corresponding to $a^{\prime}$ must match to that of $a$ as above. Furthermore, there is a fixed element $\hat{b} \in B$ such that

$$
a_{n}^{m_{n}} \circ \ldots \circ a_{1}^{m_{1}}=a_{1}^{m_{1}} \circ \ldots \circ a_{n}^{m_{n}} \circ \hat{b} .
$$

Now, for any given representative $y=b \circ a^{\prime} \circ \tilde{b}$ with the coordinates for $a^{\prime}$ being $m_{0}=-1, m_{1}, \ldots, m_{n}$, the coordinate $m_{0}$ can be converted to +1 , by replacing $\tilde{b}$ by $\hat{b} \circ \tilde{b}$. Furthermore, using the arguments given in the proof of Theorem 6, there is a fixed automatic homomorphism $\varphi_{a}: B \rightarrow B$ such that $b \circ a=a \circ \varphi_{a}(b)$. The products $\varphi_{a}(b) \circ \tilde{b}$ or $\varphi_{a}(b) \circ \hat{b} \circ \tilde{b}$ can be carried out by componentwise addition of the vectors involved. Once this is done, the algorithm from Theorem 6 can be used to compare $y$ in the resulting representation with that of $x$. Thus equality is semiautomatic in the representation chosen.

The full proof of part (b) is omitted due to space constraints. The idea is to code a hard problem into a system of equations for a specific semiautomatic structure of the form $(A, B, \bullet,=; \circ)$. Each of the equations involves multiplication of an existentially quantified variable $x$ from $A$ with constants from $A$ and existentially quantified variables from $B$, the latter is automatic as it falls into the domain of $\bullet$. Then it is shown that solving such equations permits to solve the original computationally hard problem which cannot be solved by an automaton, implying that the structure cannot be semiautomatic. The coding mentioned above is either of an NP-hard problem (which is more direct to do) or in a more involved way of a Diophantine unsolvable problem, which then gives the undecidability result.

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# The Boolean Algebra of Piecewise Testable Languages 

Anton Konovalov and Victor Selivanov ${ }^{(\boxtimes)}$<br>A.P. Ershov Institute of Informatics Systems, Siberian Division Russian Academy of Sciences, Novosibirsk State University, Novosibirsk, Russia Jack@sibmail.ru, vseliv@iis.nsk.su


#### Abstract

We characterize up to isomorphism the Boolean algebra (BA, for short) of regular piecewise testable languages and show the decidability of classes of regular languages related to this characterization. This BA turns out isomorphic to several other natural BAs of regular languages, in particular to the BA of regular aperiodic languages.


Keywords: Boolean algebra • Frechét ideal • Regular language • Aperiodic language • Piecewise testable language

## 1 Introduction

Boolean algebras are of principal importance for several branches of mathematics. Accordingly, characterization of naturally arising BA's attracts attention of many researchers. As examples we mention characterizations of natural BA's in logic and computability theory $[3,7,12-14]$.

In automata theory, people consider many natural classes of languages which form BA's whose characterizations could be of some interest because they provide new information on well-known classes of regular languages. Due to the Stone duality, this contributes to understanding of the corresponding Stone spaces which are closely related to the profinite topology $[1,10]$.

The first papers in this direction appeared relatively recently $[8,16]$. In $[5,6,16]$ some fundamental BA's of regular languages and $\omega$-languages were characterized up to isomorphism. A surprising discovery was that those BAs are either quite simple (informally, very similar to the countable atomless BA) or else they are isomorphic to a distinguished $\mathrm{BA} \mathbb{A}$ described as follows (we use some well-known terminology on BA's which may be found e.g. in [2], see also the next section).

If $\mathbb{B}$ is a BA and $\alpha$ an ordinal, let $F_{\alpha}(\mathbb{B})$ be the $\alpha$-th iterated Frechét ideal of $\mathbb{B}$, and $\mathbb{B}^{(\alpha)}=\mathbb{B} / F_{\alpha}(\mathbb{B})$ is the $\alpha$-th Frechét derivative of $\mathbb{B}$. Frechét ideals are central for a useful classification of isomorphism types of countable BA's [4]. A very particular case of this is the following assertion: There is a unique,

[^47]up to isomorphism, countable BA $\mathbb{A}$ such that $F_{0}(\mathbb{A}) \subset F_{1}(\mathbb{A}) \subset \cdots \subset F_{\omega}(\mathbb{A})=$ $F_{\omega+1}(\mathbb{A})$, for each $n<\omega$ the $\operatorname{BA} \mathbb{A}^{(n)}$ is atomic with infinitely many atoms, and $\mathbb{A}^{(\omega)}$ is a countable atomless BA.

The BA's considered in $[5,6,16]$ are large in the sense that any of them extends the $\mathrm{BA} \mathbb{A}_{\Sigma}$ of regular aperiodic languages over an alphabet $\Sigma$ (or the corresponding BA of $\omega$-languages). In this note, we start to investigate "small" BA's of regular languages like the self-dual levels of Straubing-Thérien or Brzozowski hierarchies [18,20]. Probably, the most popular of those is the BA of piecewise testable languages $\mathbb{P}_{\Sigma}$ over $\Sigma$. Recall that a language $L$ is piecewise testable if there exists a number $k$ (depending only on $L$ ) such that membership of a word in the language depends only on the subwords (also called pieces) of length less than $k$ (we recall some characterizations of piecewise testable languages in the next section).

The main result of this paper may be formulated as follows: If $\Sigma$ contains at least two letters then $\mathbb{P}_{\Sigma}$ is isomorphic to $\mathbb{A}$.

Therefore, the BA $\mathbb{A}$ appears also among the "small" BA's of regular languages. Note that for the unary alphabet $\Sigma$ the $\mathrm{BA} \mathbb{P}_{\Sigma}$ coincides with $\mathbb{A}_{\Sigma}$ of regular aperiodic languages and is much simpler [16].

In Sect. 2 we provide the necessary background on BA's and regular languages, in Sect. 3 we give some important examples of piecewise testable languages, and in Sect. 4 we prove the main result.

## 2 Preliminaries

### 2.1 Preliminaries on Boolean Algebras

Here we briefly recall some relevant notions about BA's used in the sequel. We assume the reader to be familiar with basic notions related to BA's like ideal of a BA, quotient-algebra of a BA modulo a given ideal, and canonical homomorphism of a BA onto its quotient-algebra (for a detailed treatment of countable BA's see e.g. [2]). BA's are considered in the signature $\{\cup, \cap,-, 0,1\}$.

Recall that an element $a$ of a BA $\mathbb{B}$ is an atom if $a \neq 0$ and $x<a$ implies $x=0$. A BA $\mathbb{B}$ is called atomless if it has no atom, and it is called atomic if below any non-zero element there is an atom. The ideal of a $\mathrm{BA} \mathbb{B}$ generated by atoms is called Frechét ideal of $\mathbb{B}$. It consists of all finite unions (including the empty union which coincides with 0 ) of atoms and is denoted by $F(\mathbb{B})$. The quotient-algebra $\mathbb{B} / F(\mathbb{B})$ is called Frechét derivative of $\mathbb{B}$ and is also denoted by $\mathbb{B}^{\prime}$.

Define the transfinite sequence $\left\{F_{\alpha}(\mathbb{B})\right\}$ of iterated Frechét ideals of a BA $\mathbb{B}$ as follows: $F_{0}(\mathbb{B})=\{0\}, F_{\beta+1}(\mathbb{B})=\left\{x \mid x / F_{\beta}(\mathbb{B}) \in F\left(\mathbb{B}^{(\beta)}\right)\right\}$ where $\mathbb{B}^{(\beta)}=$ $\mathbb{B} / F_{\beta}(\mathbb{B})$, and $F_{\alpha}(\mathbb{B})=\bigcup_{\beta<\alpha} F_{\beta}(\mathbb{B})$ for a limit ordinal $\alpha$. This sequence is ascending under inclusion, and $F_{\alpha}(\mathbb{B})=F_{\alpha+1}(\mathbb{B})$ for some ordinal $\alpha$.

### 2.2 Preliminaries on Regular Languages

Here we briefly recall some notation, notions and facts on regular languages used in the sequel. Let $\Sigma^{*}$ denote the set of words over $\Sigma$ including the empty word
$\varepsilon$, and let $\Sigma^{+}:=\Sigma^{*} \backslash\{\varepsilon\}$. For any $n<\omega$, let $\Sigma^{n}$ be the set of words over $\Sigma$ of length $n$. The length of a word $w$ is denoted by $|w|$. Let $\sqsubseteq$ be the prefix partial order on $\Sigma^{*}$, i.e. $u \sqsubseteq v$ iff $u x=v$ for some $x \in \Sigma^{*}$.

We freely use some standard definitions and notation on words, regular languages and finite automata like regular expressions or deterministic finite automata (DFA) denoted as $\mathcal{A}=\left(Q, \Sigma, f, q_{0}, F\right)$ (for a more systematic treatment see e.g. $[18,20,21])$. The automaton $\mathcal{A}$ recognizes the language $L_{\mathcal{A}}=\{u \in$ $\left.\Sigma^{*} \mid q_{0} \cdot u \in F\right\}$ where $q \cdot u=f(q, u)$ is the state reached by $\mathcal{A}$ when it reads the word $u$ started from a state $q$. Let $\mathcal{R}_{\Sigma}$ denote the class of regular languages over $\Sigma$ (i.e., languages recognized by DFA's or, equivalently, denoted by regular expressions). For $L \in \mathcal{R}_{\Sigma}, \mathcal{M}_{L}$ denotes the minimal DFA that recognizes $L$; the minimal DFA is unique up to isomorphism.

For $k<\omega$, a language $L \in \mathcal{R}_{\Sigma}$ is called $k$-sparse if the function $p_{L}(n)=$ $\left|\Sigma^{n} \cap L\right|$ is $O\left(n^{k}\right)$ (cf. [19,21]). Let $\mathcal{S}_{k}$ denote the class of regular $k$-sparse languages over $\Sigma$. Note that $\mathcal{S}_{0} \subset \mathcal{S}_{1} \subset \cdots$. Languages from $\mathcal{S}_{\omega}=\bigcup_{k<\omega} \mathcal{S}_{k}$ are called sparse languages. Note that 0 -sparse languages are also known as slender languages. We will use the following characterizations of $k$-sparse languages in terms of regular expressions (cf. [19,21]).

Proposition 1. For any $L \in \mathcal{R}_{\Sigma}$ and $k<\omega, L \in \mathcal{S}_{k}$ iff $L$ is a finite union of languages $x y_{0}^{*} z_{0} \cdots y_{k}^{*} z_{k}$ where $x, y_{i}, z_{i} \in \Sigma^{*}$.

There is also a useful characterization of sparse languages in terms of (graphs of) their minimal automata (see e.g. $[9,16]$ ). We say that a DFA $\mathcal{A}$ has an $\omega$ pattern if there are $u, v_{1}, v_{2}, w \in \Sigma^{*}$ such that $v_{1}, v_{2}$ are $\sqsubseteq$-incomparable (i.e., $v_{1}(i) \neq v_{2}(i)$ for some $\left.i<\left|v_{1}\right|,\left|v_{2}\right|\right), q_{0} \cdot u=q_{0} \cdot u v_{1}=q_{0} \cdot u v_{2}$ and $q_{0} \cdot u v_{1} w \in F$.

Proposition 2. For any $L \in \mathcal{R}_{\Sigma}, L \in \mathcal{S}_{\omega}$ iff the minimal DFA of $L$ has no $\omega$-pattern.

For $q \in Q$ and $u \in \Sigma^{*}$, let $(q, u)$ denote the path in $\mathcal{A}$ along $u$ started at $q$, and let $Q(q, u)=\{q \cdot v \mid v \sqsubseteq u\}$. We say that the path $(q, u)$ meets a set of states $G \subseteq Q$ if $Q(q, u) \cap G \neq \emptyset$. The path $(q, u)$ is a cycle of $\mathcal{A}$ if $u$ is nonempty and $q \cdot u=q$. A cycle is simple if it has no repeated vertices other than the starting and ending vertices. Let $C_{\mathcal{A}}$ be the set of all $Q(q, u)$ where $(q, u)$ is a cycle of $\mathcal{A}$. Define the preorder $\leq$ on $C_{\mathcal{A}}$ as follows: $G \leq H$ if there is a path of $\mathcal{A}$ starting in $G$ and ending in $H$. Let $\equiv$ denote the equivalence relation on $C_{\mathcal{A}}$ induced by $\leq$.

Note that if $G \equiv H$ then $K \equiv G$ for some $K \supseteq G \cup H$, i.e. any element [ $G$ ] of the quotient-set $C_{\mathcal{A}} / \equiv$ has a largest set under inclusion; these largest sets are called strongly connected components (SCC's) of $\mathcal{A}$ and they may serve as canonical representatives for the equivalence classes $[G]$. The next result is known and easily follows from the previous proposition:

Proposition 3. For any sparse language $L=L_{\mathcal{A}} \in \mathcal{R}_{\Sigma}$ and any $u \in L$, if the path $\left(q_{0}, u\right)$ meets a SCC $G$ of $\mathcal{A}$ then $[G]=\{G\}$ and $G=Q(s, v)$ for some simple cycle of $\mathcal{A}$.

We also recall the following classical fact of automata theory. A DFA $\mathcal{A}$ is called counter-free if $q \cdot u^{n}=q$ implies $q \cdot u=q$, for all $q \in Q$, nonempty word $u \in \Sigma^{*}$ and $n>0$.

Proposition 4. For any $L \in \mathcal{R}_{\Sigma}$ the following conditions are equivalent:

1. There is $n<\omega$ such that $x y^{n} z \in L$ iff $x y^{n+1} z \in L$ for all $x, y, z \in \Sigma^{*}$.
2. The minimal DFA of $L$ is counter-free.

Languages satisfying the conditions in the last proposition are called aperiodic. There are several other important characterizations of the class $\mathcal{A}_{\Sigma}$ of aperiodic languages, in particular $L \subseteq \Sigma^{+}$is aperiodic iff $L$ is defined by a first order sentence of signature $\sigma=\left\{<, Q_{a}\right\}_{a \in \Sigma}$ (see e.g. [18,20]) for additional details). The class $\mathcal{A}_{\Sigma}$ is closed under the Boolean operations, let $\mathbb{A}_{\Sigma}$ be the corresponding BA.

A popular subclass of the aperiodic languages is the class $\mathcal{P}_{\Sigma}$ of piecewise testable languages (see e.g. [11,17]). It has several nice characterizations, in particular $L \subseteq \Sigma^{+}$is piecewise testable iff $L$ is defined by a boolean combination of existential sentences of signature $\sigma$. Below we use the following characterization $[11,17]$ in terms of forbidden patterns $T_{0}, T_{1}$ (where $x, z \in \Sigma^{*}, u, v \in \Sigma^{+}$) shown on Figs. 1 and 2 below.


Fig. 1. Pattern $T_{0}$, where $s_{0}$ is the initial state and $s \in F \Leftrightarrow s^{\prime} \notin F$.

Proposition 5. For any regular language $L, L \in \mathcal{P}_{\Sigma}$ iff any $D F A$ recognizing $L$ does not contain any of the patterns $T_{0}$ and $T_{1}$.

Restricting to minimal automata we immediately obtain:
Corollary 1. For any regular language $L, L \in \mathcal{P}_{\Sigma}$ iff the minimal $D F A$ of $L$ contains neither non-singleton cycles nor the patterns $T_{1}$.

Remark. In the literature one often meets two versions of piecewise testable languages, namely the languages which can or cannot contain the empty word. Usually, proofs for the two cases are slightly different but formulations essentially coincide. This also applies to our case. In the sequel we stick for simplicity to the case of languages $L \subseteq \Sigma^{+}$of non-empty words; the other version is considered in a similar fashion.


Fig. 2. Pattern $T_{1}$, where $s_{0}$ is the initial state and $s \in F \Leftrightarrow s^{\prime} \notin F$.

## 3 A Family of Piecewise Testable Languages

Here we show that for all $x_{0}, \ldots, x_{n} \in \Sigma^{*}$ with $x_{0} \cdots x_{n} \neq \varepsilon$, and for all $a_{1}, \ldots, a_{n} \in \Sigma$, the language $x_{0} a^{*}{ }_{1} x_{1} \cdots a_{n}^{*} x_{n}$ is piecewise testable. This easy fact is crucial for the characterization of BA $\mathbb{P}_{\Sigma}$. Since we have not seen it in the literature, we provide a detailed proof using the automata-theoretic characterization of piecewise testable languages from the previous section. From a discussion with J.-E. Pin we learned that a natural semigroup-theoretic proof is also possible.

By simple expression we mean a regular expression $x_{0} a_{1}^{*} x_{1} \cdots a_{n}^{*} x_{n}$ where $x_{1}, \ldots, x_{n-1} \in \Sigma^{+}, x_{0}, x_{n} \in \Sigma^{*}, x_{0} \cdots x_{n} \neq \varepsilon, a_{1}, \ldots, a_{n} \in \Sigma$, and the first letter of $x_{i}$ is distinct from $a_{i}$ for all $0<i<n$, and also for $i=n$ if $x_{n}$ is non-empty.

Lemma 1. Let $L \subseteq \Sigma^{+}$be the regular language defined by a simple expression as above. Then the minimal DFA of $L$ does not contain non-singleton cycles and patterns $T_{1}$ from the previous section.

Proof. For $n=0$ the assertion is clear. Let $n>0$ and let $m_{i}$ be the length of the word $x_{i}$ for $i \leq n$, so $x_{i}=x_{i, 1} \cdots x_{i, m_{i}}$ for unique letters $x_{i, 1}, \ldots, x_{i, m_{i}} \in \Sigma$. Note that $m_{i}>0$ for all $0<i<n$. For $n=1$ the minimal DFA for $L=x_{0} a^{*}{ }_{1} x_{1}$ looks as shown on Fig. 3 where $s_{0}$ is the initial state, $s_{2}$ is the unique accepting state, and the arrows to the rejecting sink state $s$ carry all letters required by the definition of DFA. Note that in the case $x_{0}=\varepsilon$ we have $x_{1} \neq \varepsilon, s_{0}=s_{1}$ and the nodes $q_{1}, \ldots, q_{m_{0}-1}$ and the corresponding arrows disappear, while in the case $x_{1}=\varepsilon$ we have $x_{0} \neq \varepsilon, s_{1}=s_{2}$ is the accepting state, and the nodes $p_{1}, \ldots, p_{m_{1}-1}$ and the corresponding arrows disappear. It is easy to see that the minimal DFA has the desired property.

For $n=2$ the minimal DFA for $L=x_{0} a^{*}{ }_{1} x_{1} a_{2}^{*} x_{2}$ looks similar: $x_{0}$ leads from $s_{0}$ to $s_{1}, x_{1}$ leads from $s_{1}$ to $s_{2}, x_{2}$ leads from $s_{2}$ to the unique accepting state $s_{3}$, and all other arrows lead to the rejecting sink state $s$.


Fig. 3. Minimal DFA for $n=1$.

The argument works in a similar way for any $n \geq 3$.
Lemma 2. For any $n>0, x_{0}, \ldots x_{n} \in \Sigma^{*}$ and $a_{1}, \ldots a_{n} \in \Sigma$, the language $L=x_{0} a_{1}^{+} x_{1} \cdots a_{n}^{+} x_{n}$ may be represented by a simple expression $y_{0} b_{1}^{*} y_{1} \cdots b_{m}^{*} y_{m}$ such that $m \leq n,\left\{b_{1}, \ldots, b_{m}\right\}=\left\{a_{1}, \ldots, a_{n}\right\}$ and $y_{i-1}$ has the letter $b_{i}$ for all $1 \leq i \leq m$.

Proof. By induction on $n$. For $n=1$, write $x_{1}$ as $x_{1}=a_{1}^{l} y_{1}$ where $l \geq 0$ and $y_{1}$ is empty or starts with a letter distinct from $a_{1}$. Then a desired representation is $L=\left(x_{0} a_{1}^{l+1}\right) a_{1}^{*} y_{1}$.

For $n>1$, set $K=x_{1} a_{2}^{+} x_{2} \ldots x_{n-1} a_{n}^{+} x_{n}$, then $L=x_{0} a_{1}^{+} K$. By induction, $K$ has a desired representation $y_{1} b_{2}^{*} y_{2} \ldots b_{m}^{*} y_{m}$, so $L=x_{0} a_{1}^{+} y_{1} b_{2}^{*} y_{2} \ldots b_{m}^{*} y_{m}$. If $a_{1} \neq b_{2}$, represent $x_{0} a_{1}^{+} y_{1}$ as $\left(x_{0} a_{1}^{l+1}\right) a^{*} z_{1}$ similar to the case $n=1$, then $L=\left(x_{0} a_{1}^{l+1}\right) a^{*} z_{1} b_{2}^{*} y_{2} \ldots b_{m}^{*} y_{m}$ is a desired representation. The same argument works for the case when $a_{1}=b_{2}$ and $y_{1} \notin b_{2}^{*}$. Finally, let $a_{1}=b_{2}$ and $y_{1}=b_{2}^{l}$ for some $l \geq 0$. Then $L=\left(x_{0} b_{2}^{l+1}\right) b_{2}^{*} y_{2} \ldots b_{m}^{*} y_{m}$ is a desired representation.
Lemma 3. For all $x_{0}, \ldots, x_{n} \in \Sigma^{*}$ with $x_{0} \cdots x_{n} \neq \varepsilon$, and for all $a_{1}, \ldots, a_{n} \in$ $\Sigma$, the language $L=x_{0} a^{*}{ }_{1} x_{1} \cdots a_{n}^{*} x_{n}$ is piecewise testable.

Proof. It is easy to see that $L$ is the union of languages $L(S)$ for all $S \subseteq\{1, \ldots, n\}$ where $L(S)$ is obtained from $x_{0} a_{1}^{+} x_{1} \cdots a_{n}^{+} x_{n}$ by deleting the expressions $a_{i}^{+}$with $i \in S$ (e.g., for $n=2$ we have $x_{0} a_{1}^{*} x_{1} a_{2}^{*} x_{2}=x_{0} x_{1} x_{2} \cup x_{0} a_{1}^{+} x_{1} x_{2} \cup x_{0} x_{1} a_{2}^{+} x_{2} \cup$ $\left.x_{0} a_{1}^{+} x_{1} a_{2}^{+} x_{2}\right)$.

By Lemmas 1, 2 and Corollary 1, any $L(S)$ is piecewise testable. Since $\mathbb{P}_{\Sigma}$ is closed under finite union, $L$ is piecewise testable.

Remark. As is well known (see e.g. [15] and references therein), the class of piecewise testable languages is stratified to the $\omega$ levels of the difference hierarchy over the class of languages defined by existential sentences of signature $\sigma$. It is not hard to show that no fixed level of this hierarchy contains all languages from the previous lemma.

## 4 Proof of the Main Result

Now it is easy to adjust the proof in [16] to the case of piecewise testable languages, so we give only short comments.

Since $\mathbb{P}_{\Sigma}$ is a subalgebra of $\mathbb{A}_{\Sigma}$, the next three lemmas are checked in the same way as Lemmas 3-5 in [16].

Lemma 4. For all $k<\omega$ and $x, z_{i} \in \Sigma^{*}, y_{i} \in \Sigma \cup\{\varepsilon\}, x y_{1}^{*} z_{1} \cdots y_{k}^{*} z_{k} / F_{k+1}\left(\mathbb{P}_{\Sigma}\right)$ is either zero or an atom in the $B A \mathbb{P}_{\Sigma}^{(k)}$. If $L$ is a finite union of such languages $x y_{1}^{*} z_{1} \cdots y_{k}^{*} z_{k}$ then $L \in F_{k+1}\left(\mathbb{P}_{\Sigma}\right)$.

Lemma 5. Let $u, w \in \Sigma^{*}$ and let $v_{1}, v_{2} \in \Sigma$ satisfy $v_{1} \neq v_{2}$. Then $u v_{1}^{*} w \notin$ $F_{1}\left(\mathbb{P}_{\Sigma}\right), u v_{1}^{*} v_{2} v_{1}^{*} w \notin F_{2}\left(\mathbb{P}_{\Sigma}\right), u v_{1}^{*} v_{2} v_{1}^{*} v_{2} v_{1}^{*} w \notin F_{3}\left(\mathbb{P}_{\Sigma}\right)$ and so on.

Lemma 6. For any $L \in F_{\omega}\left(\mathbb{P}_{\Sigma}\right)$ the minimal DFA of $L$ has no $\omega$-pattern.
The next two theorems and corollaries are also proved in the same way as the corresponding facts in [16].

Theorem 1. For any $L \in \mathcal{R}_{\Sigma}$ the following conditions are equivalent:

1. $L \in F_{\omega}\left(\mathbb{P}_{\Sigma}\right)$.
2. The minimal DFA of $L$ has neither non-singleton cycles nor $\omega$-patterns.
3. $L \in \mathcal{P}_{\Sigma} \cap \mathcal{S}_{\omega}$.
4. $L$ is a finite union of languages $x y_{0}^{*} z_{0} \cdots y_{k}^{*} z_{k}$, where $k<\omega, x, z_{i} \in \Sigma^{*}$, and $y_{0}, \ldots, y_{k} \in \Sigma$.

Corollary 2. The class of regular languages $F_{\omega}\left(\mathbb{P}_{\Sigma}\right)$ is decidable.
Next we characterize the ideals $F_{k}\left(\mathbb{P}_{\Sigma}\right)$ for $k<\omega$. Observe that $F_{0}\left(\mathbb{P}_{\Sigma}\right)=\{\emptyset\}$ and $F_{1}\left(\mathbb{P}_{\Sigma}\right)$ is the class of finite languages over $\Sigma$.

Theorem 2. For any $k<\omega$ and $L \in \mathcal{R}_{\Sigma}$ the following conditions are equivalent:

1. $L \in F_{k+2}\left(\mathbb{P}_{\Sigma}\right)$.
2. The minimal DFA of $L$ has neither non-singleton cycles nor $\omega$-patterns, nor chains $F_{0}<\cdots<F_{k+1}$ of SCC's of $\mathcal{M}_{L}$ such that some accepting state of $\mathcal{M}_{L}$ is reachable from $F_{k+1}$.
3. $L \in \mathcal{P}_{\Sigma} \cap \mathcal{S}_{k}$.
4. $L$ is a finite union of languages $x y_{0}^{*} z_{0} \cdots y_{k}^{*} z_{k}$, where $x, z_{i} \in \Sigma^{*}, y_{0}, \ldots, y_{k} \in$ $\Sigma$.

Corollary 3. For any $k<\omega$, the class of regular languages $F_{k}\left(\mathbb{P}_{\Sigma}\right)$ is decidable uniformly on $k$.

We are ready to complete the proof of the main result:
Theorem 3. The $B A \mathbb{P}_{\Sigma}$ is a unique, up to isomorphism, countable $B A \mathbb{A}$ such that $F_{0}(\mathbb{A}) \subset F_{1}(\mathbb{A}) \subset \cdots \subset F_{\omega}(\mathbb{A})=F_{\omega+1}(\mathbb{A})$, for each $n<\omega$ the $B A \mathbb{A}^{(n)}$ is atomic with infinitely many atoms, and $\mathbb{A}^{(\omega)}$ is a countable atomless $B A$.

Proof. First we check that $F_{k}\left(\mathbb{P}_{\Sigma}\right) \subset F_{k+1}\left(\mathbb{P}_{\Sigma}\right)$ for each $k<\omega$. For $k=0$ this is clear. Choose distinct $a, b \in \Sigma$, it suffices to show that

$$
a^{*} \in F_{2}\left(\mathbb{P}_{\Sigma}\right) \backslash F_{1}\left(\mathbb{P}_{\Sigma}\right), a^{*} b a^{*} \in F_{3}\left(\mathbb{P}_{\Sigma}\right) \backslash F_{2}\left(\mathbb{P}_{\Sigma}\right), a^{*} b a^{*} b a^{*} \in F_{4}\left(\mathbb{P}_{\Sigma}\right) \backslash F_{3}\left(\mathbb{P}_{\Sigma}\right)
$$

so on. By Theorem 1 and Lemma 4,

$$
a^{*} \in F_{2}\left(\mathbb{P}_{\Sigma}\right), a^{*} b a^{*} \in F_{3}\left(\mathbb{P}_{\Sigma}\right), a^{*} b a^{*} b a^{*} \in F_{4}\left(\mathbb{P}_{\Sigma}\right),
$$

so on. By Lemma 5,

$$
a^{*} \notin F_{1}\left(\mathbb{P}_{\Sigma}\right), a^{*} b a^{*} \notin F_{2}\left(\mathbb{P}_{\Sigma}\right), a^{*} b a^{*} b a^{*} \notin F_{3}\left(\mathbb{P}_{\Sigma}\right),
$$

so on. Note that precise proof needs induction on $k$, as in [16].
By Lemmas 4-6 the elements $a^{*} / F_{1}\left(\mathbb{P}_{\Sigma}\right), a^{*} b a^{*} / F_{2}\left(\mathbb{P}_{\Sigma}\right), \ldots$ are atoms respectively in the BA's $\mathbb{P}_{\Sigma}^{(1)}, \mathbb{P}_{\Sigma}^{(2)}, \ldots$, and this also holds for the elements

$$
b^{n} a^{*} / F_{1}\left(\mathbb{P}_{\Sigma}\right), b^{n} a^{*} b a^{*} / F_{2}\left(\mathbb{P}_{\Sigma}\right), \ldots
$$

As the languages $b^{n} a^{*}$ (and also the languages $b^{n} a^{*} b a^{*}$ and so on.) are pairwise disjoint for distinct $n$, the BA's $\mathbb{P}_{\Sigma}^{(1)}, \mathbb{P}_{\Sigma}^{(2)}, \ldots$ have infinitely many atoms (as also the $\left.\operatorname{BA} \mathbb{P}_{\Sigma}^{(0)}=\mathbb{P}_{\Sigma}\right)$.

Next we check that the $\operatorname{BA} \mathbb{P}_{\Sigma}^{(k)}$ is atomic for each $k<\omega$. For $k<2$ this is clear, it remains to check that for any $k<\omega$ and $L \in \mathbb{P}_{\Sigma} \backslash F_{k+2}\left(\mathbb{P}_{\Sigma}\right)$ there is $A \in \mathbb{P}_{\Sigma}$ such that $A \subseteq L$ and $A / F_{k+2}\left(\mathbb{P}_{\Sigma}\right)$ is an atom in $\mathbb{P}_{\Sigma}^{(k+2)}$. We distinguish the cases $L \notin F_{\omega}\left(\mathbb{P}_{\Sigma}\right)$ and $L \in F_{\omega}\left(\mathbb{P}_{\Sigma}\right)$.

In the first case Proposition 2 and Corollary 1 imply that the minimal DFA of $L$ has an $\omega$-pattern with the words $u, v_{1}, v_{1}, w$ such that $v_{1}, v_{2}$ are distinct letters from $\Sigma$, so $A$ exists by Theorem 1 and Lemma 5. In the second case, by Theorem 2 there are SCC's $F_{0}<\cdots<F_{k+1}$ (having by Corollary 1 the form $F_{i}=\left\{s_{i}\right\}, s_{i} y_{i}=s_{i}$ for some $\left.y_{0}, \ldots, y_{k+1} \in \Sigma\right)$ and words $z_{0}, \ldots, z_{k} \in \Sigma^{+}$, $x, z_{k+1} \in \Sigma^{*}$, such that the first letter of $z_{i}$ is distinct from $y_{i}$ for each $i \leq k$, $q_{0} x=s_{0}$ and $x z_{0} \cdots z_{k+1} \in L$. Then the language $A=x y_{0}^{*} z_{0} \cdots y_{k+1}^{*} z_{k+1}$ has the desired properties.

It remains to show that for any $L \in \mathbb{P}_{\Sigma} \backslash F_{\omega}\left(\mathbb{P}_{\Sigma}\right)$ there is $M \in \mathbb{P}_{\Sigma} \backslash F_{\omega}\left(\mathbb{P}_{\Sigma}\right)$ such that $M \subset L$ and $M, L \backslash M \notin F_{\omega}\left(\mathbb{P}_{\Sigma}\right)$. The minimal DFA $\mathcal{M}_{L}$ of $L$ has again an $\omega$-pattern for some $u, v_{1}, v_{2}, w$ as above. We can thus take $M=u v_{1}\left(v_{1}+\right.$ $\left.v_{2}\right)^{*} w$.

The main result and the corresponding result in [16] immediately imply:
Corollary 4. Over any alphabet $\Sigma, \mathbb{P}_{\Sigma}$ is isomorphic to $\mathbb{A}_{\Sigma}$.

## 5 Conclusion

We see that the BA $\mathbb{A}$ from Introduction is astonishingly robust in characterising BAs of regular languages: all such non-degenerate BAs considered so far are
isomorphic to $\mathbb{A}$. At the same time, the structure of many BA's of regular languages remains unclear. In particular, this is the case for higher self-dual levels of the Straubing-Thérien, Brzozowski and other similar hierarchies of regular languages [15]. The same applies to the corresponding BA's of regular $\omega$-languages. An intriguing open question is whether any such non-degenerate BA is isomorphic to $\mathbb{A}$.

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# On the Lattices of Effectively Open Sets 

Oleg V. Kudinov ${ }^{1}$ and Victor L. Selivanov ${ }^{2(\boxtimes)}$<br>1 S.L. Sobolev Institute of Mathematics SB RAS, Novosibirsk, Russia<br>kud@math.nsc.ru<br>2 A.P. Ershov Institute of Informatics Systems SB RAS, Novosibirsk State University, Novosibirsk, Russia vseliv@iis.nsk.su


#### Abstract

We show that for many natural computable metric spaces and computable domains the first order theory of the lattice of effectively open sets is hereditarily undecidable. Moreover, for several important spaces (e.g., finite-dimensional Euclidean spaces and the domain $P \omega$ ) this theory is $m$-equivalent to the first-order arithmetic.


Keywords: Lattice • Effectively open set • First-order theory • Interpretation

## 1 Introduction

The lattice $\mathcal{E}$ of computably enumerable (c.e.) sets is a popular object of study in Computability Theory [12]. A principal fact about this lattice is the undecidability of its first-order theory $\operatorname{Th}(\mathcal{E})$ [5,6]. Moreover, $\operatorname{Th}(\mathcal{E})$ is known [7] to be $m$-equivalent to the first-order arithmetic $\operatorname{Th}(\mathbb{N})$ where $\mathbb{N}:=(\omega ;+, \times)$, i.e., to the $\omega^{\prime}$ th iteration $\emptyset^{(\omega)}$ of Turing jump starting with the empty set.

The lattice $\mathcal{E}$ may be considered as the lattice $\Sigma_{1}^{0}(\omega)$ of effectively open subsets of the discrete topological space $\omega$. The lattices of effectively open sets $\Sigma_{1}^{0}(X)$ of various "effective" topological spaces $X$ are of interest in Computable Analysis [13] and Effective Descriptive Set Theory [9,11], hence it is natural to ask which results about $\mathcal{E}$ hold true for the lattices of effectively open sets of "natural" effective topological spaces. It seems that not much is known about this question. To our knowledge, only the case of Cantor space $\mathcal{C}$ and Baire space $\mathcal{N}$ was studied to some extent, in the context of the theory of $\Pi_{1}^{0}$-classes. In [10] (see remarks after Main Theorem 3.1 in Introduction) it is shown that $T h\left(\Pi_{1}^{0}(\mathcal{C})\right)$ is $m$-equivalent to $\emptyset^{(\omega)}$. Since the lattice $\Pi_{1}^{0}(\mathcal{C})$ is anti-isomorphic to $\Sigma_{1}^{0}(\mathcal{C})$, this settles the question for $\mathcal{C}$. To our knowledge, similar questions for $\mathcal{N}$ (even the decidability of $\operatorname{Th}\left(\Sigma_{1}^{0}(\mathcal{N})\right.$ )) were open.

In this paper, we attempt to make next steps in this direction. First, in Sect. 2 we show that for many natural effective spaces $X$ the theory $\operatorname{Th}\left(\Sigma_{1}^{0}(X)\right)$ (including Baire space) is undecidable. This includes the computable metric

[^48]spaces without isolated points, and many natural classes of domains (as well as many non-complete partial orders with Scott topology).

In Sect. 3, we try to precisely estimate the $m$-degree of $T h\left(\Sigma_{1}^{0}(X)\right)$. We prove that for Euclidean finite-dimensional spaces $X$ and for some natural domains, $T h\left(\Sigma_{1}^{0}(X)\right)$ is $m$-equivalent to $\emptyset^{(\omega)}$. The methods of this section apply mainly to locally compact spaces. The precise estimation of the complexity of $\operatorname{Th}\left(\Sigma_{1}^{0}(X)\right)$ turns out to be very subtle and strongly depending on the topology of $X$. For many natural spaces $X$ we still have a big gap between the known (to us) lower and the upper bound for $\operatorname{Th}\left(\Sigma_{1}^{0}(X)\right)$. In particular, for Baire space we currently only know the estimate $\emptyset^{\prime} \leq_{m} T h\left(\Sigma_{1}^{0}(X)\right) \leq_{m} O^{(\omega)}$ where $O$ is the Kleene ordinal notation system which is a $\Pi_{1}^{1}$-complete set.

## 2 Undecidability of $\operatorname{Th}\left(\Sigma_{1}^{0}(X)\right)$

Here we show that for many natural effective spaces $X$ the theory $\operatorname{Th}\left(\Sigma_{1}^{0}(X)\right)$ is undecidable. First we explain what we mean by effectively open sets. For any countably based topological space $X$ and any numbering $\beta: \omega \rightarrow P(X)$ of a basis of $X$, define a numbering $\pi: \omega \rightarrow P(X)$ by $\pi(n)=\bigcup \beta\left[W_{n}\right]$ where $\left\{W_{n}\right\}$ is the standard numbering of c.e. sets [12] and $\beta\left[W_{n}\right]$ is the image $\left\{\beta(a) \mid a \in W_{n}\right\}$. The sets in $\pi[\omega]$ are called effectively open sets in $X$. Thus, the set of effectively open sets is always equipped with the induced numbering $\pi$, hence it makes sense to speak about computable sequences of effectively open sets.

We define some particular classes of effective spaces relevant to this paper. A computable metric space [13] is a triple $(X, d, \nu)$, where $(X, d)$ is a metric space and $\nu: \omega \rightarrow X$ is a numbering of a dense subset $\operatorname{rng}(\delta)$ of $X$ such that the set $\{(i, j, k, l) \mid \varkappa(k)<d(\nu(i), \nu(j))<\varkappa(l)\}$ is c.e. Here $\varkappa$ is a computable numbering of the set $\mathbb{Q}$ of rationals. Any computable metric space $(X, d, \nu)$ gives rise to a numbering $\beta$ of the standard basis $\beta_{\langle m, n\rangle}=B\left(\nu_{m}, \varkappa_{n}\right)$ were $\langle m, n\rangle$ is the Cantor pairing and $B\left(\nu_{m}, \varkappa_{n}\right)$ is the basic open ball with center $\nu_{m}$ and radius $\varkappa_{n}$. By a strongly computable metric space (SCMS) we mean a computable metric space such that there exists an infinite computable sequence $\left\{B_{n}\right\}$ of pairwise disjoint basic open balls. The metric spaces $\omega, \mathbb{Q}, \mathcal{C}:=2^{\omega}, \mathcal{N}:=\omega^{\omega}, \mathbb{R}^{n}$, and $\mathbb{R}^{\omega}$ equipped with the standard metrics and with natural numberings of dense subsets are SCMS. Any computable metric space without isolated points is a SCMS. Recall that a theory of signature $\sigma$ is hereditarily undecidable if any its subtheory of signature $\sigma$ is undecidable.

Theorem 1. Let $X$ be a SCMS. Then $T h\left(\Sigma_{1}^{0}(X)\right)$ is hereditarily undecidable.
Proof. Since the theory $\operatorname{Th}(\mathcal{E})$ is hereditarily undecidable [5,6], it suffices to define the structure $\mathcal{E}$ in the structure $\Sigma_{1}^{0}(X)$ by first-order formulas with parameters; in our case two parameters $v, w$ will suffice. Consider the formulas $\phi_{U}(x, v, w):=v \subseteq x \wedge x \subseteq w$ and $\phi_{\subseteq}(x, y, v, w):=\phi_{U}(x, v, w) \wedge \phi_{U}(x, v, w) \wedge x \subseteq$ $y$. For any values $V, W \in \Sigma_{1}^{0}(X)$ for parameters $v, w$ with $V \subseteq W$, the formulas define the sublattice $\mathcal{D}$ of $\Sigma_{1}^{0}(X)$ formed by the sets lying between $V$ and $W$.

Therefore, it suffices to find effectively open sets $V, W$ such that the lattice $\mathcal{D}$ is isomorphic to $\mathcal{E}$.

Let $\left\{B_{n}\right\}$ be the sequence from the definition of SCMS and let $B_{n}^{\prime}$ be obtained from the ball $B_{n}$ by removing its center $c_{n}$. Then $\left\{B_{n}^{\prime}\right\}$ is a computable sequence of effectively open sets, hence $V:=\bigcup_{n} B_{n}^{\prime}$ and $W:=\bigcup_{n} B_{n}$ are effectively open and $W \backslash V=\left\{c_{0}, c_{1}, \ldots\right\}$. From the definition of SCMS it is easy to see that the function $D \mapsto\left\{n \mid c_{n} \in D\right\}$ is a desired isomorphism between $\mathcal{D}$ and $\mathcal{E}$.

Next we recall some definitions related to domain theory (for more details see e.g. $[1,3,11])$. Let $X$ be a $T_{0}$-space. For $x, y \in X$, let $x \leq y$ denote that $x \in U$ implies $y \in U$, for all open sets $U$. The relation $\leq$ is a partial order known as the specialization order. Let $F(X)$ be the set of finitary elements of $X$ (known also as compact elements), i.e. elements $p \in X$ such that the upper cone $\uparrow p=\{x \mid p \leq x\}$ is open. Such open cones are called $f$-sets. The space $X$ is called a $\varphi$-space if every open set is a union of $f$-sets. The $\varphi$-space $X$ is an $f$-space if any compatible elements $c, d \in F(X)$ have a least upper bound w.r.t. $\leq$ (compatibility means that $c, d$ have an upper bound in $F(X)$ ). An $f$-space $X$ is an $f_{0}$-space if $F(X)$ has a least element.

By a computable $\varphi$-space we mean a pair $(X, \delta)$ consisting of a $\varphi$-space $X$ and a numbering $\delta: \omega \rightarrow F(X)$ of all the finitary elements such that the specialization order is c.e. on the finitary elements (i.e., the relation $\delta_{x} \leq \delta_{y}$ is c.e.). Setting $\beta(n):=\uparrow \delta_{n}$ we obtain a numbering of a topological basis of $X$. Thus, we have a notion of an effective open set in every computable $\varphi$-space.

By a strongly computable $\varphi$-space (SCPS) we mean a computable $\varphi$-space $X$ such that the specialization order is computable on the finitary elements, and there is a computable sequence $\left\{c_{n}\right\}$ of pairwise incomparable finitary elements. Although the restrictions on SCPSs are rather strong, many popular domains are SCPSs. In particular, this applies to all $\varphi$-spaces mentioned below in this section. A SCPS $X$ is a strongly computable $f_{0}$-space (SCFS) if it is an $f_{0}$ space, the relation of compatibility is computable on $F(X)$, and the supremum of compatible finitary elements is computable.

Let $\omega^{\leq \omega}$ be the completion of the partial ordering ( $\omega^{*} ; \sqsubseteq$ ) where $\omega^{*}$ is the set of finite strings of naturals and $\sqsubseteq$ is the prefix relation. Of course, $\omega \leq \omega=\omega^{*} \cup \omega^{\omega}$ consists of all finite and infinite strings of natural numbers. For every $2 \leq n<\omega$, let $n \leq \omega$ be obtained in the same way from $\left(n^{*} ; \sqsubseteq\right)$. Thus, $n \leq \omega=n^{*} \cup n^{\omega}$ consists of all finite and infinite words over the alphabet $\{0, \ldots, n-1\}$. Let $P \omega$ be the powerset of $\omega$ with the Scott topology for the subset relation, hence the finitary elements of $P \omega$ are exactly the finite subsets of $\omega$.

Let $\omega_{\perp}$ be the "flat" domain obtained from the discrete space of $\omega$ by adjoining a new bottom element $\perp$. Let $\omega_{\perp}^{\omega}$ be the space of partial functions $g: \omega \rightharpoonup \omega$ with the usual structure of a $\varphi$-space given by the Scott topology for the subgraph relation (as usual, we identify the partial function $g$ with the total function $\tilde{g}: \omega \rightarrow \omega_{\perp}=\omega \cup\{\perp\}$ where $g(x)$ is undefined iff $\tilde{g}(x)=\perp$, for some "bottom" element $\perp \notin \omega$ ). For each $n, 2 \leq n<\omega$, let $n_{\perp}^{\omega}$ be the space of partial functions $g: \omega \rightharpoonup\{0, \ldots, n-1\}$ defined similarly to $\omega_{\perp}^{\omega}$. Obviously, $\omega \leq \omega, n \leq \omega, P \omega, \omega_{\perp}^{\omega}, n_{\perp}^{\omega}$ are complete SCFSs. As is well known (see e.g. [2]), for any (complete) $f_{0}$-spaces
$X, Y$ the space $Y^{X}$ of continuous functions from $X$ to $Y$ with the topology of pointwise convergence is again a (complete) $f_{0}$-space. An inspection of the corresponding proofs shows that if $X, Y$ are (complete) SCFSs then so is also $Y^{X}$. Therefore, any space of continuous partial functionals over $\omega$ of a finite type is a complete SCFS. In particular, this applies to the spaces $\mathbb{F}_{n}$ defined by induction $\mathbb{F}_{0}:=\omega_{\perp}, \mathbb{F}_{n+1}:=\omega_{\perp}^{\mathbb{F}_{n}}$.

Theorem 2. Let $X$ be an SCPS. Then $T h\left(\Sigma_{1}^{0}(X)\right)$ is hereditarily undecidable.
Proof. We use the same interpretation scheme as in the previous proof. Let $\left\{c_{n}\right\}$ be the sequence of finitary elements from the definition of SCPS. This time we take as the parameters values $W:=\bigcup_{n} \uparrow c_{n}$ and $V:=\bigcup_{n}\left(\uparrow c_{n} \backslash\left\{c_{n}\right\}\right.$. From the definition of SCPS it is easy to see that $\left\{\uparrow c_{n} \backslash\left\{c_{n}\right\}\right\}$ is a computable sequence of effectively open sets, hence again $V$ and $W$ are effectively open and $W \backslash V=\left\{c_{0}, c_{1}, \ldots\right\}$ ). Moreover, the function $D \mapsto\left\{n \mid c_{n} \in D\right\}$ is again an isomorphism between $\mathcal{D}$ and $\mathcal{E}$.

## $3 \quad m$-Degree of $\operatorname{Th}\left(\Sigma_{1}^{0}(X)\right)$

Here we give precise estimate of the algorithmic complexity of $\operatorname{Th} \Sigma_{1}^{0}(X)$ for some spaces $X$. First we establish a natural upper bound that applies to many locally compact spaces. We need a technical notion related to local compactness. By arithmetically locally compact space (ALCS) we mean a triple ( $X, \beta, \kappa$ ) consisting of a topological space $X$, a numbering $\beta$ of a basis in $X$, and a numbering $\kappa$ of compact sets in $X$ such that any set $\beta_{n}$ is a union of some sets in $\left\{\kappa_{i} \mid i<\omega\right\}$, and the relation $\kappa_{i} \subseteq \bigcup \beta\left[D_{n}\right]$ is arithmetical where $\left\{D_{n}\right\}$ is the canonical numbering of finite subsets of $\omega$. Note that although ALCSs are not automatically locally compact, many popular locally compact spaces may be considered as ALCSs. In particular, the computable $\varphi$-spaces, the finite dimensional Euclidean spaces, and Cantor space are ALCSs (say, for a computable $\varphi$-space $(X, \delta)$ we can set $\kappa_{n}:=\uparrow \delta_{n}$ which is compact; the relation $\kappa_{i} \subseteq \bigcup \beta\left[D_{n}\right]$ in this case is c.e.).

Proposition 1. If $(X, \beta, \kappa)$ is an $A L C S$ then $T h\left(\Sigma_{1}^{0}(X)\right) \leq_{m} \emptyset^{(\omega)}$.
Proof. It suffices to show that the relation $\pi_{i} \subseteq \pi_{j}$ is arithmetical because then the elementary diagram of the numbered structure $\left(\Sigma_{1}^{0}(X) ; \subseteq, \pi\right)$, and hence also $T h\left(\Sigma_{1}^{0}(X)\right)$, is $m$-reducible to $\emptyset^{(\omega)}$.

Obviously, $\pi_{i} \subseteq \pi_{j}$ is equivalent to $\forall n\left(\kappa_{n} \subseteq \pi_{i} \rightarrow \kappa_{n} \subseteq \pi_{j}\right)$, hence it suffices to show that the relation $\kappa_{n} \subseteq \pi_{i}$ is arithmetical. We have $\kappa_{n} \subseteq \pi_{i}$ iff $\kappa_{n} \subseteq$ $\bigcup \beta\left[W_{i}\right]$ iff $\exists m\left(D_{m} \subseteq W_{i} \wedge \kappa_{n} \subseteq \bigcup \beta\left[D_{m}\right]\right)$, by compactness of $\kappa_{n}$. The last relation is arithmetical by the definition of ALCS.

We turn to the precise estimation of $m$-degrees of $T h\left(\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)\right)$ and start with the following lemma.

Lemma 1. Any connected component of an effectively open set in $\mathbb{R}^{n}$ is effectively open.

Proof. Let $U$ be a connected component of $V \in \Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)$ and let $a$ be a rational point in $U$. Then

$$
\begin{aligned}
& U=\bigcup\left\{B(b, r) \mid b \in \mathbb{Q}^{n} \wedge r \in \mathbb{Q}^{+} \wedge \exists a_{1} \cdots a_{m-1} \in \mathbb{Q}^{n} \exists r_{1} \cdots r_{m-1} \in \mathbb{Q}^{+}\right. \\
& \left.\left(\bigwedge_{i=1}^{m} \overline{B\left(a_{i}, r_{i}\right)} \subseteq V \wedge \bigwedge_{i=1}^{m-1}\left(B\left(a_{i}, r_{i}\right) \cap B\left(a_{i+1}, r_{i+1}\right) \neq \emptyset\right) \wedge a \in B\left(a_{1}, r_{1}\right)\right)\right\}
\end{aligned}
$$

where $b=a_{m}$ and $r=r_{m}$. Since we can computably enumerate basic closed balls $\overline{B(b, r)} \subseteq V[8]$ with $b \in \mathbb{Q}^{n}, r \in \mathbb{Q}^{+}, U$ is effectively open.

Now we prove the main technical result of this paper:
Theorem 3. For any $n \geq 1$, $\operatorname{Th}\left(\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)\right) \equiv_{m} \emptyset^{(\omega)}$.
Since the upper bound holds by the previous proposition, we only have to prove the lower bound. Since $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)$ is a distributive lattice, we can use in the definitions not only the symbol of inclusion but also the symbols of Boolean operations and the constants $\emptyset, \mathbb{R}^{n}$. Note that for any $x \in \mathbb{R}^{n}, x$ is computable iff $\mathbb{R}^{n} \backslash\{x\}$ is effectively open, hence we can use (to simplify notation) in our defining formulas the computable points (as complements of sets maximal w.r.t. inclusion among the sets below $\mathbb{R}^{n}$ ). More precisely, we will use the variable $x$ to range over the computable points, the atomic formulas $x \in u$ (where $u$ range as usual through $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)$ ), and we can quantify over $x$.

Proof for $n=1$. Since $\operatorname{Th}(\mathcal{E}) \equiv_{m} \emptyset^{(\omega)}$ by [7], it suffices to $m$-reduce $\operatorname{Th}(\mathcal{E})$ to $\operatorname{Th}\left(\Sigma_{1}^{0}(\mathbb{R})\right)$. For this we again find a definition of the lattice $\mathcal{E}$ in $\Sigma_{1}^{0}(\mathbb{R})$ with parameters but this time the set of parameters itself should be definable.

First we show that the property of being a connected effectively open set is definable without parameters. We start with some auxiliary formulas. Let $\widetilde{C o n}(u)$ be the formula $u \neq \emptyset \wedge \neg \exists v, w(u \subseteq v \cup w \wedge v \cap w=\emptyset \wedge u \cap v \neq \emptyset \wedge u \cap w \neq \emptyset)$ saying that $u$ is "effectively connected" i.e. it cannot be partitioned into two disjoint non-empty effectively open sets. In particular, the connected effectively open sets satisfy this formula. Let $\xi(u, v)$ be the formula
$u \neq \emptyset \wedge v \neq \emptyset \wedge u \cap v=\emptyset \wedge \forall u^{\prime}\left(u^{\prime} \cap v=\emptyset \rightarrow u^{\prime} \subseteq u\right) \wedge \forall v^{\prime}\left(v^{\prime} \cap u=\emptyset \rightarrow v^{\prime} \subseteq v\right)$
saying that $u, v$ are disjoint non-empty effectively open sets such that $u=$ $\operatorname{Int}(\mathbb{R} \backslash v$ ) (where Int is the interior operator) and $v=\operatorname{Int}(\mathbb{R} \backslash u)$.

Let now $C \operatorname{Int}(u):=\widetilde{\operatorname{Con}}(u) \wedge \exists v \neq \emptyset \xi(u, v)$. Then $\Sigma_{1}^{0}(\mathbb{R}) \vDash C \operatorname{Int}(U)$ iff $U$ is a computable interval (i.e. the endpoints of $U$ are computable or infinite) distinct from $\emptyset, \mathbb{R}$. For the nontrivial direction, suppose $\Sigma_{1}^{0}(\mathbb{R}) \models C \operatorname{Int}(U), V$ is a satisfying value for $v$ and let $a:=\inf (U), b:=\sup (U)$. Then $a<b$ and at least one of $a, b$ is finite. For each computable point $p \in V$, sets $U_{1}:=\{x \in U \mid x<p\}$ and $U_{2}:=\{x \in U \mid p<x\}$ are effectively open and partition $U$, hence one of them is empty. From this one easily deduces that $V=(b,+\infty)$ if $a=-\infty$, $V=(-\infty, a)$ if $b=+\infty$, and $V=(-\infty, a) \cup(b,+\infty)$ if $a \neq-\infty, b \neq+\infty$. Since $U, V \in \Sigma_{1}^{0}(\mathbb{R}), a$ and $b$ are computable. It is now clear that $U=(a, b)$, as desired.

Let $\psi(u, x):=x \in u \wedge \forall y \in u \exists w(\operatorname{CInt}(w) \wedge x, y \in w \subseteq u)$ and $\widetilde{C m p}(u, v, x):=$ $u \subseteq v \wedge \psi(u, x) \wedge \forall u^{\prime}\left(x \in u^{\prime} \subseteq v \wedge C \operatorname{Int}\left(u^{\prime}\right) \rightarrow u^{\prime} \subseteq u\right)$. Using Lemma 1 one easily checks that $\Sigma_{1}^{0}(\mathbb{R}) \models \widetilde{C m p}(U, V, x)$ iff $U$ is the connected component of $V$ containing the computable point $x$. Let $\operatorname{Con}(u):=(\exists x \in u) \widetilde{\operatorname{Cmp}}(u, u, x)$. Then $\Sigma_{1}^{0}(\mathbb{R}) \vDash \operatorname{Con}(U)$ iff $U$ is connected, hence the property of being connected is definable.

Let $\operatorname{Cmp}(u, v):=u \subseteq v \wedge \operatorname{Con}(u) \wedge \forall u^{\prime}\left(u^{\prime} \subseteq v \wedge \operatorname{Con}\left(u^{\prime}\right) \wedge u \cap u^{\prime} \neq \emptyset \rightarrow u^{\prime} \subseteq u\right)$. Then $\Sigma_{1}^{0}(\mathbb{R}) \models C m p(U, V)$ iff $U$ is a connected component of $V$. Let $C m p^{*}(u, v)$ be the formula $u \subseteq v \wedge \forall u^{\prime}\left(C m p\left(u^{\prime}, v\right) \rightarrow\left(u^{\prime} \subseteq u \vee u \cap u^{\prime}=\emptyset\right)\right)$ saying that $u$ is the union of some connected components of $v$.

Let $\operatorname{ICmp}(v)$ be the formula $\exists u\left(C m p^{*}(u, v) \wedge \neg \exists w\left(w=v \backslash u \wedge \operatorname{Com}^{*}(w, v)\right)\right)$. Then $\Sigma_{1}^{0}(\mathbb{R}) \models \operatorname{ICmp}(V)$ implies that $V$ has infinitely many connected components. Suppose the contrary, then $V=V_{0} \cup \cdots \cup V_{n}$ for some $n \geq 0$ and pairwise disjoint components $V_{0}, \ldots, V_{n} \in \Sigma_{1}^{0}(\mathbb{R})$. Then any $U$ with $\Sigma_{1}^{0}(\mathbb{R}) \models$ $C m p^{*}(U, V)$ is a union of some of $V_{0}, \ldots, V_{n}$, hence $V \backslash U$ is the finite union of the remaining $V_{i}$, a contradiction.

Finally, let $\theta(u, v)$ be the formula

$$
\xi(u, v) \wedge \operatorname{ICmp}(u) \wedge \operatorname{ICmp}(v) \wedge \forall w(C m p(w, u \cup v) \rightarrow C \operatorname{Int}(w))
$$

Then $\Sigma_{1}^{0}(\mathbb{R}) \quad=\theta(U, V)$ iff both $U, V$ have infinitely many connected components which are computable intervals, $U=\operatorname{Int}(\mathbb{R} \backslash V), V=\operatorname{Int}(\mathbb{R} \backslash U)$, between any $U$-components there is a $V$-component (this means that for any $U$-components $U_{1}<U_{2}$, i.e. $\forall x \in U_{0} \forall y \in U_{1}(x<y)$, there is a $V$-component $V_{1}$ with $U_{1}<V_{1}<U_{2}$ ) and between any $V$-components there is a $U$-component. Indeed, from right to left this is obvious. Conversely, if $\Sigma_{1}^{0}(\mathbb{R}) \models \theta(U, V)$ then the only non-trivial fact to check is that between any $U$-components there is a $V$ component. Suppose the contrary, then the interval $W:=\left(\inf \left(U_{1}\right), \sup \left(U_{2}\right)\right)$ is disjoint with $V$, hence $W \subseteq U$. This is a contradiction because $\sup \left(U_{1}\right) \in W \backslash U$.

Let now $\Sigma_{1}^{0}(\mathbb{R}) \vDash \theta(U, V)$ and let $\left\{q_{0}, q_{1}, \cdots\right\}$ be a computable enumeration of the set $U \cap \mathbb{Q}$ without repetitions. Define the equivalence relation $\sim$ on $\omega$ by: $m \sim n$ iff $q_{m}, q_{n}$ are in the same $U$-component. We claim that this relation is c.e. Indeed, since $m \sim n$ is equivalent to the disjunction of $q_{m} \leq q_{n} \wedge\left[q_{m}, q_{n}\right] \subseteq U$ and $q_{n} \leq q_{m} \wedge\left[q_{n}, q_{m}\right] \subseteq U$, so it suffices to check that the relation $q_{m} \leq$ $q_{n} \wedge\left[q_{m}, q_{n}\right] \subseteq U$ is c.e. We have $U=\bigcup_{i} B_{i}$ for a computable sequence $\left\{B_{i}\right\}$ of basic open balls (i.e., intervals with rational endpoints). Since closed intervals are compact, the relation $q_{m} \leq q_{n} \wedge\left[q_{m}, q_{n}\right] \subseteq U$ is equivalent to

$$
\exists l \exists i_{0}, \ldots, i_{l}\left(q_{m} \in B_{i_{0}} \wedge q_{n} \in B_{i_{l}} \wedge B_{i_{0}} \cap B_{i_{1}} \neq \emptyset \wedge \cdots \wedge B_{i_{l-1}} \cap B_{i_{l}} \neq \emptyset\right) .
$$

Alternately, the last assertion follows from the results in [8]. Therefore, $\sim$ is c.e. It is also co-c.e. because, by the previous paragraph, $m \nsim n$ is equivalent to the disjunction of predicates $q_{m}<q_{n} \wedge \exists r \in V \cap \mathbb{Q}\left(q_{m}<r<q_{n}\right)$ and $q_{n}<q_{m} \wedge \exists r \in V \cap \mathbb{Q}\left(q_{n}<r<q_{m}\right)$.

Therefore, the relation $\sim$ is computable with infinitely many equivalence classes, hence the lattice $\mathcal{E}$ is isomorphic to the lattice $\mathcal{F}$ of c.e. sets closed under
$\sim$. The lattice $\mathcal{F}$ is in turn isomorphic to the lattice $(\mathcal{G} ; \subseteq)$ where $\mathcal{G}:=\{G \mid$ $\left.\Sigma_{1}^{0}(\mathbb{R}) \models C m p^{*}(G, U)\right\}$ consists of effective open subsets of $U$ closed under components, for each $U$ as above.

Now it is easy to define a computable transformation of $\subseteq$-sentences $\phi \mapsto \phi^{*}$ such that $\mathcal{E} \models \phi$ iff $\Sigma_{1}^{0}(\mathbb{R}) \models \phi^{*}$ which gives a desired $m$-reduction of $T h(\mathcal{E})$ to $T h\left(\Sigma_{1}^{0}(\mathbb{R})\right)$. According to the method of interpretation, there is a computable transformation of $\subseteq$-formulas $\phi(\ldots) \mapsto \bar{\phi}(u, v, \ldots)$ that adds two new free variables $u, v$ to those of $\phi$ and has the property that $\mathcal{E} \equiv \phi(\ldots)$ iff $\Sigma_{1}^{0}(\mathbb{R}) \models \bar{\phi}(U, V, \ldots)$ for any values $U, V$ of $u, v$ with $\Sigma_{1}^{0}(\mathbb{R}) \models \theta(U, V)$. Now it suffices to set $\phi^{*}:=\forall u, v(\theta(u, v) \rightarrow \bar{\phi}(u, v))$.

Proof for $n \geq 2$. In this case we $m$-reduce $\operatorname{Th}(\mathbb{N})$ to $\operatorname{Th}\left(\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)\right)$. Let $\operatorname{Cof}$ be the unary relation on $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)$ which is true precisely on the cofinite subsets of $\mathbb{R}^{n}$, then $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right) \models \operatorname{Cof}(U)$ iff $\mathbb{R}^{n} \backslash U$ is a finite set of computable points. We will first use in our reduction of $\operatorname{Th}(\mathbb{N})$ to $\operatorname{Th}\left(\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)\right.$ ) this relation Cof (or, equivalently, the relation Fin on $\Pi_{1}^{0}\left(\mathbb{R}^{n}\right)$ such that $\operatorname{Fin}(A)$ iff $A$ is a finite sets of computable points). At the end of the proof we will show that the relation Cof is definable in $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)$.

Our proof is closely related to that in [4] where a similar reduction was established for the lattice of all closed sets in $\mathbb{R}^{n}$. To keep our notation as close as possible to that of [4], we let our formulas to use, along with the usual variables $U, V, W, G, H, K, T$ ranging over $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)$, the variable $x$ ranging over the computable points of $\mathbb{R}^{n}$, and variables $A, B, C, D, E, F$ ranging over $\Pi_{1}^{0}\left(\mathbb{R}^{n}\right)$. Distinctions with the proof in [4] are mainly caused by the fact that we currently do not know whether the relation of being connected is definable in $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)$ for $n \geq 2$. Instead, we use effective versions of some notions from the previous proof. Note that formulas $\widehat{C o n}$ and $\xi$ have the same meaning as above in any reasonable effective space. Obviously, if effectively open sets $U, V$ are effectively connected (i.e., satisfy $\widetilde{C o n}$ ) and are not disjoint then $U \cup V$ is also effectively connected.

Let $\widetilde{C m p}(u, v)$ be the formula

$$
u \subseteq v \wedge \widetilde{\operatorname{Con}}(u) \wedge \forall u^{\prime}\left(u^{\prime} \subseteq v \wedge \widetilde{\operatorname{Con}}\left(u^{\prime}\right) \wedge u \cap u^{\prime} \neq \emptyset \rightarrow u^{\prime} \subseteq u\right)
$$

saying that $u$ is an effectively connected component (we say "effective component" for brevity) of $v$. Let $\Phi(V)$ be the formula $\forall x \in V \exists U \subseteq V(x \in$ $U \wedge \widetilde{C m p}(U, V))$ saying that any computable point of $V$ belongs to some effective component of $V$. From Lemma 1 and remarks above it follows that if $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right) \models \Phi(V)$ then $V$ is the union of its effective components.

Similar to axiom A6 in [4], for any finite disjoint sets $A, B$ of computable points the following formula is easily seen to be true in $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)$ :

$$
\begin{array}{r}
(A \cup B \subseteq U \wedge \widetilde{\operatorname{Con}}(U)) \rightarrow \exists V \exists W(A \subseteq V \wedge B \subseteq W \wedge \\
V \cap W=\emptyset \wedge V \cup W \subseteq U \wedge \widetilde{\operatorname{Con}}(V) \wedge \widetilde{\operatorname{Con}}(W))
\end{array}
$$

Let now $A \approx_{G} B$ be the ternary relation meaning that $A, B$ are finite disjoint sets of computable points and $G$ is an effectively open set such that

$$
A \cup B \subseteq G \wedge \Phi(G) \wedge \forall H(\widetilde{C m p}(H, G) \rightarrow(|H \cap A|=1 \wedge|H \cap B|=1))
$$

Then $A \approx_{G} B$ implies that $A, B$ are of the same cardinality, $|A|=|B|$. Note that for any finite sets $A, B$ of computable points we have: $|A|=|B|$ iff $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right) \models$ $E q(A, B)$ where $E q(A, B)$ is the formula
$F i n(A) \wedge F i n(B) \wedge \exists C, D, G\left(F \operatorname{in}(C) \wedge F i n(D) \wedge C \approx_{G} D \wedge C=A \backslash B \wedge D=B \backslash A\right)$.
As in [4], it is now straightforward to interpret the structure $\mathbb{N}$ in the structure ( $\Sigma_{1}^{0}\left(\mathbb{R}^{n}\right)$; Cof) without parameters by interpreting natural numbers as the cardinalities of finite sets $A, B$ of computable points (i.e., by taking the quotient-set of all such $A$ under the equivalence relation $E q$ ) and interpreting,$+ \times$ as follows: $|A|+|B|=|C|$ iff $\exists A^{\prime}, B^{\prime}\left(E q\left(A^{\prime}, A\right) \wedge E q\left(B^{\prime}, B\right) \wedge A^{\prime} \cap B^{\prime}=\emptyset \wedge E q\left(A^{\prime} \cup B^{\prime}, C\right)\right.$, and $|A| \times|B|=|C|$ iff
$\exists U \exists F(E q(B, F) \wedge F, C \subseteq U \wedge \forall V(C m p(U, V) \rightarrow(|V \cap F|=1 \wedge|V \cap C|=|A|)))$.
It remains to define the relation $\operatorname{Fin}$ in $\Pi_{1}^{0}\left(\mathbb{R}^{n}\right)$ with parameters ranging through a definable set (similar to the case $n=1$ ). Let $W \sqsubseteq U$ abbreviates the formula $\forall D(\widetilde{C m p}(D, W) \rightarrow \widetilde{C m p}(D, U))$ meaning that any effective $W$-component is an effective $U$-component. Consider the following formulas (denoted below by (a), (b), (c), (d) respectively) with free variables $U, V, W, W_{1}, x$ :

$$
\begin{aligned}
& x \in U \wedge U \cap V=\emptyset \wedge \Phi(U) \wedge \Phi(V) \wedge \Phi(W) \wedge \Phi\left(W_{1}\right), \\
& U \cup V \subseteq W \wedge V \subseteq W_{1} \wedge \forall G\left((\widetilde{C m p}(G, U) \wedge x \notin G) \rightarrow G \subseteq W_{1}\right), \\
& (\forall G) \widetilde{\operatorname{Cmp}(G, W)}(\exists!H)_{\widetilde{C m p}(H, U)}(\exists!K) \widetilde{C m p}(K, V) \\
& (\forall \subseteq G \subseteq G \wedge K \subseteq G), \\
& (\forall G)_{\widetilde{C m p}\left(G, W_{1}\right)}(\exists!H)_{\widetilde{C m p}(H, U)}(\exists!K) \widetilde{C m p}(K, V) \\
& (H \subseteq G \wedge K \subseteq G \wedge x \notin H) .
\end{aligned}
$$

Informally, the effective $W$-components induce a bijection (denoted by $f$ ) between the effective $U$-components and the effective $V$-components while $W_{1}$ induces a bijection (denoted by $g$ ) between the effective $V$-components and the effective $U$-components which do not contain the point $x$. Note that the graphs of these bijections are definable with parameters from the tuple $\mathcal{A}=\left(U, V, W, W_{1}, x\right)$, hence we can use the new function symbols $f_{\mathcal{A}}, g_{\mathcal{A}}$ to denote the bijections. For $T \sqsubseteq U$, we denote by $f_{\mathcal{A}}[T]$ the union of $f_{\mathcal{A}}$-images of effective $T$-components. Set now $s_{\mathcal{A}}:=g_{\mathcal{A}} \circ f_{\mathcal{A}}$ (this is a definable function on the effective $U$-components), and consider the following formulas (the second of which is not formalized completely)
(e) $s_{\mathcal{A}}(D)=s_{\mathcal{A}}(E) \rightarrow D=E$,
(f) $\neg \exists T \sqsubseteq U\left(s_{\mathcal{A}}\right.$ is a bijection on the effective $T$-components).

In particular, formula (f) forbids the finite $s_{\mathcal{A}}$-cicles, hence all the iterates $G, s_{\mathcal{A}}(G), s_{\mathcal{A}}^{2}(G), s_{\mathcal{A}}^{3}(G), \ldots$ of the function $s_{\mathcal{A}}$ started with the "initial" effective $U$-component (i.e., the unique effective component $G$ containing the point $x$ ), are distinct. Let $\phi\left(U, V, W, W_{1}, x\right)$ be the conjunction of the formulas (a)-(f).

Define the unary relation $\widetilde{I s o}$ on $\Pi_{1}^{0}\left(\mathbb{R}^{n}\right)$ (an effective analogue of the relation Iso from [4]) as follows: $\widetilde{\operatorname{Iso}}(A)$ iff $\exists V(A \subseteq V \wedge \Phi(V) \wedge \forall U(\widetilde{C m p}(U, V) \rightarrow \exists!x(x \in$
$A \cap U)))$ and $\forall U(U \cap A \neq \emptyset \rightarrow \exists x(x \in U \cap A))$. This relation is definable, $\widetilde{\operatorname{Iso}}(A)$ implies that any point in $A$ is computable and isolated, any infinite set satisfying $\widetilde{I s o}$ is not bounded, and any finite set of computable points satisfies $\widetilde{I s o}$.

Now it suffices to check that for any $A \in \Pi_{1}^{0}\left(\mathbb{R}^{n}\right), \operatorname{Fin}(A)$ is equivalent to

$$
\begin{aligned}
& \widetilde{I s o}(A) \wedge \forall U, V, W, W_{1}, x\left(\left(\phi\left(U, V, W, W_{1}, x\right) \wedge A \subseteq U \cup V\right) \rightarrow\right. \\
& \exists T \sqsubset U\left(x \in T \wedge A \subseteq T \cup f_{\mathcal{A}}[T]\right) \wedge \\
& \left.(\forall G)_{\widetilde{C m p}(G, T)}\left(x \notin G \rightarrow(\exists H)_{\widetilde{C m p}(H, T)}\left(s_{\mathcal{A}}[H]=G\right)\right)\right) .
\end{aligned}
$$

Indeed, the direction from left to right is obvious. Conversely, if $A$ is infinite and satisfies $\overline{\operatorname{Iso}}(A)$, take $x:=0$ to be the zero-vector in $\mathbb{R}^{n}$ and find a strictly increasing computable unbounded sequence $\left\{r_{k}\right\}$ of positive reals such that the boundaries of all balls $B\left(0, r_{k}\right)$ are disjoint with $A$ (this is possible since the relation "the boundary of $B(0, r)$ is contained in $\mathbb{R}^{n} \backslash A$ " is effectively open in $\mathbb{R})$. The assignments $U:=B\left(0, r_{0}\right) \cup \bigcup_{k>0}\left(B\left(0, r_{2 k}\right) \backslash \overline{B\left(0, r_{2 k-1}\right)}\right), V:=$ $\bigcup_{k}\left(B\left(0, r_{2 k+1}\right) \backslash \overline{B\left(0, r_{2 k}\right)}\right), W:=B\left(0, r_{1}\right) \cup \bigcup_{k>0}\left(B\left(0, r_{2 k+1}\right) \backslash \overline{B\left(0, r_{2 k-1}\right)}\right)$, $W_{1}:=\bigcup_{k}\left(B\left(0, r_{2 k+2}\right) \backslash \overline{B\left(x, r_{2 k}\right)}\right)$ make the defining formula false because $A \subseteq U \cup V$ and $\phi\left(U, V, W, W_{1}, x\right)$ are true but there is no corresponding $T \sqsubset U$ (such bounded sets $T$ do not satisfy the condition $A \subseteq T \cup f_{\mathcal{A}}[T]$ ).

For many domains we have even better definability result:
Theorem 4. For any $X \in\left\{\omega^{\leq \omega}, n^{\leq \omega}, P \omega, \omega_{\perp}^{\omega}, n_{\perp}^{\omega}\right\}$, the lattice $\mathcal{E}$ is definable without parameters in the lattice $\Sigma_{1}^{0}(X)$.

Proof. We give the proof only for the space $P \omega$ but the argument works for the other spaces as well. First we check that the set elements $\uparrow F$, for all finite $F \subseteq \omega$, is definable in $\Sigma_{1}^{0}(P \omega)$ (without parameters). Indeed, the defining formula is $\operatorname{ir}(v):=v \neq 0 \wedge \forall x, y(v \subseteq x \cup y \rightarrow v \subseteq x \vee v \subseteq y)$ which says that $v$ is a non-smallest join-irreducible element. Obviously, any set $\uparrow F$ satisfies this formula. Conversely, let an element $V \in \Sigma_{1}^{0}(P \omega)$ satisfy the formula. Since $V \neq \emptyset$, for some canonically enumerable sequence $\left\{F_{n}\right\}$ of finite sets we have $V=\bigcup_{n} \uparrow F_{n}$. Since the partial order ( $\left\{F_{n} \mid n<\omega\right\} ; \subseteq$ ) is well founded, it has a minimal element $F$. Then of course $\uparrow F \subseteq V$, so it suffices to show that $V \subseteq \uparrow F$. Let $S:=\left\{\uparrow F_{n} \mid F \nsubseteq F_{n}\right\}$. Since $S \in \Sigma_{1}^{0}(P \omega), V \subseteq \uparrow F \cup S$ and $V$ is join-irreducible, it suffices to show that $V \nsubseteq S$. Suppose the contrary, then $F \in S$, so $F_{n} \subseteq F$ for some $n$ with $F \nsubseteq F_{n}$. By the minimality of $F, F_{n}=F$ which is a contradiction.

Now we check that the set $V_{n}:=\{\uparrow F:|F|=n\}$ is definable in $\Sigma_{1}^{0}(P \omega)$ for each $n<\omega$. The sequence $\left\{\phi_{n}(v)\right\}$ of defining formulas is given by induction on $n$ as follows: $\phi_{0}:=\forall x(x \subseteq v)$ (saying that $v$ is the largest element of a lattice), and

$$
\phi_{n+1}:=\operatorname{ir}(v) \wedge \neg \phi_{0}(v) \wedge \cdots \wedge \neg \phi_{n}(v) \wedge \forall x\left(v \subset x \rightarrow \phi_{0}(v) \vee \cdots \vee \phi_{n}(v)\right)
$$

(saying that $v$ is a maximal join-irreducible element among those not in the set defined by the formula $\left.\phi_{0}(v) \vee \cdots \vee \phi_{n}(v)\right)$. By the previous paragraph and induction on $n, \phi_{n}$ defines $V_{n}$ for each $n$.

Now let $U_{n}:=\{S \subseteq \omega: n \leq|S|\}$, so $U_{0}=P \omega$ and $U_{n}=\bigcup V_{n}$. Then the singleton set $\left\{U_{n+1}\right\}$ is clearly defined by the formula

$$
\psi_{n+1}(u):=\forall x\left(\phi_{n+1}(x) \rightarrow x \subseteq u\right) \wedge \neg \exists x\left(\phi_{n}(x) \wedge x \subseteq u\right) .
$$

By the proof of Theorem 2, the lattice $\mathcal{E}$ is isomorphic to the sublattice $\left\{S \mid U_{1} \subseteq S \subseteq U_{2}\right\}$ of $\Sigma_{1}^{0}(P \omega)$, and is thus definable without parameters.

Corollary 1. For any $X \in\left\{\omega \leq \omega, n \leq \omega, P \omega, \omega_{\perp}^{\omega}, n_{\perp}^{\omega}\right\}, \operatorname{Th}\left(\Sigma_{1}^{0}(X)\right) \equiv_{m} \emptyset^{(\omega)}$.
Proof. The upper bound holds by Proposition 1, the lower bound by the previous theorem.

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# On the Executability of Interactive Computation 

Bas Luttik and Fei Yang ${ }^{(\boxtimes)}$<br>Eindhoven University of Technology, Eindhoven, The Netherlands<br>\{s.p.luttik,f.yang\}@tue.nl


#### Abstract

The model of interactive Turing machines (ITMs) has been proposed to characterise which stream translations are interactively computable; the model of reactive Turing machines (RTMs) has been proposed to characterise which behaviours are reactively executable. In this article we provide a comparison of the two models. We show, on the one hand, that the behaviour exhibited by ITMs is reactively executable, and, on the other hand, that the stream translations naturally associated with RTMs are interactively computable. We conclude from these results that the theory of reactive executability subsumes the theory of interactive computability. Inspired by the existing model of ITMs with advice, which provides a model of evolving computation, we also consider RTMs with advice and we establish that a facility of advice considerably upgrades the behavioural expressiveness of RTMs: every countable transition system can be simulated by some RTM with advice up to a fine notion of behavioural equivalence.


## 1 Introduction

According to the Church-Turing thesis, the classical Turing machine model adequately formalises which functions from natural numbers to natural numbers are effectively computable. There is, however, a considerable semantic gap between computing the result of a function applied to a natural number and the way computing systems operate nowadays. Modern computing systems are reactive, they are in continuous interaction with their environment, and their operation is not supposed to terminate. Quite a number of extended models of computation have been proposed in recent decades to study the combination of computation and interaction (see, e.g., the collection in [8]). In this paper we compare interactive Turing machines and reactive Turing machines.

Van Leeuwen and Wiedermann have developed a theory of interactive computation from the stance that an interactive computation can be viewed as a never-ending exchange of symbols between a component and its unpredictable interactive environment [9]. Semantically, this amounts to studying the recognition, generation and translation of infinite streams of symbols. In [10], the notion of interactive Turing machine (ITM) is put forward as a tool to formally characterise which stream translations are interactively computable. The notion is subsequently extended with an (non-computable) advice mechanism in order

[^49]to obtain a non-uniform machine model. Van Leeuwen and Wiedermann argue that the resulting model of interactive Turing machines with advice is as powerful as their model of evolving finite automata, and they conclude from this, on intuitive grounds, that ITMs with advice are adequate to model evolving system such as the Internet [17]. Moreover, in a recent article by Cabessa and Villa it is shown that ITMs with advice are as powerful as the model of interactive evolving recurrent neural networks for computing stream translations [4].

The model of interactive Turing machines focusses on capturing the computational content of sequential interactive behaviour. The included mechanism of interaction is therefore limited to achieving this goal, and does not easily generalise to more than one distributed component, nor does it allow for more fine-grained considerations of the behaviour of reactive systems. The behavioural theory of reactive systems, on the other hand, has focussed on aspects of modelling, specification and verification (see, e.g., [1]).

To integrate computability theory and the behavioural theory of reactive systems, the notion of reactive Turing machine (RTM) has been proposed in [2,3]. It extends Turing machines with concurrency-style interaction. Semantically, the operational behaviour of an RTM is given by a transition system. From this transition system one may extract a set of computations, or stream translations, but a more refined analysis is also possible. In fact, to study the effect of interaction of multiple components many refined notions of behavioural equivalence have been developed in the concurrency theory literature [7]. The notion of RTM gives rise to a general theory of executability: a transition system is executable (usually up to some preferred notion of behavioural equivalence) if there exists an RTM that has the transition system as its semantics. (We refer to [3] for more elaborate motivation of the notion of RTM.)

The aim of this paper is to make a connection between the theory of interactive computability and the theory of reactive systems, providing a comparison of the models of ITMs and RTMs in both their semantic domains. We shall first, in Sect. 2, recapitulate both models. Then, in Sect. 3 we present a transition-system semantics for ITMs; the transition system associated with an ITM is executable up to a fine notion of behavioural equivalence. In Sect. 4 we shall identify a subclass of RTMs that can be considered suitable for stream translation, and prove that the stream translation associated with an RTM in this subclass is interactively computable. In Sect. 5 we consider an extension of RTMs with an advice mechanism adapted from the advice mechanism considered for ITMs. RTMs with advice can execute every countable transition system, at the cost of introducing divergence in the computation. The paper ends with a conclusion in Sect. 6.

A full version of this paper is available as [14].

## 2 Preliminaries

### 2.1 The Theory of Interactive Computation

In [12], van Leeuwen and Wiedermann present an analysis of interactive computation on the basis of a component $C$ (thought to behave according to a
deterministic program) interacting with an unpredictable environment $E$. They discuss the consequences of a few general postulates pertaining to the behaviour and interaction of $C$ and $E$ for interactive recognition, interactive generation and interactive translation. In their analysis, the component $C$ acts as a stream transducer, transforming an infinite input stream of data symbols from $\Sigma=\{0,1\}$ presented by $E$ at its input port into an infinite output stream of symbols from $\Sigma$ produced at its output port. Henceforth, by an $\omega$-translation we mean a mapping $\phi: \Sigma^{\omega} \rightarrow \Sigma^{\omega}$ (with $\Sigma^{\omega}$ denoting the set of streams, i.e., infinite sequences, over $\Sigma)$.

Interactive computation is a step-wise process. It is not required that the environment offers a symbol in every step, nor that the component produces a symbol in every step. For the purpose of modelling components, however, it is convenient to record that nothing is offered or produced. The symbol $\lambda$ is used to indicate the situation that no symbol is offered at the input port or produced at the output port, and we let $\Sigma_{\lambda}=\Sigma \cup\{\lambda\}$. It is assumed that when $E$ offers a non- $\lambda$ symbol in some step, then the component $C$ produces a non- $\lambda$ symbol at its output port within finitely many steps, and vice versa; this assumption is referred to as the interactiveness (or finite delay) condition in the work of van Leeuwen and Wiedermann.

In order to formally define which $\omega$-translations are interactively computable by a computational device, van Leeuwen and Wiedermann proposed the notion of interactive Turing machine [10,11]. It extends the classical notion of Turing machine with an input port and an output port, through which it exchanges an infinite, never ending stream of data symbols with its environment. Interactive Turing machines use a two-way infinite tape as memory on which they can write symbols from some presupposed set $\mathcal{D}_{\square}$ of tape symbols, not necessarily disjoint from $\Sigma$ and including the special $\square$ symbol to denote an empty tape cell. Our formal definition below is adapted from [16] (but we leave out the distinction between internal and external states).

Definition 1. $A$ (deterministic) interactive Turing machine (ITM) with a single work tape is a triple $\mathcal{I}=\left(Q, \longrightarrow \mathcal{I}, q_{\text {in }}\right)$, where

1. $Q$ is its set of states;
2. $\longrightarrow_{\mathcal{I}}: Q \times \mathcal{D}_{\square} \times \Sigma_{\lambda} \rightarrow Q \times \mathcal{D}_{\square} \times\{L, R\} \times \Sigma_{\lambda}$ is a transition function; and
3. $q_{\text {in }} \in Q$ is its initial state.

The contents of the tape of an ITM may be represented by an element of $\left(\mathcal{D}_{\square}\right)^{*}$. We denote by $\check{\mathcal{D}_{\square}}=\left\{\check{d} \mid d \in \mathcal{D}_{\square}\right\}$ the set of marked symbols; a tape instance is a sequence $\delta \in\left(\mathcal{D}_{\square} \cup \mathcal{D}_{\square}\right)^{*}$ such that $\delta$ contains exactly one element of $\mathcal{\mathcal { D } _ { \square }}$. The marker indicates the position of the tape head.

A computation of an $\operatorname{ITM} \mathcal{I}=\left(Q, \longrightarrow_{\mathcal{I}}, q_{i n}\right)$ is an infinite sequence of transitions

$$
\begin{equation*}
\left(q_{i n}, \check{\square}\right)=\left(q_{0}, \delta_{0}\right) \xrightarrow{i_{0} / o_{0}} \mathcal{I}\left(q_{1}, \delta_{1}\right) \xrightarrow{i_{1} / o_{1}} \mathcal{I} \cdots\left(q_{k}, \delta_{k}\right) \xrightarrow{i_{k} / o_{k}} \mathcal{I} \cdots . \tag{1}
\end{equation*}
$$

The input stream associated with the computation in (1) is obtained from $i_{0}, i_{1}, \ldots$ by omitting all occurrences of $\lambda$, and the output stream associated
with the computation in (1) is obtained from $o_{0}, o_{1}, \ldots$ by omitting all occurrences of $\lambda$. A pair $(\boldsymbol{x}, \boldsymbol{y}) \in \Sigma^{\omega} \times \Sigma^{\omega}$ is an interaction pair associated with $\mathcal{I}$ if there exists a computation of $\mathcal{I}$ with $\boldsymbol{x}$ as input stream and $\boldsymbol{y}$ as output stream. The set of all interaction pairs associated with an ITM $\mathcal{I}$ is called its interactive behaviour. (In Sect. 3 we shall present a more refined view on its behaviour when we associate with every ITM a transition system.) The computation in (1) is interactive if, for all $k \in \mathbb{N}$, if $i_{k} \neq \lambda$, then there exists $\ell \geq k$ such that $o_{\ell} \neq \lambda$. The computation in (1) is input-active if $i_{k} \neq \lambda$ for all $k \in \mathbb{N}$.

An ITM satisfies the interactiveness condition if all its computations are interactive. Clearly, if a deterministic ITM $\mathcal{I}$ satisfies the interactiveness condition, then its interactive behaviour is total, in the sense that for every $\boldsymbol{x} \in \Sigma^{\omega}$ there is at least one $\boldsymbol{y} \in \Sigma^{\omega}$ such that $(\boldsymbol{x}, \boldsymbol{y})$ is an interaction pair of $\mathcal{I}$. By confining our attention to the input-active computations-which, in the terminology of [12], corresponds to adopting the full environmental activity postulate - , we may then associate with every such ITM an $\omega$-translation: we say that ITM $\mathcal{I}$ produces $\boldsymbol{y}$ on input $\boldsymbol{x}$ if $(\boldsymbol{x}, \boldsymbol{y})$ is the interaction pair associated with an input-active computation of $\mathcal{I}$.

Definition 2. An $\omega$-translation $\phi: \Sigma^{\omega} \rightarrow \Sigma^{\omega}$ is interactively computable if there exists a deterministic ITM that satisfying the interactiveness condition that produces $\phi(\boldsymbol{x})$ on input $\boldsymbol{x}$ for all $\boldsymbol{x} \in \Sigma^{\omega}$.

In [12] a criterion of the interactively computable $\omega$-translations is presented by using limit-continuous functions (a formal definition is included in [14]).

Theorem 1. A total $\omega$-translation is interactively computable iff it is limitcontinuous.

### 2.2 The Theory of Executability

The theory of executability combines computation and concurrency-style interaction in such a way that both are treated on equal footing; thus, an integration of computability and concurrency theory is realised.

The transition system is the central notion in the mathematical theory of discrete-event behaviour. It is parameterised by a set $\mathcal{A}$ of action symbols, denoting the observable events of a system. We extend $\mathcal{A}$ with a special symbol $\tau$, which intuitively denotes unobservable internal activity. We shall abbreviate $\mathcal{A} \cup\{\tau\}$ by $\mathcal{A}_{\tau}$.

Definition 3. $A n \mathcal{A}_{\tau}$-labelled transition system $\mathcal{T}$ is a triple $(\mathcal{S}, \longrightarrow, \uparrow)$, where,

1. $\mathcal{S}$ is a set of states,
2. $\longrightarrow \subseteq \mathcal{S} \times \mathcal{A}_{\tau} \times \mathcal{S}$ is an $\mathcal{A}_{\tau}$-labelled transition relation,
3. $\uparrow \in \mathcal{S}$ is the initial state.

Transition systems can be used to give semantics to programming languages and process calculi. The standard method is to first associate with every program or process expression a transition system (its operational semantics), and then
consider programs and process expressions modulo one of the many behavioural equivalences on transition systems that have been studied in the literature. In this paper, we shall use the notion of (divergence-preserving) branching bisimilarity [5,6], which is the finest behavioural equivalence in van Glabbeek's linear time - branching time spectrum [7] that abstracts from internal computation steps (represented in the transition system by transitions labelled with $\tau$ ). We adopt the notation $\overleftrightarrow{\mathrm{b}}_{\mathrm{b}}$ for divergence-preserving branching bisimilarity, and $\overleftrightarrow{\mathrm{b}}_{\mathrm{b}}$ for the divergence-insensitive version (see [14] for a formal definition).

The notion of reactive Turing machine (RTM) was put forward in [3] to mathematically characterise which behaviour is executable by a conventional computing system. We recall the definition of RTMs and the ensued notion of executable transition system.

Definition 4. $A$ reactive Turing machine $(R T M) \mathcal{M}$ is a triple $(\mathcal{S}, \longrightarrow, \uparrow)$, where

1. $\mathcal{S}$ is a finite set of states,
2. $\longrightarrow \subseteq \mathcal{S} \times \mathcal{D}_{\square} \times \mathcal{A}_{\tau} \times \mathcal{D}_{\square} \times\{L, R\} \times \mathcal{S}$ is a $\left(\mathcal{D}_{\square} \times \mathcal{A}_{\tau} \times \mathcal{D}_{\square} \times\{L, R\}\right)$-labelled transition relation (we write $s \xrightarrow{a[d / e] M}$ t for $(s, d, a, e, M, t) \in \longrightarrow)$,
3. $\uparrow \in \mathcal{S}$ is a distinguished initial state.

Intuitively, the meaning of a transition $s \xrightarrow{a[d / e] M} t$ is that whenever $\mathcal{M}$ is in state $s$, and $d$ is the symbol currently read by the tape head, then it may execute the action $a$, write symbol $e$ on the tape (replacing $d$ ), move the read/write head one position to the left or the right on the tape, and then end up in state $t$.

To formalise the intuitive understanding of the operational behaviour of RTMs, we associate with every RTM $\mathcal{M}$ an $\mathcal{A}_{\tau}$-labelled transition system $\mathcal{T}(\mathcal{M})$. The states of $\mathcal{T}(\mathcal{M})$ are the configurations of $\mathcal{M}$, pairs consisting of a state and a tape instance.

Definition 5. Let $\mathcal{M}=(\mathcal{S}, \longrightarrow, \uparrow)$ be an RTM. The transition system $\mathcal{T}(\mathcal{M})$ associated with $\mathcal{M}$ is defined as follows:

1. its set of states consists of the set of all configurations of $\mathcal{M}$;
2. its transition relation $\longrightarrow$ is the least relation satisfying, for all $a \in \mathcal{A}_{\tau}, d, e \in$ $\mathcal{D}_{\square}$ and $\delta_{L}, \delta_{R} \in \mathcal{D}_{\square}^{*}:$
$-\left(s, \delta_{L} \check{d} \delta_{R}\right) \xrightarrow{a}\left(t, \delta_{L}<e \delta_{R}\right)$ iff $s \xrightarrow{a[d / e] L} t$, and
$-\left(s, \delta_{L} \check{d} \delta_{R}\right) \xrightarrow{a}\left(t, \delta_{L} e^{>} \delta_{R}\right)$ iff $s \xrightarrow{a[d / e] R} t$
( $\delta_{L}{ }^{<}$is obtained from $\delta_{L}$ by placing the tape head marker on the right-most symbol in $\delta_{L}$, and ${ }^{>} \delta_{R}$ is obtained analogously from $\delta_{R}$ );
3. its initial state is the configuration ( $\uparrow, \square \square)$.

Turing introduced his machines to define the notion of effectively computable function in [15]. By analogy, we have a notion of effectively executable behaviour [3].

Definition 6. A transition system is executable if it is the transition system associated with some RTM.

## 3 Executability of Interactive Turing Machines

In this section we associate a transition system with every ITM, and then prove that it is executable modulo divergence-preserving branching bisimilarity. It is convenient to consider input and output as separate actions in the transition system associated with an ITM. We denote by ?i the action of inputting the symbol $i \in \Sigma$, and by $!o$ the action of outputting the symbol $o \in \Sigma$.

Definition 7. Let $\mathcal{I}=\left(Q, \longrightarrow_{\mathcal{I}}, q_{\text {in }}\right)$ be an ITM. The transition system $\mathcal{T}(\mathcal{I})$ associated with $\mathcal{I}$ is defined as follows:

1. its set of states is the set $\left\{(s, \delta) \mid s \in Q \cup\left\{s_{o} \mid o \in \Sigma_{\lambda}, s \in\right.\right.$ $Q\}, \delta$ is a tape instance $\}$;
2. its transition relation $\longrightarrow$ is the least relation satisfying, for all $i, o \in \Sigma_{\lambda}$, $d, e \in \mathcal{D}_{\square}$, and $\delta_{L}, \delta_{R} \in \mathcal{D}_{\square}^{*}:$
$-\left(s, \delta_{L} \check{d} \delta_{R}\right) \xrightarrow{? i}\left(t_{o}, \delta_{L}{ }^{<} e \delta_{R}\right)$ iff $(s, d, i) \longrightarrow_{\mathcal{I}}(t, e, L, o)$ and $i \in \Sigma$,
$-\left(s, \delta_{L} \check{d} \delta_{R}\right) \xrightarrow{? i}\left(t_{o}, \delta_{L} e^{>} \delta_{R}\right)$ iff $(s, d, i) \longrightarrow \mathcal{I}(t, e, R, o)$ and $i \in \Sigma$,
$-\left(s, \delta_{L} \check{d} \delta_{R}\right) \xrightarrow{\tau}\left(t_{o}, \delta_{L}^{<} e \delta_{R}\right)$ iff $(s, d, i) \longrightarrow \mathcal{I}(t, e, L, o)$ and $i=\lambda$,
$-\left(s, \delta_{L} \check{d} \delta_{R}\right) \xrightarrow{\tau}\left(t_{o}, \delta_{L} e^{>} \delta_{R}\right)$ iff $(s, d, i) \longrightarrow \mathcal{I}(t, e, R, o)$ and $i=\lambda$,
$-\left(s_{o}, \delta\right) \xrightarrow{!o}(s, \delta)$ iff $o \in \Sigma$, and $\left(s_{o}, \delta\right) \xrightarrow{\tau}(s, \delta)$ iff $o=\lambda$.
3. its initial state is the configuration ( $\left.q_{\text {in }}, \square\right)$.

The following theorem shows that every transition systems associated with an ITM can be simulated by an RTM. (A proof of the theorem is included in [14].)
Theorem 2. For every ITM $\mathcal{I}$ there exists an RTM $\mathcal{M}$, such that $\mathcal{T}(\mathcal{I}) \overleftrightarrow{{ }_{b}}{ }_{b}^{\Delta}$ $\mathcal{T}(\mathcal{M})$.

As a consequence we have the following corollary.
Corollary 1. The transition system associated with an ITM is executable modulo divergence-preserving branching bisimilarity.

## 4 Executable $\boldsymbol{\omega}$-Translations

Recall that an $\omega$-translation is defined to be interactively computable if, and only if, it can be realised by an ITM. RTMs are designed for exhibiting the expressive power of executable transition systems, rather than $\omega$-translations, and not every RTM naturally has an $\omega$-translation associated with it. Imposing some restrictions on the formalism of RTMs, however, we shall define a subclass of RTMs with which an $\omega$-translation is naturally associated. The $\omega$-translation realised by such an RTM is then called executable, and we shall establish that an $\omega$-translation is interactively computable if, and only if, it is executable.

By analogy to the systems described in the theory of interactive computation, we let the RTMs for $\omega$-translations execute in steps, in such a way that with every step a pair of input and output actions can be associated. With every infinite computation of the RTM we can then associate a interaction pair, and the RTM will thus give rise to an $\omega$-translation.

Definition 8. $\operatorname{Let} \mathcal{A}_{\tau}=\{?,!,!o \mid i, o \in\{0,1\}\} \cup\{\tau\}$, and let $\mathcal{M}=(\mathcal{S}, \longrightarrow, \uparrow)$ be an $R T M$ with $\mathcal{A}_{\tau}$ as its set of labels. Then $\mathcal{M}$ is an RTM for $\omega$-translation if it satisfies the following properties:

1. the set of states $\mathcal{S}$ is partitioned into disjoint sets $I$ of input states and $E$ of execution states, i.e., $\mathcal{S}=I \cup E$ and $I \cap E=\emptyset$;
2. the initial state $\uparrow$ is an input state, i.e., $\uparrow \in I$;
3. for a transition $s \xrightarrow{a[d / e] M} t$, if $s \in I$, then $a \in\{? 0, ? 1\}$ and $t \in E$; if $s \in E$, then $a \in\{!0,!1, \tau\}$ and $t \in I$; and
4. for all $(s, d) \in E \times \mathcal{D}_{\square}$, there is at most one transition of the form $s \xrightarrow{a[d / e] M} t$; and
5. for all $(s, d) \in I \times \mathcal{D}_{\square}$, there are exactly two transitions of the form $s \xrightarrow{a[d / e] M} t$, one with $a=? 0$ and one with $a=$ ? 1 .

In the following lemma we establish some properties of the transition system associated with an RTM for $\omega$-translation. (See [14] for a proof of the lemma.)

Lemma 1. Let $\mathcal{M}$ be an RTM for $\omega$-translation. Then $\mathcal{T}(\mathcal{M})=\left(\mathcal{S}_{\mathcal{M}}, \longrightarrow \mathcal{M}, \uparrow_{\mathcal{M}}\right)$ satisfies the following properties:

1. (Alternation). The set of states $\mathcal{S}_{\mathcal{M}}$ is partitioned into a set of input states $I_{\mathcal{M}}$ and $a$ set of output states $E_{\mathcal{M}}$, i.e., $\mathcal{S}_{\mathcal{M}}=I_{\mathcal{M}} \cup E_{\mathcal{M}}$ and $I_{\mathcal{M}} \cap E_{\mathcal{M}}=\emptyset$. For every transition $s \xrightarrow{a} s^{\prime}$, if $s \in I_{\mathcal{M}}$, then $a \in\{? 0, ? 1\}$ and $s^{\prime} \in E_{\mathcal{M}}$; if $s \in E_{\mathcal{M}}$, then $a \in\{!0,!1, \tau\}$ and $s^{\prime} \in I_{\mathcal{M}}$.
2. (Unambiguity). For every $s \in E_{\mathcal{M}}$, there is exactly one outgoing transition $s \xrightarrow{a} s^{\prime}$ with $a \in\{!0,!1, \tau\}$.
3. (Totality). For every $s \in I_{\mathcal{M}}$, there are exactly two outgoing transitions, labelled with ?0 and ?1, respectively.

We call a transition system that satisfies the conditions of Lemma 1 an $i / o$ transition system. Moreover, by analogy to the interactiveness condition for ITMs, we impose an interactiveness condition on RTMs for $\omega$-translation.

Definition 9. An i/o transition system is interactive, if for every $s \in \mathcal{S}$ and $s \xrightarrow{? i} s_{0}$ with $i \in\{0,1\}$, and for every sequence $s_{0} \longrightarrow s_{1} \longrightarrow \cdots$, there exists a natural number $k$, such that $s_{k} \xrightarrow{!o} s_{k+1}$ with $o \in\{0,1\}$.

An RTM for $\omega$-translation is interactive if the associated $i / o$ transition system is.

We define the $\omega$-translation realized by an RTM by defining the $\omega$-translation realized by the i/o transition system associated with it. Let $\mathcal{T}=(\mathcal{S}, \longrightarrow, \uparrow)$ be an i/o transition system, let $s \in \mathcal{S}$, and let $\sigma \in \mathcal{A}^{\omega}$, say $\sigma=a_{0}, a_{1}, \ldots$; we write $s \xrightarrow{\sigma}$ if there exist $s_{0}, s_{0}^{\prime}, s_{1}, s_{1}^{\prime}, \ldots \in \mathcal{S}$ such that $s=s_{0}$, and $s_{i} \longrightarrow{ }^{*} s_{i}^{\prime} \xrightarrow{a_{i}} s_{i+1}$ for all $i \geq 0$. ( $\mathrm{By} \longrightarrow^{*}$ we denote the reflexive-transitive closure of the relation $\xrightarrow{\tau}$.) If $\sigma \in \mathcal{A}^{\omega}$ and $s \xrightarrow{\sigma}$, then $\sigma$ is a weak infinite trace from $s$. We denote by $\operatorname{Tr}_{w}^{\infty}(s)$ the set of weak infinite traces from $s$.

Definition 10. Let $\mathcal{T}$ be an $i / o$ transition system, and $s_{0}$ be the initial state. For $\sigma \in \operatorname{Tr}_{w}^{\infty}\left(s_{0}\right)$, the input stream realised by $\sigma$ is the stream $\boldsymbol{x} \in \Sigma^{\omega}$ such that $\boldsymbol{x}=x_{1} x_{2} \ldots$, where $x_{j}=i$ if ? i is the $j$-th input action in $\sigma$, and similarly for the output stream realized by $\sigma$. We say that $\mathcal{T}$ realizes $\omega$-translation $\phi: \Sigma^{\omega} \rightarrow \Sigma^{\omega}$ iff, for every $\boldsymbol{x} \in \Sigma^{\omega}$, there exists a trace $\sigma \in \operatorname{Tr}_{w}^{\infty}\left(s_{0}\right)$ with $\boldsymbol{x}$ as its input stream, and for every such trace, its output stream is $\boldsymbol{y}=\phi(\boldsymbol{x})$.

We can now define when an $\omega$-translation is executable.
Definition 11. An $\omega$-translation is executable if it can be realized by an executable i/o transition system.

The following lemma establishes that an $\omega$-translation can be associated with every interactive i/o transition system.

Lemma 2. If an $i / o$ transition system is interactive, then it realises an $\omega$ translation.

Moreover, we have the following theorem. (A proof can be found in [14].)
Theorem 3. An $\omega$-translation is executable iff it is a limit-continuous total function.

By Theorem 1, we have the following corollary.
Corollary 2. An $\omega$-translation is executable iff it is interactively computable.
Therefore, the classes of computable limit-continuous functions, interactively computable $\omega$-translations and executable $\omega$-translations coincide.

## 5 Advice

In [10], the computational power of evolving interactive systems is studied using ITMs. Particularly, a mechanism called advice function is introduced to enhance the computational power of an ITM. In this way, the insertion of external information into the course of a computation is allowed, which leads to a non-uniform operation. In this section, we introduce the notion of advice as a process in parallel composition with an RTM, and show that advice processes indeed give the systems more expressive power.

In this section, we consider advices as functions over natural numbers. In order to record a number on the tape, a natural number $n$ is encoded by a sequence $n$ " 1 "s ending with a " 0 ". In [10], the notion of ITM with advice is defined as follows.

Definition 12. An advice function is a function $f: \mathbb{N} \rightarrow \mathbb{N}$. An ITM with advice (ITM/A) is equipped with a separate advice tape and a distinguished advice state. By writing the value of the argument $x$ on the advice tape and by entering into the advice state, the value of $f(x)$ will appear on the advice tape in a single step. By this action, the original contents of the advice tape is completely overwritten.

Here we do not put the restriction on the length of the advice function as in [12], since it does not make a difference in the issue of computability, and we are not yet interested in the issue of complexity. It is obvious that ITMs with uncomputable advice functions cannot be simulated by any RTM, as uncomputable advice function cannot be evaluated by the mechanism of RTMs. As an extension, we equip RTMs with advice processes which enable the simulation of ITM/As.

An advice process $A_{f}$ is designed to compute the function $f$, and can only interact with a certain RTM $\mathcal{M}$. As an advice function is not necessarily computable, we cannot associate with every advice process an executable transition system. An RTM $\mathcal{M}$ communicates with $A_{f}$ as follows: when it needs to get the result of $f(i)$, it enters a special control state $a_{f}$, and starts to send a sequence of $i$ " 1 " s and a " 0 ", which is already written on the tape, to the channel $\overline{i n}$, and then, it receives the result sequence $f(i)$ " 1 "s and a " 0 " from out channel, and write them on the tape. This procedure ends up with another control state. We can model an advice process as follows.

Definition 13. Let $f: \mathbb{N} \rightarrow \mathbb{N}$ be a function, $A_{f}$ is an advice process for $f$ with transition system $\mathcal{T}\left(A_{f}\right)=(\mathcal{S}, \rightarrow, \uparrow)$, where

1. $\mathcal{S}=\left\{s_{i} \mid i=0,1,2, \ldots\right\} \cup\left\{t_{i} \mid i=0,1,2, \ldots\right\}$, and
2. $s_{i} \xrightarrow{i n ? 1} s_{i+1}, i=0,1,2 \ldots \quad s_{i} \xrightarrow{i n ? 0} t_{f(i)}, i=1,2 \ldots$
$t_{i} \xrightarrow{\text { out!1 }} t_{i-1}, i=1,2 \ldots \quad t_{0} \xrightarrow{\text { out! } 0} s_{0}$
3. $\uparrow=s_{0}$.

The behaviour of $A_{f}$ is deterministic. It receives a sequence of $i$ " 1 "s from the channel $i n$, followed by a " 0 " symbol, indicating the end of the sequence, and then, it produces $f(i)$ " 1 "s to the channel out, also followed by a " 0 " symbol. This procedure is repeated indefinitely.

The parallel composition of an RTM $\mathcal{M}$ and an advice process $A_{f}$, we write as $\left[\mathcal{M} \| A_{f}\right]_{\mathcal{C}}$. The parallel composition is defined in the same way as the parallel composition of two RTMs in [3], where $\mathcal{C}=\{$ in, out $\}$ is the set of restricted names for communication. If $\mathcal{M}$ is an RTM and $A_{f}$ is an advice process, then we call $\left[\mathcal{M} \| A_{f}\right]_{\mathcal{C}}$ a reactive Turing machine with advice (RTM/A).

Note that, since advice functions and advice processes have the same computational power, by Corollary 2, an $\omega$-translation is realisable by an ITM/A if, and only if, it is realisable by an RTM/A.

Let $\mathcal{T}$ be any bounded branching transition system (not necessarily effective). Based on a presupposed encoding of its sets of states and actions and its transition relation, let the advice function $f_{\mathcal{T}}$ be such that for the code of a state it yields the code of the set of all outgoing transitions of that state. It is straightforward to define an RTM that simulates $\mathcal{T}$ with the help of $f_{\mathcal{T}}$. Then we obtain the following result.

Theorem 4. If $\mathcal{T}$ is a boundedly branching labelled transition system, then there exists an $R T M / A\left[\mathcal{M} \| A_{f}\right]_{\mathcal{C}}$ such that $\mathcal{T}\left(\left[\mathcal{M} \| A_{f}\right]_{\mathcal{C}}\right) \overleftrightarrow{ }_{b}^{\Delta} \mathcal{T}$.

If we, instead, let the advice function $f_{\mathcal{T}}$ be such that on the code of a pair of a state $s$ and a natural number $i$ yields the code of the $i$ th outgoing transition of $s$, then we can extend the simulation to transition systems with countable many states and transitions.

Theorem 5. If $T$ is a countable labelled transition system, then there exists an $R T M / A\left[\mathcal{M} \| A_{f}\right]_{\mathcal{C}}$ such that $\mathcal{T}\left(\left[\mathcal{M} \| A_{f}\right]_{\mathcal{C}}\right) \leftrightarrows_{b} T$.

Note that the transition system associated with an RTM/A is boundedly branching. Hence, by Theorem 2 in [13], if a transition system has no divergence up to $\overleftrightarrow{\mathrm{b}}_{\mathrm{b}}^{\Delta}$ and is unboundedly branching up to $\overleftrightarrow{L}_{\mathrm{b}}^{\Delta}$, then it is not executable modulo $\overleftrightarrow{\mathrm{b}}_{\mathrm{b}}^{\Delta}$. It follows that there exist countable unboundedly branching transition systems that cannot be simulated by an RTM/A modulo $\overleftrightarrow{\mathrm{b}}_{\mathrm{b}}^{\Delta}$.

## 6 Conclusion

We have discussed the relationship between two models of computation that take interaction into account. We have established that the model of RTMs subsumes and is more expressive than the model of ITMs when it comes specifying behaviour, and coincides with the model of ITMs when it comes defining $\omega$-translations.

Furthermore, we have shown that RTMs admit an extension with advice that facilitates modelling non-uniform behaviour. In [3] it was established that every effective transition system can be simulated by an RTM. Our result that every countable transition system can be simulated by an RTM with advice further confirms the universal expressiveness of the notion of RTM.

In [16], a complexity theory for interactive computation has been defined on the basis of ITMs and $\omega$-translations. Clearly, such a complexity theory could also be based on the restricted class of RTMs for $\omega$-translation. Such a complexity theory could then further be generalised towards a complexity theory for general executable behaviour.

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# Circuit Satisfiability and Constraint Satisfaction Around Skolem Arithmetic 

Christian Glaßer ${ }^{1}$, Peter Jonsson ${ }^{2}$, and Barnaby Martin ${ }^{3(\bowtie)}$<br>${ }^{1}$ Theoretische Informatik, Julius-Maximilians-Universität, Würzburg, Germany<br>${ }^{2}$ Department of Computer and Information Science, Linköpings Universitet, 58183 Linköping, Sweden<br>${ }^{3}$ School of Science and Technology, Middlesex University, The Burroughs, Hendon, London NW4 4BT, UK<br>barnabymartin@gmail.com


#### Abstract

We study interactions between Skolem Arithmetic and certain classes of Circuit Satisfiability and Constraint Satisfaction Problems (CSPs). We revisit results of Glaßer et al. [16] in the context of CSPs and settle the major open question from that paper, finding a certain satisfiability problem on circuits-involving complement, intersection, union and multiplication - to be decidable. This we prove using the decidability of Skolem Arithmetic. Then we solve a second question left open in [16] by proving a tight upper bound for the similar circuit satisfiability problem involving just intersection, union and multiplication. We continue by studying first-order expansions of Skolem Arithmetic without constants, $(\mathbb{N} ; \times)$, as CSPs. We find already here a rich landscape of problems with non-trivial instances that are in P as well as those that are NP-complete.


## 1 Introduction

Skolem Arithmetic is the weak fragment of first-order arithmetic involving only multiplication. Thoralf Skolem gave a quantifier-elimination technique and argued for decidability of the theory in [27]. However, his proof was rather vague and a robust demonstration was not given of this result until Mostowski [22]. Skolem Arithmetic is somewhat less fashionable than Presburger Arithmetic, which involves only addition, and was proved decidable by Presburger in [25]. Indeed, Mostowski's proof made use of a reduction from Skolem Arithmetic to Presburger Arithmetic through the notion of weak direct powers (an excellent survey on these topics is [3]). The central thread of this paper is putting to work results about Skolem Arithmetic from the past, to solve open and naturally arising problems from today. Many of our results, like that of Mostowski, will rely on the interplay between Skolem and Presburger Arithmetic.

A constraint satisfaction problem (CSP) is a computational problem in which the input consists of a finite set of variables and a finite set of constraints, and

[^50]where the question is whether there exists a mapping from the variables to some fixed domain such that all the constraints are satisfied. When the domain is finite, and arbitrary constraints are permitted in the input, the CSP is NPcomplete. However, when only constraints from a restricted set of relations are allowed in the input, it can be possible to solve the CSP in polynomial time. The set of relations that is allowed to formulate the constraints in the input is often called the constraint language. The question which constraint languages give rise to polynomial-time solvable CSPs has been the topic of intensive research over the past years. It has been conjectured by Feder and Vardi [13] that CSPs for constraint languages over finite domains have a complexity dichotomy: they are either in P or NP-complete. This conjecture remains unsettled, although dichotomy is now known on substantial classes (e.g. structures with domains of size $\leq 3[9,26]$ and smooth digraphs $[2,17])$. Various methods, combinatorial (graph-theoretic), logical and universal-algebraic have been brought to bear on this classification project, with many remarkable consequences. A conjectured delineation for the dichotomy was given in the algebraic language in [10].

By now the literature on infinite-domain CSPs is also beginning to mature. Here the complexity can be much higher (e.g. undecidable) but on natural classes there is often the potential for structured classifications, and this has proved to be the case for reducts of, e.g. the rationals with order [5], the random (Rado) graph [7] and the integers with successor [6]; as well as first-order (fo) expansions of linear program feasibility [4]. Skolem and Presburger Arithmetic represent perfect candidates for continuation in this vein. These natural classes around Skolem and Presburger Arithmetic have the property that their CSPs sit in NP and a topic of recent interest for the second and third authors has been natural CSPs sitting in higher complexity classes.

Meanwhile, a literature existed on satisfiability of circuit problems over sets of integers involving work of the first author [16], itself continuing a line of investigation begun in [29] and pursued in [21,31,32]. The problems in [16] can be seen as variants of certain functional CSPs whose domain is all singleton sets of the non-negative integers and whose relations are set operations of the form: complement, intersection, union, addition and multiplication (the latter two are defined set-wise, e.g. $A \times B:=\{a b: a \in A \wedge b \in B\}$ ). An open problem was the complexity of the problem when the permitted set operators were precisely complement, intersection, union and multiplication. In this paper we resolve that this problem is in fact decidable, indeed in triple exponential space. We prove this result by using the decidability of the theory of Skolem Arithmetic with constants. We take here Skolem Arithmetic to be the non-negative integers with multiplication (and possibly constants). In studying this problem we are able to bring to light existing results of [16] as results about their related CSPs, providing natural examples with interesting super-NP complexities. In addition, we improve one of the upper bounds of [16] to a tight upper bound. This is the circuit satisfiability problem where the permitted set operators are just intersection, union and multiplication, and where we improve the bound from NEXP
to PSPACE. Interestingly, this result does not immediately translate to a similar upper bound for the corresponding functional CSP.

In the second part of the paper, Skolem Arithmetic takes centre stage as we initiate the study of the computational complexity of the CSPs of its reducts, i.e. those constraint languages whose relations have a fo-definition in $(\mathbb{N} ; \times)$. $\operatorname{CSP}(\mathbb{N} ; \times)$ is in P, indeed it is trivial. The object therefore of our early study is its fo-expansions. We show that $\operatorname{CSP}(\mathbb{N} ;+, \neq)$ is $\operatorname{NP}$-complete, as is $\operatorname{CSP}(\mathbb{N} ; \times, c)$ for each $c>1$. We further show that $\operatorname{CSP}(\mathbb{N} ; \times, U)$ is NP-complete when $U$ is any non-empty set of integers greater than 1 such that each has a prime factor $p$, for some prime $p$, but omits the factor $p^{2}$. Clearly, $\operatorname{CSP}(\mathbb{N} ; \times, U)$ is in P (and is trivial) if $U$ contains 0 or 1 . As a counterpoint to our NP-hardness results, we prove that $\operatorname{CSP}(\mathbb{N} ; \times, U)$ is in P whenever there exists $m>1$ so that $U \supseteq\left\{m, m^{2}, m^{3}, \ldots\right\}$.
Related Work. Apart from the research on circuit problems mentioned above there has been work on other variants like circuits over integers [30] and positive natural numbers [8], equivalence problems for circuits [15], functions computed by circuits [24], and equations over sets of natural numbers [18, 19].

## 2 Preliminaries

Let $\mathbb{N}$ be the set of non-negative integers, and let $\mathbb{N}^{+}$be the set of positive integers. For $m \in \mathbb{N}$, let $\operatorname{Div}_{m}$ be the set of factors of $m$. Finally, let $\{\mathbb{N}\}$ be the set of singletons $\{\{x\}: n \in \mathbb{N}\}$. In this paper we use a version of the CSP permitting both relations and functions (and constants). Thus, a constraint language consists of a domain together with functions, relations and constants over that domain. One may thus consider a constraint language to be a firstorder structure. A homomorphism from a constraint language $\Gamma$ to a constraint language $\Delta$, over the same signature, is a function $f$ from the domain of $\Gamma$ to the domain of $\Delta$ that preserves the relations, i.e. if $\left(x_{1}, \ldots, x_{k}\right) \in R^{\Gamma}$, then also $\left(f\left(x_{1}\right), \ldots, f\left(x_{k}\right)\right) \in R^{\Delta}$. A homomorphism from a constraint language to itself is an endomorphism. An endomorphism that also preserves the negations of relations is termed an embedding and a bijective embedding is an automorphism.

A constraint language is a core if all of its endomorphisms are embeddings (equivalently, if the domain is finite, automorphisms). The functional version of the CSP has previously been seen in, e.g., [12]. For a purely functional constraint language, a primitive positive ( pp ) sentence is the existential quantification of a conjunction of term equalities. More generally, and when relations present, we may have positive atoms in this conjunction. The problem $\operatorname{CSP}(\Gamma)$ takes as input a primitive positive sentence $\phi$, and asks whether it is true on $\Gamma$. The problem $\operatorname{CSP}^{c}(\Gamma)$ is similar but allows input constants naming the domain elements. We will allow that the functions involved on $\phi$ be defined on a larger domain than the domain of $\Gamma$. This is rather unheimlich ${ }^{1}$ but it allows the problems of [16] to

[^51]be more readily realised in the vicinity of CSPs. For example, one such typical domain is $\{\mathbb{N}\}$, but we will allow functions such as - (complement), $\cup$ (union) and $\cap$ (intersection) whose domain and range is the set of all subsets of $\mathbb{N}$. We will also employ the operations of set-wise addition $A+B:=\{a+b: a \in A \wedge b \in B\}$ and multiplication $A \times B:=\{a b: a \in A \wedge b \in B\}$.
$\Sigma_{\mathrm{i}}^{\mathrm{P}}, \Pi_{\mathrm{i}}^{\mathrm{P}}$, and $\Delta_{\mathrm{i}}^{\mathrm{P}}$ are levels of the polynomial-time hierarchy, while $\Sigma_{i}, \Pi_{i}$, and $\Delta_{i}$ are levels of the arithmetical hierarchy. Moreover, we use the classes $\mathrm{NP}=\Sigma_{1}^{\mathrm{P}}$, $\operatorname{PSPACE}=\bigcup_{k \geq 1} \operatorname{DSPACE}\left(n^{k}\right)$, and 3EXPSPACE $=\bigcup_{k \geq 1} \operatorname{DSPACE}\left(2^{2^{2^{n^{k}}}}\right)$. Where no SPACE is written explicitly, the complexity classes may be assumed to refer to time. For more on these complexity classes we refer the reader to [23].

For sets $A$ and $B$ we say that $A$ is polynomial-time many-one reducible to $B$, in symbols $A \leq_{\mathrm{m}}^{\mathrm{p}} B$, if there exists a polynomial-time computable function $f$ such that for all $x$ it holds that $(x \in A \Longleftrightarrow f(x) \in B)$. If $f$ is even computable in logarithmic space, then $A$ is logspace many-one reducible to $B$, in symbols $A \leq_{\mathrm{m}}^{\log } B$. $A$ is nondeterministic polynomial-time many-one reducible to $B$, in symbols $A \leq{ }_{\mathrm{m}}^{\mathrm{NP}} B$, if there is a nondeterministic Turing transducer $M$ that runs in polynomial time such that for all $x$ it holds that $x \in A$ if and only if there exists a $y$ computed by $M$ on input $x$ with $y \in B$. The reducibility notions $\leq \leq_{\mathrm{m}}^{\mathrm{p}}$, $\leq_{\mathrm{m}}^{\mathrm{log}}$, and $\leq_{\mathrm{m}}^{\mathrm{NP}}$ are transitive and NP is closed under these reducibilities.

A circuit $C=\left(V, E, g_{C}\right)$ is a finite, non-empty, directed, acyclic multi-graph $(V, E)$ with a specified node $g_{C} \in V$. The graph does not need to be connected and only has multiple edges between two nodes when a binary operator is applied on both sides to a single set (e.g. $A \times A$ ). Let $V=\{1,2, \ldots, n\}$ for some $n \in \mathbb{N}$. The nodes in the graph $(V, E)$ are topologically ordered, i.e., for all $v_{1}, v_{2} \in V$, if $v_{1}<v_{2}$, then there is no path from $v_{2}$ to $v_{1}$. Nodes are also called gates. Nodes with indegree 0 are called input gates and $g_{C}$ is called the output gate. If there is an edge from gate $u$ to gate $v$, then we say that $u$ is a predecessor of $v$ and $v$ is a successor of $u$.

Let $\mathcal{O} \subseteq\left\{\cup, \cap,{ }^{-},+, \times\right\}$. An $\mathcal{O}$-circuit with unassigned input gates $C=$ ( $V, E, g_{C}, \alpha$ ) is a circuit $\left(V, E, g_{C}\right)$ whose gates are labeled by the labeling function $\alpha: V \rightarrow \mathcal{O} \cup \mathbb{N} \cup\{\star\}$ such that the following holds: Each gate has an indegree in $\{0,1,2\}$, gates with indegree 0 have labels from $\mathbb{N} \cup\{\star\}$, gates with indegree 1 have label ${ }^{-}$, and gates with indegree 2 have labels from $\{\cup, \cap,+, \times\}$. Input gates with a label from $\mathbb{N}$ are called assigned (or constant) input gates; input gates with label $\star$ are called unassigned (or variable) input gates. An $\mathcal{O}$-formula is an $\mathcal{O}$-circuit that only contains nodes with outdegree one.

Let $u_{1}<\cdots<u_{n}$ be the unassigned inputs in $C$ and $x_{1}, \ldots, x_{n} \in \mathbb{N}$. By assigning value $x_{i}$ to the input $u_{i}$, we obtain an $\mathcal{O}$-circuit $C\left(x_{1}, \ldots, x_{n}\right)$ whose input gates are all assigned. In this circuit, each gate $g$ computes the following set $I(g)$ : If $g$ is an assigned input gate where $\alpha(g) \neq \star$, then $I(g)=\{\alpha(g)\}$. If $g=u_{k}$ is an unassigned input gate, then $I(g)=\left\{x_{k}\right\}$. If $g$ has label ${ }^{-}$and predecessor $g_{1}$, then $I(g)=\mathbb{N} \backslash I\left(g_{1}\right)$. If $g$ has label $\circ \in\{\cup, \cap,+, \times\}$ and predecessors $g_{1}$ and $g_{2}$, then $I(g)=I\left(g_{1}\right) \circ I\left(g_{2}\right)$. Finally, let $I\left(C\left(x_{1}, \ldots, x_{n}\right)\right)=I\left(g_{C}\right)$ be the set computed by the circuit $C\left(x_{1}, \ldots, x_{n}\right)$.

## Definition 1 (membership, equivalence, and satisfiability problems of circuits and formulas).

Let $\mathcal{O} \subseteq\left\{\cup, \cap,{ }^{-},+, \times\right\}$.
$\operatorname{MC}_{\mathbb{N}}(\mathcal{O})=\{(C, b) \mid C$ is an $\mathcal{O}$-circuit without unassigned inputs and $b \in I(C)\}$
$\mathrm{EC}_{\mathbb{N}}(\mathcal{O})=\left\{\left(C_{1}, C_{2}\right) \mid C_{1}\right.$ and $C_{2}$ are $\mathcal{O}$-circuits without unassigned inputs and we have $\left.I\left(C_{1}\right)=I\left(C_{2}\right)\right\}$
$\mathrm{SC}_{\mathbb{N}}(\mathcal{O})=\left\{(C, b) \mid C\right.$ is an $\mathcal{O}$-circuit with unassigned inputs $u_{1}<\cdots<u_{n}$ and there exist $x_{1}, \ldots, x_{n} \in \mathbb{N}$ such that $\left.b \in I\left(C\left(x_{1}, \ldots, x_{n}\right)\right)\right\}$
$\mathrm{MF}_{\mathbb{N}}(\mathcal{O}), \mathrm{EF}_{\mathbb{N}}(\mathcal{O})$, and $\mathrm{SF}_{\mathbb{N}}(\mathcal{O})$ are the variants that deal with $\mathcal{O}$-formulas instead of $\mathcal{O}$-circuits.

When an $\mathcal{O}$-circuit is used as input for an algorithm, then we use a suitable encoding such that it is possible to verify in deterministic logarithmic space whether a given string encodes a valid circuit.

In Sect. 3, for $i \in \mathbb{N}$, we often identify $\{i\}$ with $i$, where this can not cause a harmful confusion.

## 3 Circuit Satisfiability and Functional CSPs

We investigate the computational complexity of functional CSPs. In many cases we can translate known lower and upper bounds for membership, equivalence, and satisfiability problems of arithmetic circuits $[15,16,21]$ to CSPs. Our main result is the decidability of $\operatorname{SC}_{\mathbb{N}}\left({ }^{-}, \cup, \cap, \times\right)$ and $\operatorname{CSP}^{c}\left(\{\mathbb{N}\} ;{ }^{-}, \cup, \cap, \times\right)$, which solves the main open question of the paper [16]. We emphasise that the domain of $\operatorname{CSP}^{c}\left(\{\mathbb{N}\} ;{ }^{-}, \cup, \cap, \times\right)$ is the set of singletons that we defined as $\{\mathbb{N}\}$ and not, e.g., the set of subsets of all natural numbers. This would be a different CSP. Our unusual definition is motivated by the circuit problems whose relationship to CSPs we wish to formalise.

We start with the observation that the equivalence of arithmetic terms reduces to functional CSPs. This yields several lower bounds for the CSPs.

Proposition 1. For $\mathcal{O} \subseteq\{-, \cup, \cap,+, \times\}$ it holds that $\mathrm{EF}_{\mathbb{N}}(\mathcal{O}) \leq_{\mathrm{m}}^{\log }$ $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \mathcal{O})$.

## Corollary 1.

1. $\operatorname{CSP}^{c}(\{\mathbb{N}\} ;-, \cup, \cap,+)$ and $\operatorname{CSP}^{c}(\{\mathbb{N}\} ;-, \cup, \cap, \times)$ are $\leq_{m}^{\log }$-hard for PSPACE.
2. $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \cup \cap,+), \operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \cup, \cap, \times), \operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \cup,+)$, and
$\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \cup, \times)$ are $\leq_{\mathrm{m}}^{\log }$-hard for $\Pi_{2}^{\mathrm{P}}$.
CSPs with + and $\times$ can express diophantine equations, which implies the Turing-hardness of such CSPs.

Proposition 2. $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ;+, \times)$ is $\leq_{\mathrm{m}}^{\text {log }}$-hard for $\Sigma_{1}, \operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \cup, \cap,+, \times) \in$ $\Sigma_{1}$ and $\operatorname{CSP}^{\mathrm{c}}\left(\{\mathbb{N}\} ;{ }^{-}, \cup, \cap,+, \times\right) \in \Sigma_{2}$.

We now show that the decidability of Skolem arithmetic [14] can be used to decide the satisfiability of arithmetic circuits without + . From this we obtain the decidability of CSPs where exactly one arithmetic operation is forbidden.

Theorem 1. $\operatorname{SC}_{\mathbb{N}}(-, \cup, \cap, \times), \operatorname{CSP}^{c}\left(\{\mathbb{N}\} ;^{-}, \cup, \cap, \times\right)$ and $\operatorname{CSP}^{\mathrm{c}}\left(\{\mathbb{N}\} ;{ }^{-}, \cup, \cap,+\right)$ are in 3EXPSPACE.

The following proposition transfers the NP-hardness from satisfiability problems for arithmetic circuits to $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \times)$ and $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ;+)$.
Proposition 3. $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \times)$ and $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ;+)$ are $\leq_{\mathrm{m}}^{\text {log }}-$ hard for NP .
The remaining results in this section show that certain functional CSPs belong to NP. This needs non-trivial arguments of the form: If a CSP can be satisfied, then it can be satisfied even with small values. These arguments are provided by the known results that integer programs, existential Presburger arithmetic, and existential Skolem arithmetic are decidable in NP.

Proposition 4. $\operatorname{CSP}^{\mathrm{c}}\left(\{\mathbb{N}\} ;{ }^{-}, \cap, \cup\right)$ is $\leq_{\mathrm{m}}^{\text {log }}$-complete for NP .
Proposition 5. $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ;+) \in \mathrm{NP}$.
Proposition 6. $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \cap,+) \leq_{\mathrm{m}}^{\mathrm{NP}} \operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ;+,=, \neq)$ and $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \cap, \times)$ $\leq_{m}^{\mathbb{N P}} \operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \times,=, \neq)$. Therefore, $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \cap,+), \operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \cap, \times) \in \operatorname{NP}$.

A Second Open Problem from [16]. We now improve another of the upper bounds of [16] to a tight upper bound. Here we have the circuit satisfiability problem where the permitted set operators are just intersection, union and multiplication, where we improve the bound from NEXP to PSPACE.
Theorem 2. $\mathrm{SC}_{\mathbb{N}}(\cup, \cap, \times) \in$ PSPACE.
Table 1 summarizes the results obtained in Sect. 3 and shows open questions. In particular, we would like to improve the gap between the lower and upper bounds for $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \mathcal{O})$, where $\mathcal{O}$ contains $\cup$ and exactly one arithmetic operation (+ or $\times$ ).

## 4 CSPs over Fo-expansions of Skolem Arithmetic

We now commence our exploration of the complexity of CSPs generated from the simplest expansions of $(\mathbb{N} ; \times)$. Abandoning our set-wise definitions, we henceforth use $\times$ to refer to the syntactic multiplication of Skolem Arithmetic (which may additionally carry semantic content). When we wish to refer to multiplication in a purely semantic way, we prefer $\cdot \mathrm{s}$ or $\Pi$. We will consider $\times$ as a ternary relation rather than a binary function. We will never use syntactic $\times$ in a non-standard way, i.e. holding on a triple of integers for which it does not already hold in natural arithmetic.

Table 1. Upper and lower bounds for $\operatorname{CSP}^{\mathrm{C}}(\{\mathbb{N}\} ; \mathcal{O})$. All lower bounds are with respect to $\leq_{m}^{\log }$-reductions.

| $\operatorname{CSP}^{\mathrm{c}}(\{\mathbb{N}\} ; \mathcal{O})$ |  |  |
| :---: | :---: | :---: |
| $\mathcal{O}$ | Lower bound | Upper bound |
| ${ }^{-} \cup \cap+\times$ | $\Sigma_{1}$ | $\Sigma_{2}$ |
| $-\cup \cap+$ | PSPACE | 3EXPSPACE |
| $-\cup \cap \times$ | PSPACE | 3EXPSPACE |
| $-\cup \cap$ | NP | NP |
| $\cup \cap+\times$ | $\Sigma_{1}$ | $\Sigma_{1}$ |
| $\cup \cap+$ | $\Pi_{2}^{\text {P }}$ | 3EXPSPACE |
| $\cup \cap \times$ | $\Pi_{2}^{\mathrm{P}}$ | 3EXPSPACE |
| $\cup+\times$ | $\Sigma_{1}$ | $\Sigma_{1}$ |
| $\cup \quad+$ | $\Pi_{2}^{P}$ | 3EXPSPACE |
| $\cup \times$ | $\Pi_{2}^{\mathrm{P}}$ | 3EXPSPACE |
| $\cap+x$ | $\Sigma_{1}$ | $\Sigma_{1}$ |
| $\cap+$ | NP | NP |
| $\cap \times$ | NP | NP |
| + $\times$ | $\Sigma_{1}$ | $\Sigma_{1}$ |
| + | NP | NP |
| $\times$ | NP | NP |

Proposition 7. Let $\Gamma$ be a finite signature reduct of $(\mathbb{N} ; \times, 1,2, \ldots)$. Then $\operatorname{CSP}(\Gamma)$ is in NP.

Upper Bounds. We continue with polynomial upper bounds. Note that constants are no longer assumed to necessary exist in our structures (in contrast to the situation in Proposition 7).

Lemma 1. Let $U \subseteq N$ be non-empty and $U \cap\{0,1\}=\emptyset$. Then $\operatorname{CSP}(\mathbb{N} ; \times, U)$ is polynomial-time reducible to $\operatorname{CSP}\left(\mathbb{N}^{+} ; \times, U\right)$.

We now borrow the following slight simplification of Lemma 6 from [20].
Lemma 2 (Scalability [20]). Let $\Gamma$ be a finite signature constraint language with domain $\mathbb{R}$, whose relations are quantifier-free definable in,$+ \leq$ and $<$, such that the following holds.

- Every satisfiable instance of $\operatorname{CSP}(\Gamma)$ is satisfied by some rational point.
- For each relation $R \in \Gamma$, it holds that if $\bar{x}:=\left(x_{1}, x_{2}, \ldots, x_{k}\right) \in R$, then $\left(a x_{1}, a x_{2}, \ldots, a x_{k}\right) \in R$ for all $a \in\{y: y \in \mathbb{R}, y \geq 1\}$.
$-\operatorname{CSP}(\Gamma)$ is in P .
Then $\operatorname{CSP}(\Delta)$ is in P , where $\Delta$ is obtained from $\Gamma$ by substituting the domain $\mathbb{R}$ by $\mathbb{Z}$.

Lemma 3. Arbitrarily choose $m>1$ and $U \subseteq \mathbb{N}^{+}$such that $\left\{m, m^{2}, m^{3}, \ldots\right\} \subseteq$ $U$. Then, $\operatorname{CSP}\left(\mathbb{N}^{+} ; \times, U\right)$ is in P .
Proposition 8. Arbitrarily choose $m>1$ and $U \subseteq \mathbb{N}$ such that $\left\{m, m^{2}, m^{3}, \ldots\right\}$ $\subseteq U$. Then, $\operatorname{CSP}(\mathbb{N} ; \times, U)$ is in P .

Cores. We say that an integer $m>1$ has a degree-one factor $p$ if and only if $p$ is a prime such that $p \mid m$ and $p^{2} \bigwedge m$. Let $\operatorname{Div}_{m}$ be the set of divisors of $m$, pp-definable in $(\mathbb{N} ; \times, m)$ by $\exists y x \times y=m$. We can pp-define the relation $\{1\}$ in $\left(\operatorname{Div}_{m} ; \times, m\right)$ since $x=1$ iff $x \times x=x$ (recalling $\left.0 \notin \operatorname{Div}_{m}\right)$. It follows that $\{1, m\}$ are contained in the core of $\left(\operatorname{Div}_{m} ; \times, m\right)$.
Lemma 4. Let $m>1$ be an integer that has a degree-one factor $p$. Then $\left(\operatorname{Div}_{m} ; \times, m\right)$ has a two-element core.

Lemma 5. Let $m$ be an integer that does not have a degree-one factor. Then ( $\operatorname{Div}_{m} ; \times, m$ ) does not have a two-element core.

Lower Bounds. We now move to lower bounds of NP-completeness.
Proposition 9. $\operatorname{CSP}(\mathbb{N} ; \neq, \times)$ is NP -complete.
An operation $t: D^{k} \rightarrow D$ is a weak near-unanimity operation if $t$ is idempotent and satisfies $t(y, x, \ldots, x)=t(x, y, x, \ldots, x)=\cdots=t(x, \ldots, x, y)$.

Theorem 3 [1]. Let $\Gamma$ be a constraint language over a finite set $D$. If $\Gamma$ is a core and does not have a weak near-unanimity polymorphism, then $\operatorname{CSP}(\Gamma)$ is NP-hard.

Lemma 6. Arbitrarily choose an $m>1$ such that $m \neq k^{n}$ for all $k, n>1$ together with a finite set $\{1, m\} \subseteq S \subseteq \mathbb{N} \backslash\{0\}$. If $(S ; \times, m)$ is a core, then $\operatorname{CSP}(S ; \times, m)$ is NP-hard.

Note that the proof of this last lemma is made easier by our assumption that $x$ is a relation and not a function. Were it a function we would need to prove the domain $S$ is closed under it.

Theorem 4. $\operatorname{CSP}(\mathbb{N} ; \times, m)$ is NP -hard for every integer $m>1$.
Theorem 5. Let $U$ be any subset of $\mathbb{N} \backslash\{0,1\}$ so that every $x \in U$ has a degreeone factor. Then $\operatorname{CSP}(\mathbb{N} ; \times, U)$ is NP-hard.

For $x \in \mathbb{N} \backslash\{0,1\}$, define its minimal exponent, $\min -\exp (x)$, to be the smallest $j$ such that $x$ has a factor of $p^{j}$, for some prime $p$, but not a factor of $p^{j+1}$. Thus an integer with a degree-one factor has minimal exponent 1 . Call $x \in \mathbb{N} \backslash\{0,1\}$ square-free if it omits all repeated prime factors. For a set $U \subseteq \mathbb{N} \backslash\{0,1\}$, define its basis, $\operatorname{basis}(U)$ to be the set $\{\min -\exp (x): x \in U\}$.
Lemma 7. Let $U \subseteq \mathbb{N} \backslash\{0,1\}$, so that $\operatorname{basis}(U)$ is finite and $\operatorname{basis}(U) \neq\{1\}$. There is some set $X$ pp-definable in $(\mathbb{N} ; \times, U)$ so that $\operatorname{basis}(X)=\{1\}$.

Theorem 6. Let $U \subseteq \mathbb{N} \backslash\{0,1\}$ be so that $\operatorname{basis}(U)$ is finite. Then $\operatorname{CSP}(\mathbb{N} ; \times, U)$ is NP-complete.

## 5 Final Remarks

There are two major directions in which more work is necessary.
A perfunctory glance at the results of Sect. 3 shows that some of our bounds are not tight, and it would be great to see some natural CSPs in this region manifesting complexities such as PSPACE-complete. It is informative to compare our Table 1 with Table 1 in [16]. Our weird formulation of these CSPs belies the fact there are more natural versions where, for $\mathcal{O} \subseteq\left\{{ }^{-}, \cap, \cup,+, \times\right\}$, we ask about $\operatorname{CSP}(\mathcal{P}(\mathbb{N}) ; \mathcal{O})$, where $\mathcal{P}(\mathbb{N})$ is the power set of $\mathbb{N}$, rather than the somewhat esoteric $\operatorname{CSP}(\{\mathbb{N}\} ; \mathcal{O})$. Indeed, if we replace complement "-" by set difference " $\backslash$ ", these questions could also be phrased for just the finite sets of $\mathcal{P}(\mathbb{N})$ (see recent work [11]).

Meanwhile, the results of Sect. 4 need to be extended to a classification of complexity for all $\operatorname{CSP}(\Gamma)$, where $\Gamma$ is a reduct of Skolem Arithmetic $(\mathbb{N} ; \times)$. We anticipate the first stage is to complete the classification for $\operatorname{CSP}(\mathbb{N} ; \times, U)$ where $U$ is fo-definable in $(\mathbb{N} ; \times)$.

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# The Complexity of Counting Quantifiers on Equality Languages 

Barnaby Martin ${ }^{1(\boxtimes)}$, András Pongrácz ${ }^{2}$, and Michał Wrona ${ }^{3}$<br>${ }^{1}$ School of Science and Technology, Middlesex University, The Burroughs, Hendon, London NW4 4BT, UK<br>barnabymartin@gmail.com<br>${ }^{2}$ Debreceni Egyetem TTK, Algebra és Számelmélet Tanszék, Debrecen PF. 18, 4010, Hungary<br>3 Faculty of Mathematics and Computer Science, Theoretical Computer Science Department, Jagiellonian University, Kraków, Poland


#### Abstract

An equality language is a relational structure with infinite domain whose relations are first-order definable in equality. We classify the extensions of the quantified constraint satisfaction problem over equality languages in which the native existential and universal quantifiers are augmented by some subset of counting quantifiers. In doing this, we find ourselves in various worlds in which dichotomies or trichotomies subsist.


## 1 Introduction

The constraint satisfaction problem $\operatorname{CSP}(\Gamma)$, much studied in artificial intelligence, is known to admit several equivalent formulations, two of the best known of which are the query evaluation of primitive positive (pp) sentences - those involving only existential quantification and conjunction - on $\Gamma$, and the homomorphism problem to $\Gamma$ (see, e.g., [14]). For finite $\Gamma$ the problem $\operatorname{CSP}(\Gamma)$ is NP-complete in general, and a great deal of effort has been expended in classifying its complexity for certain restricted cases. Notably it is conjectured [9,12] that for all fixed finite $\Gamma$, the problem $\operatorname{CSP}(\Gamma)$ is in P or NP-complete. While this has not been settled in general, a number of partial results are known - e.g. over structures of size at most three [8,17] and over smooth digraphs $[2,13]$.

A popular generalisation of the CSP involves considering the query evaluation problem for positive Horn logic - involving only the two quantifiers, $\exists$ and $\forall$, together with conjunction. The resulting quantified constraint satisfaction problems $\operatorname{QCSP}(\Gamma)$ allow for a broader class, used in artificial intelligence to capture non-monotonic reasoning, whose complexities rise to Pspace-completeness.
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Once upon a time, Bodirsky and Kára gave a systematic classification for $\operatorname{CSP}(\Gamma)$, where $\Gamma$ consists of relations first-order (fo) definable in equality, over some countably infinite domain [7]. These so-called equality languages $\Gamma$ displayed dichotomy between those for which $\operatorname{CSP}(\Gamma)$ was in P and those for which it was NP-complete. Pursuing this line of investigation, Bodirsky and Chen gave a trichotomy for $\operatorname{QCSP}(\Gamma)$, where $\Gamma$ is an equality language - each problem being either in P, NP-complete or co-NP-hard [4]. In the conference version of that paper, the trichotomy was claimed to be across P, NP-complete or Pspacecomplete [3], but the proof in the tricky case of $x=y \rightarrow y=z$ was flawed, and so in the journal version this became the weaker co-NP-hard (and in Pspace). The trichotomy is thus imperfect, as most of the co-NP-hard cases are known to be Pspace-complete. Indeed, $x=y \rightarrow y=z$ would be the only open case, if it would to be Pspace-complete [5].
Working Hypothesis. $\operatorname{QCSP}(x=y \rightarrow y=z)$ is Pspace-complete.
Thus the assumption of the working hypothesis would restore the trichotomy to the P, NP-complete or Pspace-complete as stated in [3].

In this paper, we consider the generalisation of the QCSP with counting quantifiers, as pioneered in the recent paper [15]. In [15], the domains of $\Gamma$ were of finite size $n$, so the extant quantifiers $\exists \geq 1=\exists$ and $\exists \geq n=\forall$ were augmented with quantifiers of the form $\exists \geq j$, which allow one to assert the existence of at least $j$ elements such that the ensuing property holds. In the world of infinite domains, it makes sense to permit not only quantification above the finite with $\exists \geq j$, but also quantification below the co-finite with $\forall \geq j$, whose intended meaning is that the property holds for all but (at most) $j$ elements of the domain. Thus, $\forall=\forall \geq 0$. Counting quantifiers have been extensively studied in finite model theory (see $[11,16])$, where the focus is on supplementing the descriptive power of various logics. Of more general interest is the majority quantifier $\exists \geq n / 2$ (on a structure of domain size $n$ ), which sits broadly midway between $\exists$ and $\forall$. Majority quantifiers are studied across diverse fields of logic and have various practical applications, e.g. in cognitive appraisal and voting theory [10,19]. They have also been studied in computational complexity since at least [1] (see also [11]).

We study extensions of $\operatorname{QCSP}(\Gamma)$ in which the input sentence to be evaluated on $\Gamma$ remains positive conjunctive in its quantifier-free part, but is quantified by various counting quantifiers. For $X \subseteq\{\exists \geq 1, \exists \geq 2, \ldots, \forall \geq 0, \forall \geq 1, \ldots\}, X \supseteq$ $\{\exists \geq 1, \forall \geq 0\}$, the $X-\operatorname{CSP}(\Gamma)$ takes as input a sentence given by a conjunction of atoms quantified by quantifiers appearing in $X$. It then asks whether this sentence is true on $\Gamma$. Equality languages admit quantifier elimination of $\forall$ and $\exists$, that is any relation first-order definable in equality is already quantifier-free definable, say as a CNF. An equality language $\Gamma$ is

- trivial if all its relations may be given as a conjunction of equalities,
- specially negative if the co-clone $\langle\Gamma\rangle_{\{\exists, \forall, \forall \geq 1\} \text {-pp }}$ does not contain the formula $x \neq y \vee y \neq z$,
- negative if all its relations may be given as a conjunction of equalities and disjunctions of disequalities, and
- positive if all its relations may be given as a conjunction of disjunctions of equalities.

Similarly, we might use these adjectives on the relations within the equality language. We observe the containments of trivial languages within specially negative languages within negative languages. Further, it is proved in [4] (Proposition 7.3) that the positive languages that are not trivial are precisely the positive languages that are not negative. Our main results are a complete panoply of classifications for $X \supseteq\{\exists \geq 1, \forall \geq 0\}$. It will be seen that the quantifiers $\exists \geq 2, \exists \geq 3, \ldots$ more or less behave as one another and similarly with $\forall^{\geq 2}, \forall \geq 3, \ldots$. However, $\forall \geq 1$ is special and thus our task of classifications for $X$ amounts to choosing subsets of $\{\exists \geq 2, \forall \geq 1, \forall \geq 2\}$ with which to augment $\{\exists \geq 1, \forall \geq 0\}$. A priori there are then eight possibilities, but twice we will see $\forall \geq 1$ being "subsumed" by $\forall^{\geq 2}$. Thus we will give six distinct classification theorems: three dichotomies and three trichotomies (one of which is that of [4]). In Fig. 1, these classification theorems are linked to their canonical subsets of $\left\{\exists \exists^{2}, \forall \geq 1, \forall \geq 2\right\}$.

Theorem 1 [4]. If $X=\left\{\exists \geq 1, \forall \geq^{0}\right\}$, then $X-C S P(\Gamma)$ displays trichotomy on the class of equality languages $\Gamma$ :

- if all relations of $\Gamma$ are negative then $X-\operatorname{CSP}(\Gamma)$ is in $L$.
- if all relations of $\Gamma$ are positive but some relation is not trivial, then $X$ $\operatorname{CSP}(\Gamma)$ is $N P$-complete.
- otherwise $X-C S P(\Gamma)$ is co-NP-hard.

Theorem 2. If $X \subseteq\left\{\exists \geq 1, \forall \geq 0, \forall \geq 1, \forall \geq^{2}, \ldots\right\}$ and contains some $\forall \geq j$ for $j \geq 2$, then $X-C S P(\Gamma)$ displays trichotomy on the class of equality languages $\Gamma$ :

- if all relations of $\Gamma$ are trivial, then $X-C S P(\Gamma)$ is in $L$.
- if all relations of $\Gamma$ are positive, but some relation is not trivial, then $X$ $\operatorname{CSP}(\Gamma)$ is NP-complete.
- otherwise $X-C S P(\Gamma)$ is Pspace-complete.

Theorem 3. If $X=\left\{\exists \geq^{1}, \forall^{\geq 0}, \forall \geq^{1}\right\}$, then $X-\operatorname{CSP}(\Gamma)$ displays trichotomy on the class of equality languages $\Gamma$ :

- if all relations of $\Gamma$ are specially negative, then $X-C S P(\Gamma)$ is in $P$.
- if all relations of $\Gamma$ are positive, but some relation is not trivial, then $X$ $\operatorname{CSP}(\Gamma)$ is $N P$-complete.
- otherwise $X-C S P(\Gamma)$ is Pspace-complete.

Theorem 4. If $X \subseteq\left\{\forall^{\geq 0}, \exists \geq 1, \exists \geq^{2}, \ldots\right\}$ and contains some $\exists^{\geq j}$ for $j \geq 2$ then $X-\operatorname{CSP}(\Gamma)$ displays dichotomy on the class of equality languages $\Gamma$ :

- if all relations of $\Gamma$ are negative, then $X-C S P(\Gamma)$ is in $L$.
- otherwise $X-C S P(\Gamma)$ is co-NP-hard.

Theorem 5. If $X \subseteq\{\forall \geq 0, \forall \geq 1, \forall \geq 2, \ldots, \exists \geq 1, \exists \geq 2, \ldots\}$ and contains some $\exists \geq i$ and $\forall \geq j$ for $i, j \geq 2$ then $X-C S P(\Gamma)$ displays dichotomy on the class of equality languages $\Gamma$ :

- if all relations of $\Gamma$ are trivial, then $X-C S P(\Gamma)$ is in $L$.
- otherwise $X-C S P(\Gamma)$ is Pspace-complete.

Theorem 6. If $X \subseteq\{\forall \geq 0, \forall \geq 1, \exists \geq 1, \exists \geq 2, \ldots\}$ and contains $\forall \geq 1$ and some $\exists \geq i$ for $i \geq 2$ then $X-\operatorname{CSP}(\Gamma)$ displays dichotomy on the class of equality languages $\Gamma$ :

- if all relations of $\Gamma$ are specially negative, then $X-\operatorname{CSP}(\Gamma)$ is in $P$.
- otherwise $X-C S P(\Gamma)$ is Pspace-complete.

| Theorem | Subsets |
| :---: | :---: |
| 1 | $\emptyset$ |
| 2 | $\left\{\forall^{\geq 2}\right\},\left\{\forall^{\geq 1}, \forall^{\geq^{2}}\right\}$ |
| 3 | $\left\{\forall^{\geq 1}\right\}$ |
| 4 | $\left\{\exists^{\geq^{2}}\right\}$ |
| 5 | $\left\{\forall^{\geq^{2}}, \exists^{\geq 2}\right\},\left\{\forall^{\geq^{1}}, \forall^{\geq^{2}}, \exists^{\geq^{2}}\right\}$ |
| 6 | $\left\{\forall^{\geq^{1}}, \exists^{\geq^{2}}\right\}$ |

Fig. 1. Classification theorems linked to canonical subsets of $\left\{\exists^{\geq 2}, \forall^{\geq 1}, \forall \geq^{2}\right\}$

Four of our five new worlds are somewhat more conducive to analysis than that of [4], in that in them we have no gap across co-NP-hardness and Pspacecompleteness. For the remaining world of Theorem 4, we are able to demonstrate that improving co-NP- to Pspace-hardness is likely to be as difficult as in Theorem 1. Indeed, from this it follows that our working hypothesis promotes co-NP-hardness to Pspace-hardness for Theorem 4 as well as Theorem 1.

Some of our results are not especially complicated and stem from simple manipulations rather than deep technical nous. Against this we set the nice aesthetic of our results and the way in which they complement [4,6]. For example, the specially negative languages, play an important role in our classifications, but where do they sit in the context of [6]. Do they even form a co-clone? We note that equality languages have been also studied for the abduction problem [18], in the context of which we find a trichotomy among $\Sigma_{2}^{\mathrm{P}}$-complete, NP-complete and problems decidable in P.

The paper is organised as follows. After the preliminaries, we address in Sect. 3 basic upper and lower bounds that play a role in our classification. In Sect. 4 we describe the crucial specially negative languages. Finally, we ponder the difficulty of improving, in Theorem 4, co-NP- to Pspace-hardness. We do the latter by showing that $\operatorname{QCSP}(x=y \vee u=v, \neq)$ becomes no more complex when one augments with $\exists \geq k$.

## 2 Preliminaries

Let $[j]:=\{1, \ldots, j\}$. For $i \in \mathbb{N}, i \geq 1$, let $\exists{ }^{\geq} x$ quantify that there exist at least $i$ elements satisfying some property. For $i \in \mathbb{N}, i \geq 0$, let $\forall^{\geq} x$ quantify that for all but at most $i$ elements does some property hold. Thus $\exists$ is $\exists \geq 1$ and $\forall$ is $\forall \geq 0$ - the base cases being different for these quantifiers. In this paper we consider languages with first-order (fo) definitions in equality, that is on structures which admit all permutations as automorphisms. Such an equality language may be considered a structure of the form ( $\mathbb{N} ; R_{1}, \ldots$ ) where each $R_{i}$ is a CNF formula whose atoms are equalities or disequalities (owing to quantifier elimination of $\forall$ and $\exists$ this is equivalent to our saying fo-definable in equality). We typically drop the " $\mathbb{N}$ " in referring to an equality language.

Primitive positive ( pp ) logic is the restriction of fo-logic to the symbols $\{\exists, \wedge,=\}$ and positive Horn ( pH ) likewise to the symbols $\{\forall, \exists, \wedge,=\}$. For $X \subseteq\left\{\forall \geq 0, \forall \geq 1, \ldots, \exists \geq 1, \exists \geq^{2}, \ldots\right\}$, let $X$-pp denote the logic of prenex sentences whose symbols are among $X \cup\{\wedge,=\}$. We will use small letters such as $\phi$ to refer to sentences whose quantifier-free part will be denoted $\Phi$. Let $\Gamma$ be a set of relations on a domain, i.e. a structure, and let $\mathscr{L}$ be a logic. Then the evaluation problem for $\mathscr{L}$ on $\Gamma$ has as input a sentence $\phi \in \mathscr{L}$ and asks whether $\Gamma \models \phi$ ? The evaluation problem for primitive positive (resp., positive Horn) logic on $\Gamma$ is better known as $\operatorname{CSP}(\Gamma)$ (resp., $\operatorname{QCSP}(\Gamma)$ ). The evaluation problem for $X$-pp on $\Gamma$ will henceforth be known as $X-\operatorname{CSP}(\Gamma)$. In this paper we consider only $X \supseteq\left\{\exists \geq 1, \forall^{\geq 0}\right\}$, i.e. extensions of the QCSP.

Let $\Gamma$ be an equality language, and let $\langle\Gamma\rangle_{\mathrm{pp}}$ be the set of relations ppdefinable over $\Gamma$. Such a set is termed a co-clone. We may abuse notation and write $\langle R\rangle_{\mathrm{pp}}$ when properly we mean $\langle\{R\}\rangle_{\mathrm{pp}}$. A great project was launched in [6] to identify sets of the form $\langle\Gamma\rangle_{\mathrm{pp}}$, charting their inclusion relations in a lattice. The lattice is mostly identified through a dually-isomorphic algebraic lattice of local clones, but here we will only be interested in the co-clones. In line with $[4,6]$, we will consider the co-clone $\langle\emptyset\rangle_{\mathrm{pp}}=\langle=\rangle_{\mathrm{pp}}$ to be at the top of the lattice, with the co-clone of all equality-definable relations at the bottom.

In a problem $X-\operatorname{CSP}(\Gamma)$ it is desirable that $\Gamma$ involve only a finite number of relations lest there arise the question as to how they are encoded. Yet we will enjoy referring to co-clones that contain an infinite number of relations. We resolve this by considering all references to $\Gamma$ as a language inside $X-\operatorname{CSP}(\Gamma)$ to be restricting of $\Gamma$ to the relationally finite. ${ }^{1}$

We now recall some basic results from $[4,6]$. We typically write $\neq$ to indicate the binary relation of disequality (in terms of our canonical notation this abbreviates $x \neq y)$. Let $\langle\Gamma\rangle_{\mathrm{pH}}$ be the set of relations pH -definable over $\Gamma$.

[^52]
## Lemma 1.

(1) $x=y \vee u=v$ is in both $\langle x=y \vee y=z\rangle_{\mathrm{pp}}$ and $\langle x=y \vee y=z \vee x=z\rangle_{\mathrm{pp}}$.
(2) $x \neq y \vee u \neq v \in\langle x=y \rightarrow y=z, \neq\rangle_{\mathrm{pp}}$.
(3) $\left(x_{1}=y \wedge \ldots \wedge x_{m}=y\right) \rightarrow y=z \in\langle x=y \rightarrow y=z\rangle_{\mathrm{pp}}$.
(4) $\langle\{x=y \vee u=v, \neq\}\rangle_{\mathrm{pp}}$ contains all equality-definable relations.
(5) $\langle x=y \vee u=v\rangle_{\mathrm{pp}}$ contains all positive relations.
(6) If $\Gamma$ is not positive then $\neq \in\langle\Gamma\rangle_{\mathrm{pp}}$.
(7) If $\Gamma$ is positive but not negative then $x=y \vee u=v \in\langle\Gamma\rangle_{\mathrm{pH}}$.

Note that Part 7 does not hold for pp-definability. While $x=y \vee y=z$ is a ppbasis for the positive languages, there is an infinite chain of positive languages up to pp-closure [6]. $\langle x=y \vee y=z\rangle_{\text {pp }}$ is at the bottom, being the most expressive, and the trivial $\langle=\rangle_{\mathrm{pp}}$ is at the top. Inbetween, is an infinite chain without top element. However, this chain collapses for pH -definability, leaving only the two co-clones up to pH -closure ( $\langle=\rangle_{\mathrm{pp}}$ and $\langle x=y \vee y=z\rangle_{\mathrm{pp}}$ ).

We contrast Part 1 of Lemma 1 with the knowledge that $x=y \rightarrow u=v$ is in neither $\langle x=y \rightarrow y=z\rangle_{\mathrm{pp}}[6]$ nor $\langle x=y \rightarrow y=z\rangle_{\mathrm{pH}}$.
Owing to the disparity between the conference and journal versions of [4], we give the following as a specific proposition. Its proof can be derived from [4] by the assiduous reader.

Proposition 1. Both $\operatorname{QCSP}(x=y \vee u=v, \neq)$ and $\operatorname{QCSP}\left(w=z_{1} \vee w=z_{2} \vee w=\right.$ $\left.z_{3}, \neq\right)$ are Pspace-complete.

It seems that $\operatorname{QCSP}(x=y \rightarrow u=v)$ is also Pspace-complete [5], but this is harder work. The backbone of a proof appeared in an early unpublished version of [4]. The author has verified this proof but reproducing it here is beyond our scope.

## 3 Upper and Lower Bounds

Upper bounds. The following lemma is trivial but we state it because we will wish to appeal to it in the future.

Lemma 2 (Substitution of equalities). Let $X:=\{\forall \geq 0, \forall \geq 1, \exists \geq 1, \exists \geq 2, \ldots\}$ and let $\Phi$ be an instance of some $X$-CSP containing an equality $x=y$ in which $y$ appears later in the quantifier order of $\Phi$ than $x$. Then $\Phi$ is false if $y$ is quantified by anything other than $\exists \geq 1$. Otherwise, $\Phi$ is equivalent to $\Phi^{\prime}$ obtained by substituting all instances of $y$ by $x$ and removing the quantifier $\exists^{\geq 1} y$.

Lemma 3. For any $X \subseteq\left\{\forall^{\geq 0}, \forall \geq 1, \ldots, \exists \geq 1, \exists \geq 2, \ldots\right\}, X-C S P(\Gamma)$ is in Pspace.
The following lemma relates to $\{\exists \geq 1, \forall \geq i: i \geq 0\}$ on positive languages.
Lemma 4. Let $X:=\{\exists \geq 1, \forall \geq i, i \geq 0\}$ and $\Gamma$ be a positive equality language. Then $X-C S P(\Gamma)$ is in $N P$.

Lower bounds. The following proposition relates to $\{\forall \geq i, \forall \geq 0, \exists \geq 1\}$, for $i \geq 2$, on non-positive languages.

Proposition 2. Let $X:=\left\{\forall^{\geq i}, \forall \geq^{0}, \exists \geq^{\geq 1}\right\}$, for any $i \geq 2$ and $\Gamma$ a non-positive equality language. Then $X-C S P(\Gamma)$ is Pspace-complete.

The following propositions relate to $\{\forall \geq 1, \forall \geq 0, \exists \geq 1\}$ on non-positive languages.

Proposition 3. $\left\{\forall^{\geq 1}, \forall \geq^{0}, \exists \geq 1\right\}-C S P(x \neq y \vee u \neq v)$ is Pspace-complete.
Proposition 4. $(x \neq y \vee u \neq v)$ is definable from $(x \neq y \vee y \neq z)$ using $\forall \geq 1$.
Proposition 5. $\{\forall \geq 1, \forall \geq 0, \exists \geq 1\}-C S P(x \neq y \vee y \neq z)$ is Pspace-complete.
Proposition 6. Let $i \geq 2$ and $\Gamma$ be a positive non-trivial equality language. Then $\neq$ is definable from $\Gamma$ using $\exists \geq i$.

## 4 Isolating the Specially Negative Languages

We call a CNF $\Phi$ reduced if it is not logically equivalent to itself with either a clause or a literal in a clause removed. A CNF depends on one of its variables $v$ if its truth value can not be given as a propositional function from (the equality type of) only its other variables.

Suppose $R$ is a negative relation which might be given by various reduced CNFs, at least one of which, $\Phi$, is negative. Then $\Phi$ may enforce some equalities on its variables, which it plainly does syntactically. Sometimes in this section we will wish to assume that $\Phi$ has these equalities factored out by substitution, thus we could assume negative CNFs do not have any positive clauses. The point is that this is an innocuous assumption. Having said that, the process does remove variables and so can't be used strictly within a quantifier elimination procedure. Now, suppose $R$ is also given by another reduced CNF $\Phi^{\prime}$ which is not negative (we will later prove this is not possible). Then we could similarly factor out the implied equalities of $\Phi^{\prime}$, but these might not be obvious unless we have negative $\Phi$ where the equalities are syntactically explicit and not semantically implied.

### 4.1 Non-positive Cases Involving $\forall \geq 1$

Definition 1. A negative, reduced CNF $\Phi$ without equalities is flat if it consists of clauses with no free variable occurring twice in them. $\Phi$ is rich if every free variable of $\Phi$ occurs in a singleton clause of $\Phi$.

We will assume that CNFs do not possess dummy variables that do not appear explicitly. Thus, we may consider a CNF to be rich if it is rich once we have discounted such dummy variables.

Theorem 7. We have the following dichotomy for non-positive equality languages $\Gamma$.

1. Either $\langle\Gamma\rangle_{\{\exists, \forall, \forall \geq 1\}-\mathrm{pp}}$ contains the relation $x \neq y \vee y \neq z$, and then we find $\left\{\exists, \forall, \forall^{\geq 1}\right\}-\operatorname{CSP}(\Gamma)$ is Pspace-hard,
2. or $\left.\langle\Gamma\rangle_{\{\exists, \exists \geq 2}, \ldots, \forall, \forall \geq 1\right\}-\mathrm{pp}$ contains only relations whose reduced CNFs are negative, flat and rich; and $\{\exists, \exists \geq 2, \ldots, \forall, \forall \geq 1\}-\operatorname{CSP}(\Gamma)$ is in $P$.

Definition 2. Let $\Phi$ be a $C N F$ with variables in $V$. Let $P_{1} \cup \cdots \cup P_{k}$ be a partition of $V$. We say that "we weaken $\Phi$ around the given partition by keeping $P_{1}, \ldots, P_{j}$ " if we produce a formula from $\Phi$ by the following definition. First we take the conjunction $\Psi$ of all disequalities that are transversal to the partition, i.e., $x \neq y$ with $x$ and $y$ not in the same set of the partition, and produce $\Phi \wedge \Psi$. Then for all $i>j$ we identify the variables in $P_{i}$ by a new variable $w_{i}$. Then we existentially quantify over these new variables.

Note that if $\Phi$ pp-defines $\neq$, then this procedure is also a pp-definition from $\Phi$.
Lemma 5. Let $\Phi$ be a reduced negative CNF that has a non-flat clause. Then $\Phi$ pp-defines $x \neq y \vee y \neq z$.

Lemma 6. From $S(x, y, u, v):=(x \neq y \vee u \neq v) \wedge y \neq u \wedge x \neq u$ we may define with $\forall \geq 1$ and $\exists$ the relation $p \neq q \vee q \neq r$.

Lemma 7. Let $\Phi$ be a flat negative CNF with exactly four variables. Assume that $\Phi$ depends on each of its variables, and that $\Phi$ is not rich. Then $x \neq y \vee y \neq z$ has an $\{\exists, \forall, \forall \geq 1\}-p p$ definition in $\Phi$.

Lemma 8. Assume that $\langle\Gamma\rangle_{\{\exists, \forall, \forall \geq 1\}-\mathrm{pp}}$ contains a negative $C N F \Phi$ that is not rich. Then $\langle\Gamma\rangle_{\{\exists, \forall, \forall \geq 1\}-\mathrm{pp}}$ contains the formula $x \neq y \vee y \neq z$.

### 4.2 Quantifier Elimination and Reduction for Negative, Flat and Rich CNF Formulas

We wish to argue that in a certain case we can effect quantifier elimination for negative, flat and rich CNF formulas in polynomial time. Eliminating a $\forall \geq 1$ quantifier over negative, flat and rich formulas can throw up Horn CNFs, so our first task will be to argue that we can compute a reduced form from a Horn CNF $\Phi$ in polynomial time. Our task is to determine, whether there are any redundant clauses (and if so remove them) and then whether there are any redundant literals in the remaining clauses (which also must be removed). To test for redundant clauses $C$ it is sufficient to test whether $(\Phi \backslash\{C\})$ implies $C$, i.e. whether $(\Phi \backslash\{C\}) \wedge \neg C$ is a contradiction. But this is itself the complement of a Horn CSP, itself uniformly tractable by unit propagation (see [7]). Similarly, to determine if a literal $\ell$ is required in a clause $C$, we may consider whether $(\Phi \backslash\{C\}) \wedge \ell \wedge \neg(C \backslash\{\ell\})$ is a contradiction. Thus, given a Horn CNF $\Phi$ we can compute in polynomial time a reduced CNF (itself Horn) $\Phi^{\prime}$ that is equivalent to $\Phi$.

Consider each $\forall \geq 1 x_{i} \Phi\left(x_{1}, \ldots, x_{k}\right)$, where $\Phi$ is negative, flat and rich, without any equalities, where the singletons involving $x_{i}$ are precisely $x_{i} \neq x_{\lambda_{1}}^{\prime}, \ldots, x_{i} \neq$
$x_{\lambda_{t}}^{\prime}$. We can effect quantifier elimination in the following fashion. Substitute each non-singleton clause involving $x_{i}$, itself of the form, $\left(x_{i} \neq x_{\mu_{1}}^{\prime} \vee x_{\mu_{2}} \neq x_{\mu_{2}}^{\prime} \vee \ldots \vee\right.$ $x_{\mu_{s}} \neq x_{\mu_{s}}^{\prime}$ ), where no variable is repeated, by $\left(x_{\mu_{1}}^{\prime}=x_{\lambda_{1}}^{\prime}=\cdots=x_{\lambda_{t}}^{\prime} \vee x_{\mu_{2}} \neq\right.$ $\left.x_{\mu_{2}}^{\prime} \vee \ldots \vee x_{\mu_{s}} \neq x_{\mu_{s}}^{\prime}\right)$. Note that we will allow conjuncts of equalities as single relations in our clauses, to avoid having to break the clauses up. If we view this as a notational shorthand then we do not break the condition of Hornness. If $x_{i}$ appeared in multiple singleton clauses (i.e. $t>1$ ), then we need to add the equality $x_{\lambda_{1}}=\cdots=x_{\lambda_{t}}$ to the system. If $x_{i}$ appeared in a single singleton clause then we can now simply remove it. Call the new CNF finally obtained $\tilde{\Phi}$. Plainly, $\tilde{\Phi}$ is logically equivalent to $\forall \geq 1 x_{i} \Phi$ and $\tilde{\Phi}$ is Horn. We now apply our effective procedure to establish whether $\tilde{\Phi}$ is reduced and if not reduce it.

The question now naturally arises as to whether $\tilde{\Phi}$ is negative. We argue by the following lemma that it is enough to see whether our reduced CNF form has a non-singleton clause with an equality in it.

Lemma 9. Let $\Phi$ be a reduced CNF representing a relation $R$ in which there is a non-singleton clause that contains an equality. It is not possible that $R$ is negative (i.e. has another reduced CNF that is negative).

Note that reduced negative CNFs of a relation $R$ are not in general unique, and the method of transitive closure from the first paragraph of the above proof hints at an example: $(x \neq y \vee y \neq z)$ is equivalent to $(x \neq y \vee x \neq z)$.

We will now consider the quantifier elimination of $\exists, \exists \geq 2, \ldots$ etc. (note that quantification by $\forall$ on a negative, flat and rich formula will always leave it false). On negative formulas, $\exists \geq^{2}$, ..etc. have the same power as $\exists$ and allow us to remove any clause in which the corresponding variable appears. This leaves the formula negative and we again have an effective method for reduction.

We now continue in this vein eliminating quantifiers and reducing CNFs. If we at any point produce a CNF whose reduced form is not negative, flat and rich then we know that $\Gamma\left\{\exists, \forall, \forall^{\geq 1}\right\}$-pp defines $x \neq y \vee y \neq z$. We are now in a position to address Theorem 7 .

By way of example for Theorem 7, we note that $\langle\{(u \neq v \vee x \neq y) \wedge u \neq$ $y \wedge v \neq x\}\rangle_{\{\exists, \forall, \forall \geq 1\}-p p}$ contains only negative, flat and rich formulas.

## 5 Ennui of co-NP- to Pspace-Completeness

Proposition 7. For all $1 \leq k \in \mathbb{N}$, $\left\{\exists^{\geq k}, \forall^{\geq 0}, \exists^{\geq 1}\right\}-\operatorname{CSP}(x=y \rightarrow y=z)$ and $\left\{\forall \geq^{0}, \exists{ }^{1}\right\}-\operatorname{CSP}(x=y \rightarrow y=z)$ (i.e. $\operatorname{QCSP}(x=y \rightarrow y=z)$ ) are logspace equivalent.

It follows from Proposition 7 and [4] that our working conjecture that QCSP $(x=y \rightarrow y=z)$ is Pspace-complete would elevate the co-NP-hardness cases of Theorems 1 and 4 to Pspace-hardness.

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# Baire Category Theory and Hilbert's Tenth Problem Inside $\mathbb{Q}$ 

Russell Miller ${ }^{1,2(\boxtimes)}$<br>${ }^{1}$ Queens College - C.U.N.Y., $65-30$ Kissena Blvd., Queens, NY 11367, USA<br>Russell.Miller@qc.cuny.edu<br>${ }^{2}$ Graduate Center of C.U.N.Y., 365 Fifth Avenue, New York, NY 10016, USA<br>http://qcpages.qc.cuny.edu/~rmiller


#### Abstract

For a ring $R$, Hilbert's Tenth Problem $\operatorname{HTP}(R)$ is the set of polynomial equations over $R$, in several variables, with solutions in $R$. We consider computability of this set for subrings $R$ of the rationals. Applying Baire category theory to these subrings, which naturally form a topological space, relates their sets $\operatorname{HTP}(R)$ to the set $\operatorname{HTP}(\mathbb{Q})$, whose decidability remains an open question. The main result is that, for an arbitrary set $C, \operatorname{HTP}(\mathbb{Q})$ computes $C$ if and only if the subrings $R$ for which $\operatorname{HTP}(R)$ computes $C$ form a nonmeager class. Similar results hold for 1-reducibility, for admitting a Diophantine model of $\mathbb{Z}$, and for existential definability of $\mathbb{Z}$.


## 1 Introduction

The original version of Hilbert's Tenth Problem demanded an algorithm deciding which polynomial equations from $\mathbb{Z}\left[X_{1}, X_{2}, \ldots\right]$ have solutions in the integers. In 1970, Matiyasevic [4] completed work by Davis, Putnam and Robinson [1], showing that no such algorithm exists. In particular, these authors showed that there exists a 1-reduction from the Halting Problem $\emptyset^{\prime}$ to the set of such equations with solutions, by proving the existence of a single polynomial $h \in \mathbb{Z}[Y, \boldsymbol{X}]$ such that, for each $n$ from the set $\omega$ of nonnegative integers, the polynomial $h(n, \boldsymbol{X})=0$ has a solution in $\mathbb{Z}$ if and only if $n$ lies in $\emptyset^{\prime}$. Since the membership in the Halting Problem was known to be undecidable, it followed that Hilbert's Tenth Problem was also undecidable.

One naturally generalizes this problem to all rings $R$, defining Hilbert's Tenth Problem for $R$ to be the set

$$
\operatorname{HTP}(R)=\left\{f \in R[\boldsymbol{X}]:\left(\exists r_{1}, \ldots, r_{n} \in R^{<\omega}\right) f\left(r_{1}, \ldots, r_{n}\right)=0\right\}
$$

[^53]Here we will examine this problem for one particular class: the subrings $R$ of the field $\mathbb{Q}$ of rational numbers. Notice that in this situation, deciding membership in $\operatorname{HTP}(R)$ reduces to the question of deciding this membership just for polynomials from $\mathbb{Z}[\boldsymbol{X}]$, since one readily eliminates denominators from the coefficients of a polynomial. So, for us, $\operatorname{HTP}(R)$ will always be a subset of $\mathbb{Z}\left[X_{1}, X_{2}, \ldots\right]$.

Subrings $R$ of $\mathbb{Q}$ correspond bijectively to subsets $W$ of the set $\mathbb{P}$ of all primes, via the map $W \mapsto \mathbb{Z}\left[\frac{1}{p}: p \in W\right]$. We write $R_{W}$ for the subring $\mathbb{Z}\left[\frac{1}{p}: p \in W\right]$. In this article, we will move interchangeably between subsets of $\omega$ and subsets of $\mathbb{P}$, using the bijection mapping $n \in \omega$ to the $n$-th prime $p_{n}$, starting with $p_{0}=2$. For the most part, our sets will be subsets of $\mathbb{P}$, but Turing reductions and jump operators and the like will all be applied to them in the standard way. Likewise, sets of polynomials, such as $\operatorname{HTP}(R)$, will be viewed as subsets of $\omega$, using a fixed computable bijection from $\omega$ onto $\mathbb{Z}[\boldsymbol{X}]=\mathbb{Z}\left[X_{0}, X_{1}, \ldots\right]$.

We usually view subsets of $\mathbb{P}$ as paths through the tree $2^{<\mathbb{P}}$, a complete binary tree whose nodes are the functions from initial segments of the set $\mathbb{P}$ into the set $\{0,1\}$. This allows us to introduce a topology on the space $2^{\mathbb{P}}$ of paths through $2^{<\mathbb{P}}$, and thus on the space of all subrings of $\mathbb{Q}$. Each basic open set $\mathcal{U}_{\sigma}$ in this topology is given by a node $\sigma$ on the tree: $\mathcal{U}_{\sigma}=\{W \subseteq \mathbb{P}: \sigma \subset W\}$, where $\sigma \subset W$ denotes that when $W$ is viewed as a function from $\mathbb{P}$ into the set $2=\{0,1\}$ (i.e., as an infinite binary sequence), $\sigma$ is an initial segment of that sequence. Also, we put a natural measure $\mu$ on the class $\operatorname{Sub}(\mathbb{Q})$ of all subrings of $\mathbb{Q}$ : just transfer to $\operatorname{Sub}(\mathbb{Q})$ the obvious Lebesgue measure on the power set $2^{\mathbb{P}}$ of $\mathbb{P}$. Thus, if we imagine choosing a subring $R$ by flipping a fair coin (independently for each prime $p$ ) to decide whether $\frac{1}{p} \in R$, the measure of a subclass $\mathcal{S}$ of $\operatorname{Sub}(\mathbb{Q})$ is the probability that the resulting subring will lie in $\mathcal{S}$. Here we will focus on Baire category theory rather than on measure theory, however, as the former yields more useful results. For questions and results regarding measure theory, we refer the reader to Sect. 3 and to the forthcoming [5].

For all $W \subseteq \mathbb{P}$, we have Turing reductions, which in fact are 1-reductions:

$$
W \oplus \operatorname{HTP}(\mathbb{Q}) \leq_{1} \operatorname{HTP}\left(R_{W}\right) \leq_{1} W^{\prime}
$$

For instance, the Turing reduction from $\operatorname{HTP}\left(R_{W}\right)$ to $W^{\prime}$ can be described by a computable injection which maps each $f \in \mathbb{Z}[\boldsymbol{X}]$ to the code number $h(f)$ of an oracle Turing program which, on every input, searches for a solution $\boldsymbol{x}$ to $f=0$ in $\mathbb{Q}$ for which the primes dividing the denominators of the coordinates in $\boldsymbol{x}$ all lie in the oracle set $W$. The reduction from $\operatorname{HTP}(\mathbb{Q})$ to $\operatorname{HTP}\left(R_{W}\right)$ uses the fact that every element of $\mathbb{Q}$ is a quotient of elements of $R_{W}$, so that $f(\boldsymbol{X})$ has a solution in $\mathbb{Q}$ if and only if $Y^{d} \cdot f\left(\frac{X_{1}}{Y}, \ldots, \frac{X_{n}}{Y}\right)$ has a solution in $R_{W}$ with $Y>0$. The condition $Y>0$ is readily expressed using the Four Squares Theorem. Finally, $W \leq_{1} \operatorname{HTP}\left(R_{W}\right)$ by mapping $p$ to $(p X-1)$.

The topological space $2^{\mathbb{P}}$ of all paths through $2^{<\mathbb{P}}$, which we treat as the space of all subrings of $\mathbb{Q}$, is obviously homeomorphic to Cantor space, the space $2^{\omega}$ of all paths through the complete binary tree $2^{<\omega}$. Hence this space satisfies the property of Baire, that no nonempty open set is meager. We recall the relevant definitions. Here as before, $\overline{\mathcal{A}}$ represents the complement of a subset $\mathcal{A} \subseteq 2^{\mathbb{P}}$, and we will write $\operatorname{cl}(\mathcal{A})$ for the topological closure of $\mathcal{A}$ and $\operatorname{Int}(\mathcal{A})$ for its interior.

Definition 1. A subset $\mathcal{B} \subseteq 2^{\mathbb{P}}$ is said to be nowhere dense if its closure $\operatorname{cl}(\mathcal{B})$ contains no nonempty open subset of $2^{\mathbb{P}}$. In particular, every set $\mathcal{U}_{\sigma}$ with $\sigma \in 2^{<\mathbb{P}}$ must intersect $\operatorname{Int}(\overline{\mathcal{B}})$, the interior of the complement of $\mathcal{B}$.

The union of countably many nowhere dense subsets of $2^{\omega}$ is called a meager set, or a set of first category. Its complement is said to be comeager.

All sets $W \subseteq \omega$ satisfy $W \oplus \emptyset^{\prime} \leq_{T} W^{\prime}$, and for certain $W$, Turing-equivalence holds here. Indeed, it is known that the class

$$
\mathbf{G} \mathbf{L}_{1}=\left\{W \in 2^{\omega}: W^{\prime} \equiv_{T} W \oplus \emptyset^{\prime}\right\}
$$

is comeager, although its complement is nonempty. In computability theory, elements of $\mathbf{G L}_{1}$ are called generalized-low l $_{1}$ sets. The low sets - i.e., those $W$ with $W^{\prime} \leq_{T} \emptyset^{\prime}$ - clearly lie in $\mathbf{G L}_{1}$.

Lemma 1 (Folklore). There exists a Turing functional $\Psi$ such that $\{W \subseteq \omega$ : $\left.\Psi^{W \oplus \emptyset^{\prime}}=\chi_{W^{\prime}}\right\}$ is comeager. It follows that $\boldsymbol{G} \boldsymbol{L}_{1}$ is comeager.

Proof. Consider the following oracle program $\Psi$ for computing $W^{\prime}$ from $W \oplus \emptyset^{\prime}$. With this oracle, on input $e$, the program searches for a string $\sigma \subseteq W$ such that either $(1)(\exists s) \Phi_{e, s}^{\sigma}(e) \downarrow$, or $(2)(\forall \tau \supseteq \sigma)(\forall s) \Phi_{e, s}^{\tau}(e) \uparrow$. The program uses its $\emptyset^{\prime}$ oracle to check the truth of these two statements for each $\sigma \subseteq W$. If it ever finds that (1) holds, it concludes that $e \in W^{\prime}$; while if it ever finds that (2) holds, it concludes that $e \notin W^{\prime}$. Thus, $\Psi^{W \oplus \emptyset^{\prime}}$ can only fail to compute $W^{\prime}$ if there exists some $e \notin W^{\prime}$ such that, for every $n$, some $\tau \supset W \upharpoonright n$ has $\Phi_{e}^{\tau}(e) \downarrow$. This can happen, but for each single $e$, the set of those $W$ for which this happens constitutes the boundary of the open set $\left\{W: e \in W^{\prime}\right\}$. This boundary is nowhere dense (cf. Lemma 3 below), so the union of these sets (over all $e$ ) is meager, and $\Psi^{W \oplus \emptyset^{\prime}}=\chi_{W^{\prime}}$ for every $W$ outside this meager set.
$\mathbf{G L}_{1}$ also has measure 1, but no single Turing functional computes $W^{\prime}$ from $W \oplus \emptyset^{\prime}$ uniformly on a set of measure 1 .

Lemma 2 (Folklore). If $A \not \unlhd_{T} B$, then $\mathcal{C}=\left\{W: A \oplus W \geq_{T} B\right\}$ is meager.
Proof. To show that $\mathcal{C}$ is meager, define $\mathcal{C}_{e}=\left\{W \subseteq \mathbb{P}: \Phi_{e}^{A \oplus W}=\chi_{B}\right\}$, so $\mathcal{C}=\cup_{e} \mathcal{C}_{e}$. We claim that, if $\sigma \in 2^{\mathbb{P}}$ and $\mathcal{U}_{\sigma} \subseteq \operatorname{cl}\left(\mathcal{C}_{e}\right)$, the following hold.

1. $\forall x \forall \tau \supseteq \sigma\left[\Phi_{e}^{A \oplus \tau}(x) \uparrow\right.$ or $\left.\Phi_{e}^{A \oplus \tau}(x) \downarrow=\chi_{B}(x)\right]$.
2. $\forall x \exists \tau \supseteq \sigma\left[\Phi_{e}^{A \oplus \tau}(x) \downarrow\right]$.

To see that (1) holds, suppose $\Phi_{e}^{A \oplus \tau}(x) \downarrow$. With $\mathcal{U}_{\tau} \subseteq \mathcal{U}_{\sigma} \subseteq \operatorname{cl}\left(\mathcal{C}_{e}\right)$, some $W \in \mathcal{C}_{e}$ must have $\tau \subseteq W$. But then $\chi_{B}(x)=\Phi_{e}^{A \oplus W}(x) \downarrow=\Phi_{e}^{A \oplus \tau}(x)$.

To see (2), fix any $W \in \mathcal{C}_{e}$ with $\sigma \subseteq W$ : such a $W$ must exist, since $\mathcal{U}_{\sigma} \subseteq$ $\operatorname{cl}\left(\mathcal{C}_{e}\right)$. Then we can take $\tau$ to be the restriction of this $W$ to the use of the computation $\Phi_{e}^{A \oplus W}(x)$ (or $\tau=\sigma$ if the use is $<|\sigma|$ ).

But now every $\mathcal{C}_{e}$ must be nowhere dense, since any $\sigma$ satisfying (1) and (2) would let us compute $B$ from $A$ : given $x$, just search for some $\tau \supseteq \sigma$ and some $s$ for which $\Phi_{e, s}^{A \oplus \tau}(x) \downarrow$. By (2), our search would discover such a $\tau$ eventually, and by (1) we would know $\chi_{B}(x)=\Phi_{e, s}^{A \oplus \tau}(x)$. Since $A \not{ }_{T} B$, this is impossible.

Finally, on a separate topic, it will be important for us to know that whenever $R$ is a semilocal subring of $\mathbb{Q}$, we have $\operatorname{HTP}(R) \leq_{1} \operatorname{HTP}(\mathbb{Q})$. Indeed, both the Turing reduction and the 1-reduction are uniform in the complement. (The result essentially follows from work of Julia Robinson in [8]. For a proof by Eisenträger, Park, Shlapentokh, and the author, see [2].) Recall that the semilocal subrings of $\mathbb{Q}$ are precisely those of the form $R_{W}$ where the set $W$ is cofinite in $\mathbb{P}$, containing all but finitely many primes.

Proposition 1 (see Proposition 5.4 in [2]). There exists a computable function $G$ such that for every $n$, every finite set $A_{0}=\left\{p_{1}, \ldots, p_{n}\right\} \subset \mathbb{P}$ and every $f \in \mathbb{Z}[\boldsymbol{X}]$,

$$
f \in H T P\left(R_{\mathbb{P}-A_{0}}\right) \Longleftrightarrow G\left(f,\left\langle p_{1}, \ldots, p_{n}\right\rangle\right) \in H T P(\mathbb{Q}) .
$$

That is, $\operatorname{HTP}\left(R_{\mathbb{P}-A_{0}}\right)$ is 1-reducible to $\operatorname{HTP}(\mathbb{Q})$ for all semilocal $R_{\mathbb{P}-A_{0}}$, uniformly in $A_{0}$.

The proof in [2], using work from [3], actually shows how to compute, for every prime $p$, a polynomial $f_{p}\left(Z, X_{1}, X_{2}, X_{3}\right)$ such that for all rationals $q$, we have

$$
q \in R_{\mathbb{P}-\{p\}} \Longleftrightarrow f_{p}(q, \boldsymbol{X}) \in \operatorname{HTP}(\mathbb{Q})
$$

## 2 Baire Category and $\operatorname{HTP}(\mathbb{Q})$

For a polynomial $f \in \mathbb{Z}[\boldsymbol{X}]$ and a subring $R_{W} \subseteq \mathbb{Q}$, there are three possibilities. First, $f$ may lie in $\operatorname{HTP}\left(R_{W}\right)$. If this holds for $R_{W}$, the reason is finitary: $W$ contains a certain finite (possibly empty) subset of primes generating the denominators of a solution. Second, there may be a finitary reason why $f \notin \operatorname{HTP}\left(R_{W}\right)$ : there may exist a finite subset $A_{0}$ of the complement $\bar{W}$ such that $f$ has no solution in $R_{\mathbb{P}-A_{0}}$. For each finite $A_{0} \subset \mathbb{P}$, the set $\operatorname{HTP}\left(R_{\mathbb{P}-A_{0}}\right)$ is 1-reducible to $\operatorname{HTP}(\mathbb{Q})$, by Proposition 1 ; indeed the two sets are computably isomorphic, with a computable permutation of $\mathbb{Z}[\boldsymbol{X}]$ mapping one onto the other. Therefore, the existence of such a set $A_{0}$ (still for one fixed $f$ ) is a $\Sigma_{1}^{\mathrm{HTP}(\mathbb{Q})}$ problem.

The third possibility is that neither of the first two holds. An example is given in [5], where it is shown that a particular polynomial $f$ fails to lie in $\operatorname{HTP}\left(R_{W_{3}}\right)$, where $W_{3}$ is the set of all primes congruent to 3 modulo 4 , yet that, for every finite set $V_{0}$ of primes, there exists some $W$ disjoint from $V_{0}$ with $f \in \operatorname{HTP}\left(R_{W}\right)$. We consider sets such as this $W_{3}$ to be on the boundary of $f$, in consideration of the topology of the situation. The set $\mathcal{A}(f)=\left\{W: f \in \operatorname{HTP}\left(R_{W}\right)\right\}$ is open in the usual topology on $2^{\mathbb{P}}$, since, for any solution of $f$ in $R_{W}$ and any $\sigma \subseteq W$ long enough to include all primes dividing the denominators in that solution, every other $V \supseteq \sigma$ will also contain that solution. Moreover, one can computably enumerate the collection of those $\sigma$ such that the basic open set $\mathcal{U}_{\sigma}=\{W: \sigma \subseteq$ $W\}$ is contained within $\mathcal{A}(f)$. The set $\operatorname{Int}(\overline{\mathcal{A}(f)})$ is similarly a union of basic open sets, and these can be enumerated by an $\operatorname{HTP}(\mathbb{Q})$-oracle, since $\operatorname{HTP}(\mathbb{Q})$
decides $\operatorname{HTP}(R)$ uniformly for every semilocal ring $R$. The boundary $\mathcal{B}(f)$ of $f$ remains: it contains those $W$ which lie neither in $\mathcal{A}(f)$ nor in $\operatorname{Int}(\mathcal{A}(f))$. The boundary can be empty, but need not be, as seen in the example mentioned above.

It follows quickly from Baire category theory that the boundary set for a polynomial $f \in \mathbb{Z}[\boldsymbol{X}]$ must be nowhere dense. In general the boundary set $\partial \mathcal{A}$ of a set $\mathcal{A}$ within a space $\mathcal{S}$ is defined to equal $(\mathcal{S}-\operatorname{Int}(\mathcal{A})-\operatorname{Int}(\overline{\mathcal{A}}))$, and thus is always closed.

Lemma 3. For every open set $\mathcal{A}$ in a Baire space $\mathcal{S}$, the boundary set $\partial \mathcal{A}$ is nowhere dense. In particular, for each $f \in \mathbb{Z}[\boldsymbol{X}]$, the boundary set $\mathcal{B}(f)=$ $\partial(\mathcal{A}(f))$ must be nowhere dense. Hence the entire boundary set

$$
\mathcal{B}=\{W \subseteq \mathbb{P}:(\exists f \in \mathbb{Z}[\boldsymbol{X}]) W \in \mathcal{B}(f)\}=\cup_{f \in \mathbb{Z}[\boldsymbol{X}]} \mathcal{B}(f)
$$

is meager.
Proof. Since $\mathcal{A}$ is open, every open subset $\mathcal{V}$ of the closure of $\partial \mathcal{A}$ (namely $\partial \mathcal{A}$ itself) lies within the complement $\overline{\mathcal{A}}$, hence within $\operatorname{Int}(\overline{\mathcal{A}})$, which is also disjoint from $\partial \mathcal{A}$. This proves that $\partial \mathcal{A}$ is nowhere dense. Hence $\mathcal{B}$, the countable union of such sets, is meager.

For a set $W$ to fail to lie in $\mathcal{B}$, it must be the case that for every polynomial $f$, either $f \in \operatorname{HTP}\left(R_{W}\right)$ or else some finite initial segment of $W$ rules out all solutions to $f$. This is an example of the concept of genericity, common in both computability and set theory, so we adopt the term here. With this notion, we can show not only that $\operatorname{HTP}\left(R_{W}\right) \leq W \oplus \operatorname{HTP}(\mathbb{Q})$ for all $W$ in the comeager set $\overline{\mathcal{B}}$, but indeed that the reduction is uniform on $\overline{\mathcal{B}}$.

Definition 2. $A$ set $W \subseteq \mathbb{P}$ is HTP-generic if $W \notin \mathcal{B}$. In this case we will also call the corresponding subring $R_{W}$ HTP-generic. By Lemma 3, HTP-genericity is comeager.

Proposition 2. For every HTP-generic set $W, \operatorname{HTP}\left(R_{W}\right) \equiv_{T} W \oplus \operatorname{HTP}(\mathbb{Q})$, via uniform Turing reductions. Hence there is a single Turing reduction $\Phi$ such that the following set is comeager:

$$
\left\{W \subseteq \mathbb{P}: \Phi^{W \oplus H T P(\mathbb{Q})}=\chi_{H T P\left(R_{W}\right)}\right\}
$$

Proof. Given $f \in \mathbb{Z}[\boldsymbol{X}]$ as input, the program for $\Phi$ simply searches for either a solution $\boldsymbol{x}$ to $f=0$ in $\mathbb{Q}$ for which all primes dividing the denominators lie in the oracle set $W$, or else a finite set $A_{0} \subseteq \bar{W}$ such that the $\operatorname{HTP}(\mathbb{Q})$ oracle, using Proposition 1, confirms that $f \notin \operatorname{HTP}\left(R_{\mathbb{P}-A_{0}}\right)$. When it finds either of these, it outputs the corresponding answer about membership of $f$ in $\operatorname{HTP}\left(R_{W}\right)$. If it never finds either, then $W \in \mathcal{B}(f)$, and so this process succeeds for every $W$ except those in the meager set $\mathcal{B}$. (The reduction $W \oplus \operatorname{HTP}(\mathbb{Q}) \leq_{T} \operatorname{HTP}\left(R_{W}\right)$ was described in Sect.1.)

Corollary 1. For every set $C \subseteq \omega$, the following are equivalent

1. $C \leq_{T} \operatorname{HTP}(\mathbb{Q})$.
2. $\left\{W \subseteq \mathbb{P}: C \leq_{T} \operatorname{HTP}\left(R_{W}\right)\right\}=2^{\mathbb{P}}$.
3. $\left\{W \subseteq \mathbb{P}: C \leq_{T} H T P\left(R_{W}\right)\right\}$ is not meager.

This opens a new possible avenue to a proof of undecidability of $\operatorname{HTP}(\mathbb{Q})$ : one need not address $\mathbb{Q}$ itself, but only show that for most subrings $R_{W}, \operatorname{HTP}\left(R_{W}\right)$ can decide the halting problem (or some other fixed undecidable set $C$ ). Constructions in the style of [6, Theorem 1.3] offer an approach to the problem along these lines: that theorem, proven by Poonen, shows that the set of subrings $R$ with $\emptyset^{\prime} \leq_{T} \operatorname{HTP}(R)$ has size continuum and is large in certain other senses. Poonen constructs decidable subsets $T_{0}, T_{1} \subseteq \mathbb{P}$, both of asymptotic density 0 within $\mathbb{P}$, such that for every $W \subseteq \mathbb{P}$ with $T_{0} \subseteq W$ and $T_{1} \cap W=\emptyset$, the subring $R_{W}$ has $\emptyset^{\prime} \leq_{T} \operatorname{HTP}\left(R_{W}\right)$. This feels like a substantial collection of subrings, but the conditions $T_{0} \subseteq W$ and $T_{1} \cap W=\emptyset$ each imply that this set of subrings is nowhere dense, and therefore this set does not by itself enable us to apply Corollary 1. Moreover, it is not clear that any of Poonen's subrings need be HTP-generic.

Proof. Trivially $(1 \Longrightarrow 2 \Longrightarrow 3)$, since all $W$ satisfy $\operatorname{HTP}(\mathbb{Q}) \leq_{T} \operatorname{HTP}\left(R_{W}\right)$. So assume (3). Then by Proposition $2, C \leq_{T} W \oplus \operatorname{HTP}(\mathbb{Q})$ holds on a nonmeager set, as the intersection of a non-meager set with a comeager set cannot be meager. So by Lemma $2, C \leq_{T} \operatorname{HTP}(\mathbb{Q})$.

Having examined classes of subsets of $\mathbb{P}$ defined by Turing reductions involving $\operatorname{HTP}\left(R_{W}\right)$, we now replace Turing reducibility by 1 -reducibility and ask similar questions about classes so defined. It is not known whether there exists a subring $R \subseteq \mathbb{Q}$ for which $\emptyset^{\prime} \leq_{T} \operatorname{HTP}\left(R_{W}\right)$ but $\emptyset^{\prime} \not \leq_{1} \operatorname{HTP}\left(R_{W}\right)$, and we have no good candidates for such a subring. Ever since the original proof of undecidability of Hilbert's Tenth Problem in [1,4], every Turing reduction ever given from the Halting Problem to any $\operatorname{HTP}(R)$ with $R \subseteq \mathbb{Q}$ has in fact been a 1-reduction. Of course, if $\emptyset^{\prime} \leq_{1} \operatorname{HTP}(\mathbb{Q})$, then $\emptyset^{\prime} \leq_{1} \operatorname{HTP}(R)$ for all subrings $R$, so in some sense $\mathbb{Q}$ itself is the "only" candidate.

We have a result for 1-reducibility analogous to Corollary 1, but the proof is somewhat different.

Theorem 1. For every set $C \subseteq \omega$ with $C \not \mathbb{L}_{1} \operatorname{HTP}(\mathbb{Q})$, the following class is meager:

$$
\mathcal{O}=\left\{W \subseteq \mathbb{P}: C \leq_{1} H T P\left(R_{W}\right)\right\}
$$

Proof. One naturally views $\mathcal{O}$ as the union of countably many subclasses $\mathcal{O}_{e}=$ $\left\{W \subseteq \mathbb{P}: C \leq_{1} \operatorname{HTP}\left(R_{W}\right)\right.$ via $\left.\varphi_{e}\right\}$. Of course, for those $e$ for which the $e$-th Turing function $\varphi_{e}$ is not total, this class is empty. We claim that if any one of these $\mathcal{O}_{e}$ fails to be nowhere dense, then $C \leq_{1} \operatorname{HTP}(\mathbb{Q})$, contrary to the assumption of the theorem.

Suppose that indeed $\mathcal{O}_{e}$ fails to be nowhere dense, and fix a $\sigma$ for which $\mathcal{U}_{\sigma} \subseteq \operatorname{cl}\left(\mathcal{O}_{e}\right)$. Let $A_{0}=\sigma^{-1}(0)$ contain those primes excluded from all $W \in \mathcal{U}_{\sigma}$,
and set $R=R_{\left(\mathbb{P}-A_{0}\right)}$. Now whenever $n \in C$ and $W \in \mathcal{O}_{e}$, the polynomial $\varphi_{e}(n)$ must lie in $\operatorname{HTP}\left(R_{W}\right)$. Since some $W \in \mathcal{O}_{e}$ lies in $U_{\sigma}$, we must have $\varphi_{e}(n) \in \operatorname{HTP}(R)$, because $R_{W} \subseteq R$ whenever $W \in \mathcal{U}_{\sigma}$. On the other hand, suppose $n \notin C$. If $R$ contained a solution to the polynomial $\varphi_{e}(n)$, then some $\tau \supseteq \sigma$ would by itself invert the finitely many primes required to generate this solution, and thus we would have $\mathcal{U}_{\tau} \cap \mathcal{O}_{e}=\emptyset$. With $\mathcal{U}_{\sigma} \subseteq \operatorname{cl}\left(\mathcal{O}_{e}\right)$, this is impossible, and so, whenever $n \notin C$, we have $\varphi_{e}(n) \notin \operatorname{HTP}(R)$.

Thus $R$ itself lies in $\mathcal{O}_{e}$, as $\varphi_{e}$ is a 1-reduction from $C$ to $\operatorname{HTP}(R)$. But $R$ is semilocal, inverting all primes $p$ except those with $\sigma(p)=0$. By Proposition 1, we have $\operatorname{HTP}(R) \leq_{1} \operatorname{HTP}(\mathbb{Q})$, and so $C \leq_{1} \operatorname{HTP}(\mathbb{Q})$.

Now we prove two similar results, one about subrings of $\mathbb{Q}$ which admit diophantine models and one about subrings which admit existential definitions of the integers within the subring. In both cases, the result is a sort of zero-one law: that the given phenomenon must either hold almost everywhere (i.e., on a comeager set of subrings) or almost nowhere (i.e., on a meager set). We begin with the diophantine models.

Definition 3. In a ring $R$, a diophantine model of $\mathbb{Z}$ consists of three polynomials $h, h_{+}$, and $h_{\times}$, with $h \in R\left[X_{1}, \ldots, X_{n}, \boldsymbol{Y}\right]$ and $h_{+}, h_{\times} \in R\left[X_{1}, \ldots, X_{3 n}, \boldsymbol{Y}\right]$ (for some $n$ ), such that the set

$$
\left\{\boldsymbol{x} \in R^{n}:\left(\exists \boldsymbol{y} \in R^{<\omega}\right) h(\boldsymbol{x}, \boldsymbol{y})=0\right\}
$$

(equivalently, $\left\{\boldsymbol{x} \in R^{n}: h(\boldsymbol{x}, \boldsymbol{Y}) \in H T P(R)\right\}$ ) is isomorphic to the structure $(\mathbb{Z},+, \cdot)$ under the binary operations whose graphs are defined by

$$
\left\{\left(\boldsymbol{x}_{1}, \boldsymbol{x}_{2}, \boldsymbol{x}_{3}\right) \in R^{3 n}: h_{+}\left(\boldsymbol{x}_{1}, \boldsymbol{x}_{2}, \boldsymbol{x}_{3}, \boldsymbol{Y}\right) \in H T P(R)\right\}
$$

for addition and the corresponding set with $h_{\times}$for multiplication.
If a computable ring $R$ admits a diophantine model of $\mathbb{Z}$, then $\operatorname{HTP}(\mathbb{Z})$ can be coded into $\operatorname{HTP}(R)$, and so $\emptyset^{\prime} \equiv_{1} \operatorname{HTP}(\mathbb{Z}) \leq_{1} \operatorname{HTP}(R)$. For subrings $R_{W}$ of $\mathbb{Q}$ for which $\emptyset^{\prime} \not \mathbb{Z}_{T} W$, this is the only known method of showing that $\emptyset^{\prime} \leq_{T} \operatorname{HTP}\left(R_{W}\right)$ (apart from the original proof by Matiyasevich, Davis, Putnam, and Robinson for the case $W=\emptyset$, of course, which is what allows this method to succeed).

Definition 4. $\mathcal{D}=\left\{W \subseteq \mathbb{P}: R_{W}\right.$ admits a diophantine model of $\left.\mathbb{Z}\right\}$.
In this section we address the question of the size of the class $\mathcal{D}$. The main result fails to resolve this question, but shows it to have an "all-or-nothing" character.

Theorem 2. The class $\mathcal{D}$ is non-meager if and only if there exists a particular triple $\left(h, h_{+}, h_{\times}\right)$of polynomials over $\mathbb{Z}$ and a finite binary string $\sigma \in 2^{<\mathbb{P}}$ such that, for every HTP-generic $V \in \mathcal{U}_{\sigma}, R_{V}$ admits a diophantine model of $\mathbb{Z}$ via these three polynomials.

Moreover, if $\mathcal{D}$ is non-meager, then $\mathbb{P} \in \mathcal{D}$ (i.e., $\mathbb{Q}$ admits a diophantine model of $\mathbb{Z}$ ).

Proof. For each triple $\boldsymbol{h}=\left(h, h_{+}, h_{\times}\right)$of polynomials of appropriate lengths over $\mathbb{Z}$, we set $\mathcal{D}_{\boldsymbol{h}}$ to contain those $W$ for which $\boldsymbol{h}$ defines a diophantine model of $\mathbb{Z}$ within $R_{W}$. If each $\mathcal{D}_{\boldsymbol{h}}$ is nowhere dense, their countable union $\mathcal{D}$ is meager.

Now suppose that $\mathcal{D}$ is non-meager, so some class $\mathcal{D}_{\boldsymbol{h}}$ fails to be nowhere dense. Then there must be a string $\sigma$ such that $\mathcal{U}_{\sigma} \subseteq \operatorname{cl}\left(\mathcal{D}_{\boldsymbol{h}}\right)$. Using this $\sigma$ and this $\boldsymbol{h}$, we now prove the main claim: all $W \in \mathcal{U}_{\sigma}$ with HTP-generic $R_{W}$ lie in $\mathcal{D}_{\boldsymbol{h}}$. Let $R_{0}=R_{\sigma^{-1}(1)}$ and $R_{1}=R_{\mathbb{P}-\sigma^{-1}(0)}$ be the smallest and largest subrings (under $\subseteq$ ) in $\mathcal{U}_{\sigma}$, so $R_{0}$ is finitely generated and $R_{1}$ is semilocal.

Fix a single $W \supset \sigma$ with $W \in \mathcal{D}_{\boldsymbol{h}}$, and fix the tuples $\boldsymbol{x}_{0}$ and $\boldsymbol{x}_{1}$ from $R_{W}$ which represent the elements 0 and 1 in the diophantine model defined in $R_{W}$ by $\boldsymbol{h}$. It follows that $h_{\times}\left(\boldsymbol{x}_{0}, \boldsymbol{x}_{0}, \boldsymbol{x}_{0}, \boldsymbol{Y}\right) \in \operatorname{HTP}\left(R_{W}\right)$ and $h_{\times}\left(\boldsymbol{x}_{1}, \boldsymbol{x}_{1}, \boldsymbol{x}_{1}, \boldsymbol{Y}\right) \in$ $\operatorname{HTP}\left(R_{W}\right)$. Now if any other tuple $\boldsymbol{x}$ from $R_{1}$ had $h(\boldsymbol{x}, \boldsymbol{Y}) \in \operatorname{HTP}\left(R_{1}\right)$ and $h_{\times}(\boldsymbol{x}, \boldsymbol{x}, \boldsymbol{x}, \boldsymbol{Y}) \in \operatorname{HTP}\left(R_{1}\right)$, then we could set $\tau=\sigma^{\wedge} 111 \cdots 1$ to contain enough primes that $R_{\tau^{-1}(1)}$ would contain $\boldsymbol{x}, \boldsymbol{x}_{0}$, and $\boldsymbol{x}_{1}$. This would mean that $\boldsymbol{h}$ could not define a diophantine model of $\mathbb{Z}$ in any $R_{V}$ with $V \in \mathcal{U}_{\tau}$, contrary to hypothesis. Therefore, no other $\boldsymbol{x}$ from $R_{1}$ can do this. Now suppose that $\boldsymbol{x}_{0}$ does not lie within $R_{0}$. In this case, some extension $\rho=\sigma^{\wedge} 000 \cdots 0$ would exclude enough primes to ensure that $\boldsymbol{x}_{0}$ does not lie in $R_{\mathbb{P}-\rho^{-1}(0)}$, and then no $\tau \supseteq \rho$ would admit a diophantine model via $\boldsymbol{h}$, since no other tuple with the right properties lies in $R_{1}$. Again, this contradicts our hypothesis that $\mathcal{U}_{\sigma} \subseteq \operatorname{cl}\left(\mathcal{D}_{\boldsymbol{h}}\right)$, since $\mathcal{D}_{\boldsymbol{h}} \cap \mathcal{U}_{\rho}$ would be empty, and so $\boldsymbol{x}_{0}$ lies in $R_{0}$. Similarly so does $\boldsymbol{x}_{1}$.

Now one proceeds by induction on the subsequent elements of the diophantine model in $R_{1}$. Some tuple $\boldsymbol{x}_{2}$ from $R_{W}$ must satisfy $h\left(\boldsymbol{x}_{2}, \boldsymbol{Y}\right) \in \operatorname{HTP}\left(R_{W}\right)$ and $h_{+}\left(\boldsymbol{x}_{1}, \boldsymbol{x}_{1}, \boldsymbol{x}_{2}, \boldsymbol{Y}\right) \in \operatorname{HTP}\left(R_{W}\right)$, and by the same arguments as above, we see that $\boldsymbol{x}_{2}$ is the only tuple in $R_{1}$ with this property, and then that $\boldsymbol{x}_{2}$ actually lies in $R_{0}$. Likewise, $\boldsymbol{x}_{-1}$ must satisfy $h\left(\boldsymbol{x}_{-1}, \boldsymbol{Y}\right) \in \operatorname{HTP}\left(R_{W}\right)$ and $h_{+}\left(\boldsymbol{x}_{1}, \boldsymbol{x}_{-1}, \boldsymbol{x}_{0}, \boldsymbol{Y}\right) \in$ $\operatorname{HTP}\left(R_{W}\right)$, and again this forces $\boldsymbol{x}_{-1}$ to lie in $R_{0}$ and to be the unique tuple with these properties in $R_{1}$.

Continuing this induction, we see that every tuple in the domain of the diophantine model of $\mathbb{Z}$ in $R_{W}$ actually lies in $R_{0}$, and hence in every $R_{W}$ with $W \in \mathcal{U}_{\sigma}$; and moreover that these are the only tuples $\boldsymbol{x}$ in $R_{1}$ for which $h(\boldsymbol{x}, \boldsymbol{Y}) \in \operatorname{HTP}\left(R_{1}\right)$. Likewise, if some $\boldsymbol{x}_{m}, \boldsymbol{x}_{n}$ and $\boldsymbol{x}_{p}$ (representing $m, n$, and $p$ in the diophantine model) satisfy $h_{+}\left(\boldsymbol{x}_{m}, \boldsymbol{x}_{n}, \boldsymbol{x}_{p}, \boldsymbol{Y}\right) \in \operatorname{HTP}\left(R_{1}\right)$, then for some $k, \tau=\sigma^{\wedge} 1^{k}$ is long enough to ensure that every $W$ extending $\tau$ must have $h_{+}\left(\boldsymbol{x}_{m}, \boldsymbol{x}_{n}, \boldsymbol{x}_{p}, \boldsymbol{Y}\right) \in \operatorname{HTP}\left(R_{W}\right)$. But some such $W$ lies in $\mathcal{D}_{\boldsymbol{h}}$, so we must have $m+n=p$. The same works for $h_{\times}$, so $\boldsymbol{h}$ defines a diophantine model of $\mathbb{Z}$ in $R_{1}$.

It is not clear whether $\boldsymbol{h}$ defines a diophantine model in the subring $R_{0}$ (which, being finitely generated, lies in $\mathcal{B}$ ). The domain elements of the model in $R_{1}$ all lie in $R_{0}$, but the witnesses might not. However, suppose that $V \in \mathcal{U}_{\sigma}$ is HTP-generic, and fix any domain element $\boldsymbol{x}$. Let $\tau=V \upharpoonright m$, for any $m \geq|\sigma|$. Then some $U \supseteq \tau$ lies in $\mathcal{D}_{\boldsymbol{h}}$, and so some extension of $\tau$ yields a solution to $h(\boldsymbol{x}, \boldsymbol{Y})$. Since $V$ is HTP-generic (that is, $V \notin \mathcal{B})$, this forces $h(\boldsymbol{x}, \boldsymbol{Y}) \in$ $\operatorname{HTP}\left(R_{V}\right)$. Likewise, for each fact coded by $h_{+}$or $h_{\times}$about domain elements of the model, some extension of $V \upharpoonright m$ must yield a witness to that fact, and therefore
$R_{V}$ itself contains such a witness. So $\boldsymbol{h}$ defines this same diophantine model in every HTP-generic subring $R_{V}$ with $V \in \mathcal{U}_{\sigma}$, as required by the theorem.

Cases (1) and (2) of the theorem cannot both hold, because under (2), $\mathcal{U}_{\sigma} \cap \overline{\mathcal{B}}$ would be a nonmeager subset of $\mathcal{D}$. Moreover, the 1-reduction $\operatorname{HTP}\left(R_{1}\right) \leq_{1}$ $\operatorname{HTP}(\mathbb{Q})$ given in $[2$, Proposition 5.4] has sufficient uniformity that the images of $h, h_{+}$, and $h_{\times}$under this reduction define a diophantine model of $\mathbb{Z}$ inside $\mathbb{Q}$. (Specifically, $h(\boldsymbol{X}, \boldsymbol{Y})$ maps to the sum of $h^{2}$ with several other squares of polynomials in such a way as to guarantee that all solutions use values from $R_{1}$ for the variables $\boldsymbol{X}$ and $\boldsymbol{Y}$; likewise with $h_{+}$and $h_{\times}$.) This proves the final statement of the theorem.

Now we continue with the question of existential definability of the integers.
Definition 5. In a ring $R$, a polynomial $g \in \mathbb{Z}[X, \boldsymbol{Y}]$ existentially defines $\mathbb{Z}$ if, for every $q \in R$,

$$
q \in \mathbb{Z} \Longleftrightarrow g(q, \boldsymbol{Y}) \in H T P(R)
$$

$\mathbb{Z}$ is existentially definable in $R$ if such a polynomial $g$ exists.
A ring in which $\mathbb{Z}$ is existentially definable must admit a very simple diophantine model of $\mathbb{Z}$, given by the polynomial $g$ along with $h_{+}=X_{1}+X_{2}-X_{3}$ and $h_{\times}=X_{1} X_{2}-X_{3}$. The question of definability of $\mathbb{Z}$ in the field $\mathbb{Q}$ was originally answered by Julia Robinson (see [8]), who gave a $\Pi_{4}$ definition. Subsequent work by Poonen [7] and then Koenigsmann [3] has resulted in a $\Pi_{1}$ definition of $\mathbb{Z}$ in $\mathbb{Q}$, but it remains unknown whether any existential formula defines $\mathbb{Z}$ there.

Definition 6. $\mathcal{E}$ is the class of subrings of $\mathbb{Q}$ where $\mathbb{Z}$ is existentially definable:

$$
\mathcal{E}=\left\{W \subseteq \mathbb{P}: \mathbb{Z} \text { is existentially definable in } R_{W}\right\} .
$$

We now address the question of the size of the class $\mathcal{E}$. As with $\mathcal{D}$, we show $\mathcal{E}$ to be either very large or very small, in the sense of Baire category.

Theorem 3. The following are equivalent.

1. The class $\mathcal{E}$ is not meager.
2. There is a $\sigma \in 2^{<\mathbb{P}}$, and a single polynomial $g$ which existentially defines $\mathbb{Z}$ in all HTP-generic subrings $R_{V}$ with $V \in \mathcal{U}_{\sigma}$.
3. $\mathbb{P} \in \mathcal{E}$; that is, $\mathbb{Z}$ is existentially definable in $\mathbb{Q}$.
4. There is a single existential formula which defines $\mathbb{Z}$ in every subring of $\mathbb{Q}$.

Proof. The proof that $(1) \Longrightarrow(2) \Longrightarrow(3)$ proceeds along the same lines as that of Theorem 2, with $\mathcal{E}_{g}$ as the class of those $W$ for which the polynomial $g$ existentially defines $\mathbb{Z}$ within $R_{W}$. If every one of these classes is nowhere dense, then their countable union $\mathcal{E}$ is meager. Otherwise one proves (2), and from that (3), by a simplification of the same method as before, with no induction required. To see that (3) implies (4), notice that if $\mathbb{Z}$ is defined in $\mathbb{Q}$ by the formula $\exists \boldsymbol{Y} f(X, \boldsymbol{Y})=0$, and $d$ is the total degree of $f$, then the formula

$$
\exists \boldsymbol{Y} \exists Z\left[Z^{d} \cdot f\left(X, \frac{Y_{1}}{Z}, \ldots, \frac{Y_{n}}{Z}\right)=0 \& Z>0\right]
$$

defines $\mathbb{Z}$ in $R_{W}$.

It is possible to turn Theorem 2 into an equivalence analogous to that in Theorem 3, with the third condition stating that $\mathbb{P} \in \mathcal{D}$. As far as we know, however, it is necessary to consider diophantine interpretations in subrings $R_{W}$, rather than diophantine models, in order to accomplish this.

## 3 Measure Theory

Normally there is a strong connection between measure theory and Baire category theory. Each defines a certain $\Sigma$-ideal of sets to be "small": the sets of measure 0, and the meager sets. In Cantor space, neither of these properties implies the other, but empirically they appear closely connected, especially when the sets are given by natural definitions: sets of measure 0 are often meager, and vice versa. (Exceptions to this principle do exist, however, and another difference was mentioned in the context of Lemma 1.)

Our results here rely heavily on the simple Lemma 3, stating that the boundary set $\mathcal{B}(f)$ of a polynomial $f$ is nowhere dense. Most of our subsequent results have measure-theoretic analogues which would go through fairly easily, provided that these sets $\mathcal{B}(f)$ also have measure 0 . However, determining the measure of the boundary set of a polynomial appears to be a nontrivial problem. It is unknown whether there exists any polynomial $f$ for which $\mu(\mathcal{B}(f))>0$. Indeed, in work to appear elsewhere, the author has shown that if $\mu(\mathcal{B}(f))=0$ for all $f \in \mathbb{Z}[\boldsymbol{X}]$, then there is no existential definition of the set $\mathbb{Z}$ within the field $\mathbb{Q}$.

Moreover, if an $f$ exists with $\mu(\mathcal{B}(f))>0$, it is unclear what other constraints on the real number $\mu(\mathcal{B}(f))$ exist, apart from the computability-theoretic upper bound given by its definition as $\mu(\mathcal{B}(f))$. Could such a number be transcendental? Or noncomputable? If not, is there an algorithm computing $\mu(\mathcal{B}(f))$ uniformly in $f$ ? These appear to be challenging questions, often with a more number-theoretic flavor than most of this article. Resolving them might make it possible to determine whether Hilbert's Tenth Problem on subrings of $\mathbb{Q}$ has measure-theoretic zero-one laws similar to those proven here for Baire category.

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# Partial Orders and Immunity in Reverse Mathematics 

Ludovic Patey ${ }^{(\boxtimes)}$<br>Laboratoire PPS, Université Paris Diderot, Paris, France<br>ludovic.patey@computability.fr


#### Abstract

We identify computability-theoretic properties enabling us to separate various statements about partial orders in reverse mathematics. We obtain simpler proofs of existing separations, and deduce new compound ones. This work is part of a larger program of unification of the separation proofs of various Ramsey-type theorems in reverse mathematics in order to obtain a better understanding of the combinatorics of Ramsey's theorem and its consequences. We also answer a question of Murakami, Yamazaki and Yokoyama about pseudo Ramsey's theorem for pairs.


## 1 Introduction

Many theorems of "ordinary" mathematics are of the form

$$
(\forall X)[\Phi(X) \rightarrow(\exists Y) \Psi(X, Y)]
$$

where $\Phi$ and $\Psi$ are arithmetic formulas. They can be seen as mathematical problems, whose instances are sets $X$ such that $\Phi(X)$ holds, and whose solutions to $X$ are sets $Y$ such that $\Psi(X, Y)$ holds. For example, König's lemma asserts that every infinite, finitely branching tree admits an infinite path through it.

There exist many ways to calibrate the strength of a mathematical problem. Among them, reverse mathematics is a vast foundational program that seeks to determine the weakest axioms necessary to prove ordinary theorems. It uses the framework of subsystems of second-order arithmetic, within the base theory $\mathrm{RCA}_{0}$, which can be thought of as capturing computable mathematics. An $\omega$ structure is a structure whose first-order part consists of the standard integers. The $\omega$-models of $\mathrm{RCA}_{0}$ are those whose second-order part is a Turing ideal, that is, a collection of sets $\mathcal{S}$ downward-closed under the Turing reduction and closed under the effective join.

In this setting, a $\omega$-model $\mathcal{M}$ satisfies a mathematical problem P if every P instance in $\mathcal{M}$ has a solution in $\mathcal{M}$. A standard way of proving that a problem P does not imply another problem $Q$ consists of creating an $\omega$-model $\mathcal{M}$ satisfying $P$ but not $Q$. Such a model is usually constructed by taking a ground Turing ideal, and extending it by iteratively adding solutions to its P-instances. However, while taking the closure of the collection $\mathcal{M} \cup\{Y\}$ to obtain a Turing ideal, one may
add solutions to Q-instances as well. The whole difficulty of this construction consists of finding the right computability-theoretic notion preserved by P but not by Q .

We conduct a program of identification of the computability-theoretic properties enabling us to distinguish various Ramsey-type theorems in reverse mathematics, but also under computable and Weihrauch reducibilities. This program puts emphasis on the interplay between computability theory and reverse mathematics, the former providing tools to separate theorems in reverse mathematics over standard models, and the latter exhibiting new computability-theoretic properties.

Among the theorems studied in reverse mathematics, the ones coming from Ramsey's theory play a central role. Their strength are notoriously hard to gauge, and required the development of involved computability-theoretic frameworks. Perhaps the most well-known example is Ramsey's theorem.

Definition 1 (Ramsey's theorem). A subset $H$ of $\omega$ is homogeneous for a coloring $f:[\omega]^{n} \rightarrow k$ (or $f$-homogeneous) if each $n$-tuples over $H$ are given the same color by $f . \mathrm{RT}_{\mathrm{k}}^{\mathrm{n}}$ is the statement "Every coloring $f:[\omega]^{n} \rightarrow k$ has an infinite $f$-homogeneous set".

Jockusch [11] conducted a computational analysis of Ramsey's theorem. He proved in particular that $R T_{k}^{n}$ implied the existence of the halting set whenever $n \geq 3$. There has been a lot of literature around the strength of Ramsey's theorem for pairs $[4,6,9,19]$ and its consequences $[3,5,10]$. We focus on some mathematical statements about partial orders which are consequences of Ramsey's theorem for pairs.

Definition 2 (Chain-antichain). A chain in a partial order $\left(P, \leq_{P}\right)$ is a set $S \subseteq P$ such that $(\forall x, y \in S)\left(x \leq_{P} y \vee y \leq_{P} x\right)$. An antichain in $P$ is a set $S \subseteq P$ such that $(\forall x, y \in S)\left(x \neq\left. y \rightarrow x\right|_{P} y\right)$ (where $\left.x\right|_{P} y$ means that $\left.x \not \leq_{P} y \wedge y \not \mathbb{Z}_{P} x\right)$. CAC is the statement "every infinite partial order has an infinite chain or an infinite antichain."

The chain-antichain principle was introduced by Hirschfeldt and Shore [10] together with the ascending descending sequence (ADS). They studied extensively cohesive and stable versions of the statements, and proved that CAC is computationally weak, in that it does not even imply the existence of a diagonally non-computable function. However, their proof has an ad-hoc flavor, in that it is a direct separation involving the two statements. Later, Lerman et al. [13] separated ADS from CAC over $\omega$-models by using an involved iterated forcing argument.

In this paper, we revisit the two proofs and emphasis on the combinatorial nature of the principles by identifying the computability-theoretic properties separating them. Those properties happen to be very natural and coincide on coc.e. sets to some well-known computability-theoretic notions, namely, immunity and hyperimmunity. The proof of the separation of ADS from CAC is significantly simpler and more modular, as advocated by the author in [16].

### 1.1 Notation and Definitions

Given two sets $A$ and $B$, we denote by $A<B$ the formula $(\forall x \in A)(\forall y \in B)[x<$ $y]$ and by $A \subseteq^{*} B$ the formula $\left(\forall^{\infty} x \in A\right)[x \in B]$, meaning that $A$ is included in $B$ up to finitely many elements. A Mathias condition is a pair $(F, X)$ where $F$ is a finite set, $X$ is an infinite set and $F<X$. A condition $\left(F_{1}, X_{1}\right)$ extends $(F, X)\left(\right.$ written $\left.\left(F_{1}, X_{1}\right) \leq(F, X)\right)$ if $F \subseteq F_{1}, X_{1} \subseteq X$ and $F_{1} \backslash F \subset X$. A set $G$ satisfies a Mathias condition $(F, X)$ if $F \subset G$ and $G \backslash F \subseteq X$.

## 2 Preservation of Properties for Co-c.e. Sets

Ramsey's theorem for $k$ colors has a deeply disjunctive nature. One cannot know in a finite amount of time whether a coloring will admit an infinite homogeneous set for a fixed color, and one must therefore build multiple homogeneous sets simultaneously, namely, one for each color. This disjunction was exploited by the author to show for example that ADS does not preserve 2 hyperimmunities simultaneously, whereas the Erdős-Moser theorem does [16]. This idea was also used in the context of computable reducibility to show that $R T_{k+1}^{2}$ does not computably reduce to $\mathrm{RT}_{\mathrm{k}}^{2}$ whenever $k \geq 1$, by showing that $\mathrm{R} T_{k}^{2}$ preserves 2 among $k+1$ hyperimmunities simultaneously whereas $\mathrm{RT}_{\mathrm{k}+1}^{2}$ does not [18]. In this section, we shall see that this disjunctive flavor disappears whenever considering co-c.e. sets. In particular, $\mathrm{RT}_{2}^{2}$ admits preservation of countably many hyperimmune co-c.e. sets simultaneously.

Definition 3 (Hyperimmunity). An array is a sequence of mutually disjoint finitely coded sets. $A$ set $A$ is $X$-hyperimmune if for every $X$-c.e. array $F_{0}, F_{1}, \ldots$, there is some $i$ such that $F_{i} \cap A=\emptyset$.

Equivalently, a set is $X$-hyperimmune if its principal function is not dominated by any $X$-computable function, where the principal function $p_{A}$ of a set $A=\left\{x_{0}<x_{1}<\ldots\right\}$ is defined by $p_{A}(i)=x_{i}$.

Definition 4 (Preservation of hyperimmunity for co-c.e. sets). $A \Pi_{2}^{1}$ statement P admits preservation of hyperimmunity for co-c.e. sets if for every set $Z$, every sequence of $Z$-co-c.e. $Z$-hyperimmune sets $A_{0}, A_{1}, \ldots$ and every P-instance $X \leq_{T} Z$, there is a solution $Y$ to $X$ such that the $A$ 's are $Y \oplus Z$ hyperimmune.

Hirschfeldt and Shore [10] proved that CAC is equivalent to the existence of homogeneous sets for semi-transitive colorings. A coloring $f:[\mathbb{N}]^{2} \rightarrow 2$ is semi-transitive if whenever $f(x, y)=1$ and $f(y, z)=1$, then $f(x, z)=1$ for $x<y<z$.

Theorem 5. CAC admits preservation of hyperimmunity for co-c.e. sets.

Proof. Fix a set $Z$ and a countable sequence of $Z$-co-c.e. $Z$-hyperimmune sets $A_{0}, A_{1}, \ldots$ Let $f:[\omega]^{2} \rightarrow 2$ be a $Z$-computable semi-transitive coloring. We shall assume that there is no infinite $Z$-computable $f$-homogeneous set for color 0 , otherwise we are done. We will build an infinite set $G f$-homogeneous for color 1 such that the $A$ 's are $G \oplus Z$-hyperimmune. The construction is done by a Mathias forcing $(F, X)$, where $F$ is a finite set, $X$ is an infinite $Z$-computable set such that $\max (F)<\min (X)$, and for every $x \in X, F \cup\{x\}$ is $f$-homogeneous for color 1 . The condition extension is the usual Mathias extension. A set $G$ satisfies $(F, X)$ if it satisfies the Mathias condition $(F, X)$ and is $f$-homogeneous for color 1 . Lemma 6 shows that every sufficiently generic filter for this notion of forcing yields an infinite set.

Lemma 6. Every condition $c=(F, X)$ has an extension $(E, Y)$ such that $|E|>|F|$.

In what follows, we say that a condition $c$ forces a formula property $\varphi(G)$ if $\varphi(G)$ holds for every set $G$ satisfying $c$.

Lemma 7. For every condition $c=(F, X)$ and every pair of indices e, $i$, there is an extension forcing $\Phi_{e}^{G \oplus Z}$ not to dominate $p_{A_{i}}$.

Proof. Define the $Z$-partial computable function $h$ which on input $x$, searches for a finite set $E_{x} \subseteq X f$-homogeneous for color 1 such that $\Phi_{e}^{\left(F \cup E_{x}\right) \oplus Z}(x) \downarrow$. If found, $h(x)=\Phi_{e}^{\left(F \cup E_{x}\right) \oplus Z}(x)$, otherwise $h(x) \uparrow$. We have two cases.

- Case 1: $h$ is total. By $Z$-hyperimmunity of $p_{A_{i}}$, there are infinitely many $x$ such that $h(x)<p_{A_{i}}(x)$. If there is such an $x$ such that the set $Y=\{y \in$ $\left.X:\left(\forall z \in E_{x}\right) f(z, y)=1\right\}$ is infinite, then the condition $\left(F \cup E_{x}, Y\right)$ is an extension of $c$ forcing $\Phi_{e}^{G \oplus Z}(x)<p_{A_{i}}(x)$. If there is no such $x$, then by semitransitivity of $f$, for every $x$ such that $h(x)<p_{A_{i}}(x)$, for almost every $y \in X$, $f\left(\max \left(E_{x}\right), y\right)=0$. Since $A_{i}$ is co-c.e., one can find a $Z$-computable infinite subset $Y$ of $\left\{\max \left(E_{x}\right): h(x)<p_{A_{i}}(x)\right\}$. The set $Y$ is $Z$-computable and limithomogeneous for color 0 , and therefore computes an infinite $f$-homogeneous set for color 0 , contradicting our assumption.
- Case 2: there is some $x$ such that $h(x) \uparrow$. By definition of $h$, the condition $c$ already forces $\Phi_{e}^{G \oplus Z}(x) \uparrow$.

Corollary 8. $\mathrm{RT}_{2}^{2}$ admits preservation of hyperimmunity for co-c.e. sets.
Proof. Bovykin and Weiermann [2] studied the reverse mathematics of the Erdős-Moser theorem (EM) and proved that $\mathrm{RCA}_{0} \vdash \mathrm{RT}_{2}^{2} \leftrightarrow[\mathrm{CAC} \wedge E M]$. The author proved in [16] that EM admits preservation of hyperimmunity. Together with Theorem 5, we deduce that $C A C \wedge E M$, hence $\mathrm{RT}_{2}^{2}$, admits preservation of hyperimmunity for co-c.e. sets.

## 3 CAC and Constant-Bound Immunity

Hirschfeldt and Shore [10] separated CAC from DNC in reverse mathematics by a direct construction. DNC is the statement asserting, for every set $X$, the existence of a function $f$ such that $f(e) \neq \Phi_{e}^{X}(e)$ for every $e$. In this section, we extract the core of the combinatorics of their forcing argument to exhibit a computability-theoretic property separating the two notions, namely, constantbound immunity.

Definition 9 (Constant-bound immunity). $A k$-enumeration ( $k$-enum) of a set $A$ is an infinite sequence of $k$-sets $F_{0}<F_{1}<\ldots$ such that for every $i \in$ $\omega, F_{i} \cap A \neq \emptyset . A$ constant-bound enumeration (c.b-enum) of a set $A$ is a $k$ enumeration of $A$ for some $k \in \omega$. $A$ set $A$ is $k$-immune (c.b-immune) relative to $X$ if it admits no $X$-computable $k$-enumeration (c.b-enumeration).

In particular, 1-immunity coincides with the standard notion of immunity. Also note that one can easily create a c.b-immune set computing no effectively immune set. The following lemma shows that c.b-immunity and immunity coincide for co-c.e. sets.

Lemma 10. An $X$-co-c.e. set $A$ is c.b-immune relative to $X$ iff it is $X$-immune.
Definition 11 (Preservation of c.b-immunity). $A \Pi_{2}^{1}$ statement P admits preservation of c.b-immunity if for every set $Z$, every set $A$ which is c.b-immune relative to $X$, and every P -instance $X \leq_{T} Z$, there is a solution $Y$ to $X$ such that $A$ is c.b-immune relative to $Y \oplus Z$.

We can easily relate the notion of preservation of c.b-immunity with the existing notion of constant-bound enumeration avoidance defined by Liu [14] to separate $R T_{2}^{2}$ from $W W K L$ over $R C A_{0}$.

Lemma 12. If P admits preservation of $c . b-i m m u n i t y$, then it admits constantbound enumeration avoidance.

Theorem 13. CAC admits preservation of c.b-immunity.
Proof. Let $A$ be a set c.b-immune relative to some set $Z$, and let $f:[\omega]^{2} \rightarrow 2$ be a $Z$-computable semi-transitive coloring. Assume that there is no infinite $f$ homogeneous set $H$ such that $A$ is c.b-immune relative to $H \oplus Z$, otherwise we are done. We will build two infinite sets $G_{0}$ and $G_{1}$, such that $G_{i}$ is $f$-homogeneous for color $i$ for each $i<2$, and such that $A$ is c.b-immune relative to $G_{i} \oplus Z$ for some $i<2$.

The construction is done by a variant of Mathias forcing $\left(F_{0}, F_{1}, X\right)$, where $F_{0}$ and $F_{1}$ are finite sets, $X$ is infinite set such that $\max \left(F_{0}, F_{1}\right)<\min (X)$, and $A$ is c.b-immune relative to $X \oplus Z$. Moreover, we require that for every $i<2$ and every $x \in X, F_{i} \cup\{x\}$ is $f$-homogeneous for color $i$. A condition $\left(E_{0}, E_{1}, Y\right)$ extends $\left(F_{0}, F_{1}, X\right)$ if ( $E_{i}, Y$ ) Mathias extends $\left(F_{i}, X\right)$ for each $i<2$. A pair of sets $G_{0}, G_{1}$ satisfies a condition $c=\left(F_{0}, F_{1}, X\right)$ if $G_{i}$ is $f$-homogeneous for color $i$ and satisfies the Mathias condition $\left(F_{i}, X\right)$ for each $i<2$.

Lemma 14. For every condition $c=\left(F_{0}, F_{1}, X\right)$ and every $i<2$, there is an extension $\left(E_{0}, E_{1}, Y\right)$ of $c$ such that $\left|E_{i}\right|>\left|F_{i}\right|$.

In what follows, we interpret $\Phi_{0}, \Phi_{1}, \ldots$ as Turing functionals outputting non-empty finite sets such that if $\Phi_{e}^{X}(x)$ and $\Phi_{e}^{X}(x+1)$ both halt, $\max \left(\Phi_{e}^{X}(x)\right)<\min \left(\Phi_{e}^{X}(x+1)\right)$. We want to satisfy the following requirements for each $e_{0}, k_{0}, e_{1}, k_{1} \in \omega$ :

$$
\mathcal{R}_{e_{0}, k_{0}, e_{1}, k_{1}}: \quad \mathcal{R}_{e_{0}, k_{0}}^{G_{0}} \quad \vee \quad \mathcal{R}_{e_{1}, k_{1}}^{G_{1}}
$$

where $\mathcal{R}_{e, k}^{G}$ is the requirement

$$
(\exists x)\left(\Phi_{e}^{G \oplus Z}(x) \uparrow \vee\left|\Phi_{e}^{G \oplus Z}(x)\right|>k \vee \Phi_{e}^{G \oplus Z}(x) \cap A=\emptyset\right)
$$

In other words, $\mathcal{R}_{e, k}^{G}$ asserts that $\Phi_{e}^{G \oplus Z}$ is not a $k$-enumeration of $A$. A condition $c$ forces a formula $\varphi\left(G_{0}, G_{1}\right)$ if $\varphi\left(G_{0}, G_{1}\right)$ holds for every pair of infinite sets $G_{0}, G_{1}$ satisfying $c$.

Lemma 15. For every condition $c$ and every vector of indices $e_{0}, k_{0}, e_{1}, k_{1} \in \omega$, there is an extension $d$ of $c$ forcing $\mathcal{R}_{e_{0}, k_{0}, e_{1}, k_{1}}$.

Proof. Fix a condition $c=\left(F_{0}, F_{1}, X\right)$, and let $P_{0}, P_{1}, \ldots$ be an $X \oplus Z$ computable sequence of sets where $P_{n}=\Phi_{e_{0}}^{\left(F_{0} \cup E_{0}\right) \oplus Z}\left(x_{0}\right) \cup \Phi_{e_{1}}^{\left(F_{1} \cup E_{1}\right) \oplus Z}\left(x_{1}\right)$ for a pair of sets $E_{1}<E_{0} \subseteq X$ and some $x_{0}, x_{1} \in \omega$ such that $E_{0}$ is $f$-homogeneous for color $0, E_{1} \cup\{y\}$ is $f$-homogeneous for color 1 for each $y \in E_{0}$, and for each $i<2, \max \left(P_{n-1}\right)<\min \left(\Phi_{e_{i}}^{\left(F_{i} \cup E_{i}\right) \oplus Z}\left(x_{i}\right)\right)$ and $\left|\Phi_{e_{i}}^{\left(F_{i} \cup E_{i}\right) \oplus Z}\left(x_{i}\right)\right| \leq k_{i}$. We have two cases.

- Case 1: the sequence of the $P$ 's is finite and is defined, say to level $n-1$. If there is a pair of infinite sets $G_{0}, G_{1}$ satisfying $c$ and some $x_{1} \in \omega$ such that $\Phi_{e_{1}}^{G_{1} \oplus Z}\left(x_{1}\right) \downarrow, \max \left(P_{n-1}\right)<\min \left(\Phi_{e_{1}}^{G_{1} \oplus Z}\left(x_{1}\right)\right)$, and $\left|\Phi_{e_{1}}^{G_{1} \oplus Z}\left(x_{1}\right)\right| \leq k_{1}$, then let $E_{1} \subseteq G_{1}$ be such that $F_{1} \cup E_{1}$ is an initial segment of $G_{1}$ for which $\Phi_{e_{1}}^{\left(F_{1} \cup E_{1}\right) \oplus Z}\left(x_{1}\right) \downarrow$. The set $Y=\left\{y \in X: E_{1} \cup\{y\}\right.$ is $f$-homogeneous for color 1$\}$ is a superset of $G_{1}$, hence is infinite. The condition $d=\left(F_{0}, F_{1} \cup E_{1}, Y\right)$ is an extension of $c$ forcing $\mathcal{R}_{e_{0}, k_{0}}^{G_{0}}$, hence forcing $\mathcal{R}_{e_{0}, k_{0}, e_{1}, k_{1}}$. If there is no such pair of infinite sets $G_{0}, G_{1}$, then the condition $c$ already forces $\mathcal{R}_{e_{1}, k_{1}}^{G_{1}}$, hence $\mathcal{R}_{e_{0}, k_{0}, e_{1}, k_{1}}$.
- Case 2: the sequence of the $P$ 's is infinite. By c.b-immunity of $A$ relative to $X \oplus Z, P_{n} \cap A=\emptyset$ for some $n \in \omega$. Let $E_{1}<E_{0} \subseteq X$ and $x_{0}, x_{1} \in \omega$ witness the existence of $P_{n}$. If $Y_{0}=\left\{y \in X: E_{0} \cup\{y\}\right.$ is $f$-homogeneous for color 1\} is infinite, then the condition ( $F_{0} \cup E_{0}, F_{1}, Y_{0}$ ) is an extension of $c$ forcing $\mathcal{R}_{e_{0}, k_{0}}^{G_{0}}$. If $Y_{0}$ is finite, then for almost every $y \in X$, there is some $x_{y} \in E_{0}$ such that $f\left(x_{y}, y\right)=1$, and by transitivity of $f$ for color $1, E_{1} \cup\{y\}$ is $f$ homogeneous for color 1 . Indeed, $E_{1}$ is $f$-homogeneous for color 1 and for each $x \in E_{1}, f\left(x, x_{y}\right)=f\left(x_{y}, y\right)=1$. In this case, $\left(F_{0}, F_{1} \cup E_{1}, Y_{1}\right)$ is an extension of $c$ forcing $\mathcal{R}_{e_{1}, k_{1}}^{G_{1}}$, for some $Y_{1}={ }^{*} X$. In both cases, there is an extension of $c$ forcing $\mathcal{R}_{e_{0}, k_{0}, e_{1}, k_{1}}$.

This completes the proof of Theorem 13.
Theorem 16. DNC does not admit preservation of c.b-immunity.
Proof (Proof sketch). Let $\mu_{\emptyset}$ be the modulus function of $\emptyset^{\prime}$, that is, such that $\mu_{\emptyset^{\prime}}(x)$ is the minimum stage $s$ at which $\emptyset_{s}^{\prime} \upharpoonright x=\emptyset^{\prime}\lceil x$.

Computably split $\omega$ into countably many columns $X_{0}, X_{1}, \ldots$ of infinite size. For example, set $X_{i}=\{\langle i, n\rangle: n \in \omega\}$ where $\langle\cdot, \cdot\rangle$ is a bijective function from $\omega^{2}$ to $\omega$. For each $i$, let $F_{i}$ be the set of the $\mu_{\emptyset^{\prime}}(i)$ first elements of $X_{i}$. The sequence $F_{0}, F_{1}, \ldots$ is $\emptyset^{\prime}$-computable. By a simple finite injury priority argument (see appendix), one can construct a c.e. set $W$ such that the $\Delta_{2}^{0}$ set $A=\bigcup_{i} F_{i} \backslash W$ is c.b-immune, and such that $\left|X_{i} \cap W\right| \leq i$. We claim that every DNC function computes an infinite subset of $A$.

Let $f$ be any DNC function. By a classical theorem about DNC functions (see Bienvenu et al. [1] for a proof), $f$ computes a function $g(\cdot, \cdot, \cdot)$ such that whenever $\left|W_{e}\right| \leq n$, then $g(e, n, i) \in X_{i} \backslash W_{e}$. For each $i$, let $e_{i}$ be the index of the c.e. set $W_{e_{i}}=W \cap X_{i}$, and let $n_{i}=g\left(e_{i}, i, i\right)$. Since $\left|X_{i} \cap W\right| \leq i,\left|W_{e_{i}}\right| \leq i$, hence $n_{i}=g\left(e_{i}, i, i\right) \in X_{i} \backslash W_{e_{i}}=X_{i} \backslash W$. We then have two cases.

- Case $1: n_{i} \in F_{i}$ for infinitely many $i$ 's. One can $f$-computably find infinitely many of them since $\mu_{\emptyset^{\prime}}$ is left-c.e. and the sequence of the $n$ 's is $f$-computable. Therefore, one can $f$-computably find an infinite subset of $\bigcup_{i} F_{i} \backslash W=A$.
- Case 2: $n_{i} \in F_{i}$ for only finitely many $i$ 's. Then the sequence of the $n_{i}$ 's dominates the modulus function $\mu_{\emptyset^{\prime}}$, and therefore computes the halting set. Since the set $A$ is $\Delta_{2}^{0}, f$ computes an infinite subset of $A$.

Corollary 17 (Hirschfeldt and Shore [10]). $\mathrm{RCA}_{0} \wedge C A C \nvdash D N C$.

## 4 ADS and Dependent Hyperimmunity

Lerman et al. [13] separated the ascending descending sequence principle from a stable version of CAC by using a very involved iterated forcing argument. According to our previous simplification of their general framework [16], we reformulate their proof in terms of preservation of dependent hyperimmunity, and extend it to pseudo Ramsey's theorem for pairs.

Definition 18 (Ascending descending sequence). Given a linear order ( $L,<_{L}$ ), an ascending (descending) sequence is a set $S$ such that for every $x<_{\mathbb{N}}$ $y \in S, x<_{L} y\left(x>_{L} y\right)$. ADS is the statement "Every infinite linear order admits an infinite ascending or descending sequence".

Pseudo Ramsey's theorem for pairs was introduced by Friedman [7] and later studied by Friedman and Pelupessy [8], and Murakami et al. in [15] who proved that it is between the chain antichain principle and the ascending descending sequence principle over $\mathrm{RCA}_{0}$. Steila $[20]$ and the author [17] independently proved that it is actually equivalent to ADS.

Definition 19 (Pseudo Ramsey's theorem). A set $H=\left\{x_{0}<x_{1}<\ldots\right\}$ is pseudo-homogeneous for a coloring $f:[\mathbb{N}]^{n} \rightarrow k$ if $f\left(x_{i}, \ldots, x_{i+n-1}\right)=$ $f\left(x_{j}, \ldots, x_{j+n-1}\right)$ for every $i, j \in \mathbb{N}$. $\mathrm{psRT}_{k}^{n}$ is the statement "Every coloring $f:[\mathbb{N}]^{n} \rightarrow k$ has an infinite pseudo-homogeneous set".

Definition 20 (Dependent hyperimmunity). A formula $\varphi(U, V)$ is essential if for every $x \in \omega$, there is a finite set $R>x$ such that for every $y \in \omega$, there is a finite set $S>y$ such that $\varphi(R, S)$ holds. A pair of sets $A_{0}, A_{1} \subseteq \omega$ is dependently $X$-hyperimmune if for every essential $\Sigma_{1}^{0, X}$ formula $\varphi(U, V)$, $\varphi(R, S)$ holds for some $R \subseteq \bar{A}_{0}$ and $S \subseteq \bar{A}_{1}$.

In particular, if the pair $A_{0}, A_{1}$ is dependently hyperimmune, then $A_{0}$ and $A_{1}$ are both hyperimmune.

Definition 21 (Preservation of dependent hyperimmunity). $A \Pi_{2}^{1}$ statement P admits preservation of dependent hyperimmunity if for every set $Z$, every pair of dependently $Z$-hyperimmune sets $A_{0}, A_{1} \subseteq \omega$ and every Pinstance $X \leq_{T} Z$, there is a solution $Y$ to $X$ such that $A_{0}, A_{1}$ are dependently $Y \oplus Z$-hyperimmune.

A partial order $\left(P, \leq_{P}\right)$ is stable if either $(\forall i \in P)(\exists s)[(\forall j>s)(j \in P \rightarrow$ $\left.\left.i \leq_{P} j\right) \vee(\forall j>s)\left(\left.j \in P \rightarrow i\right|_{P} j\right)\right]$ or $(\forall i \in P)(\exists s)\left[(\forall j>s)\left(j \in P \rightarrow i \geq_{P}\right.\right.$ $\left.j) \vee(\forall j>s)\left(\left.j \in P \rightarrow i\right|_{P} j\right)\right]$. SCAC is the restriction of CAC to stable partial orders. A simple finite injury priority argument shows that SCAC does not admit preservation of dependent hyperimmunity.

Theorem 22. There exists a computable, stable semi-transitive coloring $f$ : $[\omega]^{2} \rightarrow 2$ such that the pair $\bar{A}_{0}, \bar{A}_{1}$ is dep. hyperimmune, where $A_{i}=\{x:$ $\left.\lim _{s} f(x, s)=i\right\}$.

Corollary 23. SCAC does not admit preservation of dependent hyperimmunity.
Proof. Let $f:[\omega]^{2} \rightarrow 2$ be the coloring of Theorem 22. By construction, the pair $\bar{A}_{0}, \bar{A}_{1}$ is dependently hyperimmune, where $A_{i}=\left\{x: \lim _{s} f(x, s)=i\right\}$. Let $H$ be an infinite $f$-homogeneous set. In particular, $H \subseteq A_{0}$ or $H \subseteq A_{1}$. We claim that the pair $\bar{A}_{0}, \bar{A}_{1}$ is not dependently $H$-hyperimmune. The $\Sigma_{1}^{0, H}$ formula $\varphi(U, V)$ defined by $U \neq \emptyset \wedge V \neq \emptyset \wedge U \cup V \subseteq H$ is essential since $H$ is infinite. However, if there is some $R \subseteq A_{1}$ and $S \subseteq A_{0}$ such that $\varphi(R, S)$ holds, then $H \cap A_{0} \neq \emptyset$ and $H \cap A_{1} \neq \emptyset$, contradicting the choice of $H$. Therefore $\bar{A}_{0}, \bar{A}_{1}$ is not dependently $H$-hyperimmune. Hirschfeldt and Shore [10] proved that SCAC is equivalent to stable semi-transitive Ramsey's theorem for pairs over RCA ${ }_{0}$. Therefore SCAC does not admit preservation of dependent hyperimmunity.

We will now prove the positive preservation result.

Theorem 24. For every $k \geq 2, \mathrm{psRT}_{\mathrm{k}}^{2}$ admits preservation of dep. hyperimmunity.

Proof. The proof is done by induction over $k \geq 2$. Fix a pair of sets $A_{0}, A_{1} \subseteq \omega$ dependently $Z$-hyperimmune for some set $Z$. Let $f:[\omega]^{2} \rightarrow k$ be a $Z$-computable coloring and suppose that there is no infinite set $H$ over which $f$ avoids at least one color, and such that the pair $A_{0}, A_{1}$ is dependently $H \oplus Z$-hyperimmune, as otherwise, we are done by induction hypothesis. We will build $k$ infinite sets $G_{0}, \ldots, G_{k-1}$ such that $G_{i}$ is pseudo-homogeneous for $f$ with color $i$ for each $i<k$ and such that $A_{0}, A_{1}$ is dependently $G_{i} \oplus Z$-hyperimmune for some $i<k$. The sets $G_{0}, \ldots, G_{k-1}$ are built by a variant of Mathias forcing $\left(F_{0}, \ldots, F_{k-1}, X\right)$ such that
(i) $F_{i} \cup\{x\}$ is pseudo-homogeneous for $f$ with color $i$ for each $x \in X$
(ii) $X$ is an infinite set such that $A_{0}, A_{1}$ is dependently $X \oplus Z$-hyperimmune

A condition $d=\left(H_{0}, \ldots, H_{k-1}, Y\right)$ extends $c=\left(F_{0}, \ldots, F_{k-1}, X\right)($ written $d \leq$ c) if $\left(H_{i}, Y\right)$ Mathias extends $\left(F_{i}, X\right)$ for each $i<k$. A tuple of sets $G_{0}, \ldots, G_{k-1}$ satisfies $c$ if for every $n \in \omega$, there is an extension $d=\left(H_{0}, \ldots, H_{k-1}, Y\right)$ of $c$ such that $G_{i} \upharpoonright n \subseteq H_{i}$ for each $i<k$. Informally, $G_{0}, \ldots, G_{k-1}$ satisfy $c$ if the sets are generated by a decreasing sequence of conditions extending $c$. In particular, $G_{i}$ is pseudo-homogeneous for $f$ with color $i$ and satisfies the Mathias condition $\left(F_{i}, X\right)$. The first lemma shows that every sufficiently generic filter yields a $k$-tuple of infinite sets.

Lemma 25. For every condition $c=\left(F_{0}, \ldots, F_{k-1}, X\right)$ and every $i<k$, there is an extension $d=\left(H_{0}, \ldots, H_{k-1}, Y\right)$ of $c$ such that $\left|H_{i}\right|>\left|F_{i}\right|$.

Fix an enumeration $\varphi_{0}(G, U, V), \varphi_{1}(G, U, V), \ldots$ of all $\Sigma_{1}^{0, Z}$ formulas. We want to satisfy the following requirements for each $e_{0}, \ldots, e_{k-1} \in \omega$ :

$$
\mathcal{R}_{\vec{e}}: \quad \mathcal{R}_{e_{0}}^{G_{0}} \quad \vee \quad \ldots \quad \vee \quad \mathcal{R}_{e_{k-1}}^{G_{k-1}}
$$

where $\mathcal{R}_{e}^{G}$ is the requirement " $\varphi_{e}(G, U, V)$ essential $\rightarrow \varphi_{e}(G, R, S)$ for some $R \subseteq$ $\bar{A}_{0}$ and $S \subseteq \bar{A}_{1}$ ". We say that a condition $c$ forces $\mathcal{R}_{\vec{e}}$ if $\mathcal{R}_{\vec{e}}$ holds for every $k$-tuple of sets satisfying $c$. Note that the notion of satisfaction has a precise meaning given above.

Lemma 26. For every condition $c$ and every $k$-tuple of indices $e_{0}, \ldots, e_{k-1} \in \omega$, there is an extension $d$ of $c$ forcing $\mathcal{R}_{\vec{e}}$.

Proof. Fix a condition $c=\left(F_{0}, \ldots, F_{k-1}, X\right)$. Let $\psi(U, V)$ be the $\Sigma_{1}^{0, X \oplus Z}$ formula which holds if there is a $k$-tuple of sets $E_{0}, \ldots, E_{k-1} \subseteq X$ and a $z \in X$ such that for each $i<k$,
(i) $z>\max \left(E_{i}\right)$
(ii) $F_{i} \cup E_{i} \cup\{z\}$ is pseudo-homogeneous for color $i$.
(iii) $\varphi_{e_{i}}\left(F_{i} \cup E_{i}, U_{i}, V_{i}\right)$ holds for some $U_{i} \subseteq U$ and $V_{i} \subseteq V$

Suppose that $c$ does not force $\mathcal{R}_{\vec{e}}$, otherwise we are done.
We claim that $\psi$ is essential. Since $c$ does not force $\mathcal{R}_{\vec{e}}$, there is a $k$-tuple of infinite sets $G_{0}, \ldots, G_{k-1}$ satisfying $c$ and such that $\varphi_{e_{i}}\left(G_{i}, U, V\right)$ is essential for each $i<k$. Fix some $x \in \omega$. By definition of being essential, there are some finite sets $R_{0}, \ldots, R_{k-1}>x$ such that for every $y \in \omega$, there are finite sets $S_{0}, \ldots, S_{k-1}>y$ such that $\varphi_{e_{i}}\left(G_{i}, R_{i}, S_{i}\right)$ holds for each $i<k$. Let $R=\bigcup R_{i}$ and fix some $y \in \omega$. There are finite sets $S_{0}, \ldots, S_{k-1}>y$ such that $\varphi_{e_{i}}\left(G_{i}, R_{i}, S_{i}\right)$ holds for each $i<k$. Let $S=\bigcup S_{i}$. By continuity, there are finite sets $E_{0}, \ldots, E_{k-1}$ such that $G_{i} \upharpoonright \max \left(E_{i}\right)=F_{i} \cup E_{i}$ and $\varphi_{e_{i}}\left(F_{i} \cup E_{i}, R_{i}, S_{i}\right)$ holds for each $i<k$. By our precise definition of satisfaction, we can even assume without loss of generality that $\left(F_{0} \cup E_{0}, \ldots, F_{k-1} \cup E_{k-1}, Y\right)$ is a valid extension of $c$ for some infinite set $Y \subseteq X$. Let $z \in Y$. In particular, by the definition of being a condition extending $c, z \in X, z>\max \left(E_{0}, \ldots, E_{k-1}\right)$ and $F_{i} \cup E_{i} \cup\{z\}$ is pseudo-homogeneous for color $i$ for each $i<k$. Therefore $\psi(R, S)$ holds, as witnessed by $E_{0}, \ldots, E_{k-1}$ and $z$. Thus $\psi(R, S)$ is essential.

Since $A_{0}, A_{1}$ is dependently $X \oplus Z$-hyperimmune, then $\psi(R, S)$ holds for some $R \subseteq \bar{A}_{0}$ and some $S \subseteq \bar{A}_{1}$. Let $E_{0}, \ldots, E_{k-1} \subseteq X$ be the $k$-tuple of sets and $z \in X$ be the integer witnessing $\psi(R, S)$. Let $i<k$ be such that the set $Y=\left\{w \in X \backslash\left[0, \max \left(E_{i}\right)\right]: f(z, w)=i\right\}$ is infinite. The condition $d=$ $\left(F_{0}, \ldots, F_{i-1}, F_{i} \cup E_{i} \cup\{z\}, F_{i+1}, \ldots, F_{k-1}, Y\right)$ is a valid extension of $c$ forcing $\mathcal{R}_{\vec{e}}$.

Theorem 27. Fix some set $Z$ and a pair of sets $A_{0}, A_{1}$ dependently $Z$ hyperimmune. If $Y$ is sufficiently random relative to $Z$, then the pair $A_{0}, A_{1}$ is dependently $Y \oplus Z$-hyperimmune.

Corollary 28. WWKL admits preservation of dependent hyperimmunity.
Proof. Let $Z$ be a set and $A_{0}, A_{1}$ be a pair of dependently $Z$-hyperimmune sets. Fix a $Z$-computable tree of positive measure $T \subseteq 2^{<\omega}$. By Theorem 27, the pair $A_{0}, A_{1}$ is dependently $Y \oplus Z$-hyperimmune for some Martin-Löf random $Y$ relative to $Z$. By Kučera [12], $Y$ is, up to finite prefix, a path through $T$.

Corollary 29. For every $k \geq 2, \mathrm{RCA}_{0} \wedge \mathrm{psR}_{\mathrm{k}}^{2} \wedge \mathrm{WWKL} \nvdash \mathrm{SCAC}$.
Whenever requiring the sets $A_{0}$ and $A_{1}$ to be co-c.e., we recover the standard notion of hyperimmunity. Therefore, the restriction of the preservation of dependent hyperimmunity to co-c.e. sets is not a good computability-theoretic property to distinguish consequences of Ramsey's theorem for pairs.

Lemma 30. Fix two sets $A_{0}, A_{1}$ such that $A_{0}$ is $X$-co-c.e. The pair $A_{0}, A_{1}$ is dependently $X$-hyperimmune iff $A_{0}$ and $A_{1}$ are $X$-hyperimmune.

Corollary 31. $\mathrm{RT}_{2}^{2}$ admits preservation of dependent hyperimmunity for co-c.e. sets.

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# A Direct Constructive Proof of a Stone-Weierstrass Theorem for Metric Spaces 

Iosif Petrakis ${ }^{(\boxtimes)}$<br>University of Munich, Munich, Germany<br>petrakis@math.lmu.de


#### Abstract

We present a constructive proof of a Stone-Weierstrass theorem for totally bounded metric spaces (SWtbms) which implies Bishop's Stone-Weierstrass theorem for compact metric spaces (BSWcms) found in [3]. Our proof has a clear computational content, in contrast to Bishop's highly technical proof of BSWcms and his hard to motivate concept of a (Bishop-)separating set of uniformly continuous functions. All corollaries of BSWcms in [3] are proved directly by SWtbms. We work within Bishop's informal system of constructive mathematics BISH.


## 1 Introduction

According to the classical Stone-Weierstrass theorem (SWchts), if $X$ is a compact Hausdorff topological space and $\Phi$ is a separating subalgebra of the continuous real-valued functions $C(X)$ on $X$ that contains a non-zero constant function, then the uniform closure of $\Phi$ is $C(X)$ (see [10], p. 282). Recall that $\Phi$ is separating, if $\forall_{x, y \in X}\left(x \neq y \rightarrow \exists_{f \in \Phi}(f(x) \neq f(y))\right)$.

There are some constructive versions of this theorem depending on the notion of space under constructive study. In [1] Banaschewski and Mulvey considered a compact, completely regular locale instead of a compact Hausdorff topological space. In [7] Coquand gave a simple, constructive localic proof of it, replacing the ring structure of $C(X)$ by its lattice structure, while in [8] he studied the usual formulation of the Stone-Weierstrass theorem in this point-free topological framework.

For reasons which we discuss in [15], Bishop did not pursue a constructive reconstruction of abstract topology. Although he introduced two constructive alternatives to the notion of topological space, the notion of neighborhood space, see [11, 13], and the notion of function space, or Bishop space, see [4, 12, 15-17], he never elaborated them, restricting his studies to metric spaces. In [2,3] Bishop formulated a theorem of Stone-Weierstrass type for compact metric spaces (i.e., complete and totally bounded metric spaces) using the notion of a Bishop-separating set of uniformly continuous functions ${ }^{1}$. Since Bishop's results, as well as ours, hold

[^54]for totally bounded metric spaces, we formulate all related concepts and results for them without restricting to compact metric spaces. Recall that a metric space $(X, d)$ is totally bounded, if for every $\epsilon>0$ there exists a finite $\epsilon$-approximation of $X$, and a set $A$ is finite if there exists a one-one mapping of $\{1, \ldots, n\}$ onto $A$, for some $n>0$ (see [6], p. 29). Hence, a totally bounded metric space is always inhabited.

Throughout this paper $(X, d)$ is a totally bounded metric space, $C_{u}(X)$ denotes the uniformly continuous real-valued functions on $X$, and $\Phi \subseteq C_{u}(X)$.

Definition 1. $\Phi$ is called Bishop-separating, if there is $\delta: \mathbb{R}^{+} \rightarrow \mathbb{R}^{+}$such that: $\left(\mathrm{Bsep}_{1}\right)$ For all $\epsilon>0$ and $x_{0}, y_{0} \in X$, if $d\left(x_{0}, y_{0}\right) \geq \epsilon$, there exists $g_{\epsilon, x_{0}, y_{0}} \in \Phi$ such that

$$
\begin{aligned}
& \forall_{z \in X}\left(d_{x_{0}}(z) \leq \delta(\epsilon) \rightarrow\left|g_{\epsilon, x_{0}, y_{0}}(z)\right| \leq \epsilon\right) \text { and } \\
& \forall_{z \in X}\left(d_{y_{0}}(z) \leq \delta(\epsilon) \rightarrow\left|g_{\epsilon, x_{0}, y_{0}}(z)-1\right| \leq \epsilon\right)
\end{aligned}
$$

$\left(\mathrm{Bsep}_{2}\right)$ For all $\epsilon>0$ and $x_{0} \in X$ there exists $g_{\epsilon, x_{0}} \in \Phi$ such that

$$
\forall_{z \in X}\left(d_{x_{0}}(z) \leq \delta(\epsilon) \rightarrow\left|g_{\epsilon, x_{0}}(z)-1\right| \leq \epsilon\right)
$$

Note that in Definition $1 g_{\epsilon, x_{0}, y_{0}}$ and $g_{\epsilon, x_{0}}$ are just notations that do not involve the use of some choice principle. Recall also that for every $x_{0} \in X$ the map $d_{x_{0}}: X \rightarrow \mathbb{R}$, defined by $x \mapsto d\left(x_{0}, x\right)$, is in $C_{u}(X)$ with modulus of continuity $\omega_{d_{x_{0}}}=\operatorname{id}_{\mathbb{R}^{+}}$. If $a \in \mathbb{R}$, we denote by $\bar{a}$ the constant map on $X$ with value $a$, and their set by Const $(X)$. We define

$$
\begin{aligned}
U_{0}(X) & :=\left\{d_{x_{0}} \mid x_{0} \in X\right\} \\
U_{0}^{*}(X) & :=U_{0}(X) \cup\{\overline{1}\} .
\end{aligned}
$$

We call $\Phi$ positively separating, if $\forall_{x, y \in X}\left(x \bowtie_{d} y \rightarrow \exists_{g \in \Phi}\left(g(x) \bowtie_{\mathbb{R}} g(y)\right)\right)$, where $x \bowtie_{d} y \leftrightarrow d(x, y)>0$, for every $x, y \in X$, and $a \bowtie_{\mathbb{R}} b \leftrightarrow|a-b|>0 \leftrightarrow a<$ $b \vee b<a$, for every $a, b \in \mathbb{R}$, are the canonical point-point apartness relations on $X$ and $\mathbb{R}$, respectively. The notion of a positively separating set $\Phi$ is the positive version of the classical notion of a separating subset of $C(X)$ for metric spaces. Clearly, $U_{0}(X)$ is positively separating.

Remark 1. If $\Phi$ is Bishop-separating, then $\Phi$ is positively separating.
Proof. By the Archimedean property of $\mathbb{R}$ (see [5], p. 57), if $x_{0}, y_{0} \in X$ such that $d\left(x_{0}, y_{0}\right)>0$, there is some natural number $N>2$ such that $d\left(x_{0}, y_{0}\right)>\frac{1}{N}$. By Bsep $_{1}$ we have that $\left|g_{\frac{1}{N}, x_{0}, y_{0}}\left(x_{0}\right)\right| \leq \frac{1}{N}$ and $\left|g_{\frac{1}{N}, x_{0}, y_{0}}\left(y_{0}\right)-1\right| \leq \frac{1}{N}$, for some $g_{\frac{1}{N}, x_{0}, y_{0}} \in \Phi$, therefore $g_{\frac{1}{N}, x_{0}, y_{0}}\left(x_{0}\right) \bowtie_{\mathbb{R}} g_{\frac{1}{N}, x_{0}, y_{0}}\left(y_{0}\right)$.

In [3], p. 106, Bishop formulated a theorem of Stone-Weierstrass type for compact metric spaces using the notion of a Bishop-separating set as the property that corresponds to the classical notion of a separating set in the formulation of SWchts. Bishop's proof of this theorem is non-trivial and does not involve the completeness property of $X$. Following Bishop, we denote by $\mathcal{A}(\Phi)$ the least
subset of $C_{u}(X)$ that includes $\Phi$ and it is closed with respect to addition, multiplication, and multiplication by reals. Bishop didn't define $\mathcal{A}(\Phi)$ inductively but explicitly, as the set of compositions of strict real polynomials in several variables with vectors of elements of $\Phi$ (see [3], p. 105).

Theorem 1 (Bishop's Stone-Weierstrass theorem for totally bounded metric spaces (BSWtbms)). If $\Phi$ is Bishop-separating, then $\mathcal{A}(\Phi)$ is dense in $C_{u}(X)$.

The condition of $\Phi$ being Bishop-separating implies that the constant map $\overline{1}$ is in the closure of $\mathcal{A}(\Phi)$ (see [3], p. 106). Bishop's formulation of BSWtbms represents a non-trivial technical achievement, namely to find a formulation of a theorem of Stone-Weierstrass type in the constructive theory of metric spaces that resembles the formulation of the classical SWchts. As Coquand and Spitters mention in [9], pp. 339-340, constructive proofs using a concrete presentation of topological notions (e.g., the Gelfand spectrum as a lattice) are "more direct than proofs via an encoding of topology in metric spaces, as is common in Bishop's constructive mathematics".

In the next two sections we present a Stone-Weierstrass theorem for metric spaces which avoids the concept of a Bishop-separating set, it has an informative and direct proof, it implies BSWtbms, and it proves directly all corollaries of BSWtbms.

## 2 A Stone-Weierstrass Theorem for Totally Bounded Metric Spaces

Definition 2. If $f, g \in C_{u}(X)$ and $\epsilon>0$, then $f \wedge g:=\min \{f, g\}, f \vee g:=$ $\max \{f, g\}$, and the uniform closure $\mathcal{U}(\Phi)$ of $\Phi$ is defined by

$$
\begin{gathered}
U(g, f, \epsilon): \leftrightarrow \forall_{x \in X}(|g(x)-f(x)| \leq \epsilon), \\
U(\Phi, f): \leftrightarrow \forall_{\epsilon>0} \exists_{g \in \Phi}(U(g, f, \epsilon)), \\
\mathcal{U}(\Phi):=\left\{f \in C_{u}(X) \mid U(\Phi, f)\right\} .
\end{gathered}
$$

The following remark is immediate to show.
Remark 2. If $\Phi$ is closed under addition, multiplication by reals and multiplication, then $\mathcal{U}(\Phi)$ is closed under addition, multiplication by reals and multiplication. Moreover, if $\Phi$ is closed under |.|, then $\mathcal{U}(\Phi)$ is closed under |.|.

The next two lemmas are proved in [3], pp. 105-106 (Lemmas 5.11 and 5.12).
Lemma 1. If $\operatorname{Const}(X) \subseteq \Phi$, and $\Phi$ is closed under addition and multiplication (or if $\Phi$ is closed under addition, multiplication by reals, and multiplication), then $\mathcal{U}(\Phi)$ is closed under $||,. \vee$ and $\wedge$.

Lemma 2. If $\Phi$ is closed under addition, multiplication by reals, and multiplication, and $f \in \mathcal{U}(\Phi)$ such that $\forall_{x \in X}(|f(x)| \geq c)$, for some $c>0$, then $\frac{1}{f} \in \mathcal{U}(\Phi)$.

Corollary 1. If $x_{0}, y_{0} \in X$ such that $d\left(x_{0}, y_{0}\right)>0$, then $\overline{1} \in \mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right)$.
Proof. If $x \in X$, then $0<d\left(x_{0}, y_{0}\right) \leq d\left(x_{0}, x\right)+d\left(x, y_{0}\right)=d_{x_{0}}(x)+d_{y_{0}}(x)$ i.e., $d\left(x_{0}, y_{0}\right) \leq d_{x_{0}}+d_{y_{0}} \in \mathcal{A}\left(U_{0}(X)\right)$. By Lemma 2 we get that $\frac{1}{d_{x_{0}}+d_{y_{0}}} \in$ $\mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right)$, therefore $\overline{1} \in \mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right)$.

The existence of $x_{0}, y_{0} \in X$ such that $d\left(x_{0}, y_{0}\right)>0$ is equivalent to the positivity of the diameter of $(X, d)$ (see the footnote in the proof of Lemma 3).

Definition 3. If $\mathbb{F}(X)$ denotes the set of real-valued functions on $X$, the set of Lipschitz functions $\operatorname{Lip}(X)$ on $(X, d)$ is defined by

$$
\begin{gathered}
\operatorname{Lip}(X, k):=\left\{f \in \mathbb{F}(X) \mid \forall_{x, y \in X}(|f(x)-f(y)| \leq k d(x, y))\right\} \\
\operatorname{Lip}(X):=\bigcup_{k \geq 0} \operatorname{Lip}(X, k)
\end{gathered}
$$

Remark 3. The set $\operatorname{Lip}(X) \subseteq C_{u}(X)$ includes $U_{0}(X), \operatorname{Const}(X)$ and it is closed under addition, multiplication by reals, and multiplication.

Proof. If $x_{0} \in X$, then $\left|d\left(x_{0}, x\right)-d\left(x_{0}, y\right)\right| \leq d(x, y)$, for every $x, y \in X$, therefore $U_{0}(X) \subseteq \operatorname{Lip}(X, 1)$. Clearly, $\operatorname{Const}(X) \subseteq \operatorname{Lip}(X, k)$, for every $k \geq 0$. Recall that $f \cdot g=\frac{1}{2}\left((f+g)^{2}-f^{2}-g^{2}\right)$, and if $M_{f}>0$ is a bound of $f$, it is immediate to see that

$$
\begin{gathered}
f \in \operatorname{Lip}\left(X, k_{1}\right) \rightarrow g \in \operatorname{Lip}\left(X, k_{2}\right) \rightarrow f+g \in \operatorname{Lip}\left(X, k_{1}+k_{2}\right), \\
f \in \operatorname{Lip}(X, k) \rightarrow \lambda \in \mathbb{R} \rightarrow \lambda f \in \operatorname{Lip}(X,|\lambda| k) \\
f \in \operatorname{Lip}(X, k) \rightarrow f^{2} \in \operatorname{Lip}\left(X, 2 M_{f} k\right)
\end{gathered}
$$

Lemma 3. If $\Phi=\mathcal{A}\left(U_{0}^{*}(X)\right)$, then $\operatorname{Lip}(X) \subseteq \mathcal{U}(\Phi)$.
Proof. It suffices to show that $\operatorname{Lip}(X, 1) \subseteq \mathcal{U}(\Phi)$, since if $f \in \operatorname{Lip}(X, k)$, for some $k>0$, then $\frac{1}{k} f \in \operatorname{Lip}(X, 1)$ and we have, for every $\epsilon>0$ and $\theta \in \Phi$,

$$
U\left(\theta, \frac{1}{k} f, \frac{\epsilon}{k}\right) \rightarrow U(k \theta, f, \epsilon)
$$

Suppose next that $f \in \operatorname{Lip}(X, 1)$ and $\epsilon>0$. We find $g \in \mathcal{U}(\Phi)$ such that $U(g, f, \epsilon)$, therefore $f \in \mathcal{U}(\mathcal{U}(\Phi))=\mathcal{U}(\Phi)$. More specifically, if $\left\{z_{1}, \ldots, z_{m}\right\}$ is an $\frac{\epsilon}{2}$-approximation of $X$, we find $g \in \mathcal{U}(\Phi)$ such that $g\left(z_{i}\right)=f\left(z_{i}\right)$, for every $i \in\{1, \ldots, m\}$, and $\left|g(x)-g\left(z_{i}\right)\right|=\left|g(x)-f\left(z_{i}\right)\right| \leq \frac{\epsilon}{2}$, for every $x \in X$ and $z_{i}$ such that $d\left(x, z_{i}\right) \leq \frac{\epsilon}{2}$. Consequently,

$$
\begin{aligned}
|g(x)-f(x)| & \leq\left|g(x)-g\left(z_{i}\right)\right|+\left|g\left(z_{i}\right)-f\left(z_{i}\right)\right|+\left|f\left(z_{i}\right)-f(x)\right| \\
& \leq \frac{\epsilon}{2}+0+d\left(z_{i}, x\right) \\
& \leq \frac{\epsilon}{2}+\frac{\epsilon}{2} \\
& =\epsilon .
\end{aligned}
$$

We define

$$
g:=\bigwedge_{k=1}^{m}\left(\overline{f\left(z_{k}\right)}+d_{z_{k}}\right)
$$

Since $\overline{f\left(z_{k}\right)}+d_{z_{k}} \in \Phi$ and since by Lemma $1 \mathcal{U}(\Phi)$ is closed under $\wedge$ we get $g \in \mathcal{U}(\Phi)$. Moreover,

$$
g\left(z_{i}\right)=\bigwedge_{k=1}^{m}\left(f\left(z_{k}\right)+d_{z_{k}}\left(z_{i}\right)\right) \leq f\left(z_{i}\right)+d_{z_{i}}\left(z_{i}\right)=f\left(z_{i}\right)
$$

For the converse inequality we suppose that $g\left(z_{i}\right)<f\left(z_{i}\right)$ and reach a contradiction (here we use the fact that $\neg(a<b) \rightarrow a \geq b$, for every $a, b \in \mathbb{R}$ (see [3], p. 26)). If $a, b, c \in \mathbb{R}$, then one shows ${ }^{2}$ that $a \wedge b<c \rightarrow a<c \vee b<c$. Hence

$$
\bigwedge_{k=1}^{m}\left(f\left(z_{k}\right)+d_{z_{k}}\left(z_{i}\right)\right)<f\left(z_{i}\right) \rightarrow \exists_{j \in\{1, \ldots, m\}}\left(f\left(z_{j}\right)+d\left(z_{j}, z_{i}\right)<f\left(z_{i}\right)\right)
$$

$$
\rightarrow d\left(z_{j}, z_{i}\right)<f\left(z_{i}\right)-f\left(z_{j}\right) \leq\left|f\left(z_{i}\right)-f\left(z_{j}\right)\right| \leq d\left(z_{j}, z_{i}\right)
$$

which is a contradiction. Using the equality $g\left(z_{i}\right)=f\left(z_{i}\right)$ we have that

$$
\begin{gathered}
g(x)=\bigwedge_{k=1}^{m}\left(f\left(z_{k}\right)+d_{z_{k}}(x)\right) \leq f\left(z_{i}\right)+d_{z_{i}}(x) \rightarrow \\
g(x)-g\left(z_{i}\right) \leq f\left(z_{i}\right)+d_{z_{i}}(x)-g\left(z_{i}\right)=f\left(z_{i}\right)+d_{z_{i}}(x)-f\left(z_{i}\right)=d\left(x, z_{i}\right) \leq \frac{\epsilon}{2} .
\end{gathered}
$$

If $k \in\{1, \ldots, m\}$, then $f\left(z_{i}\right)-f\left(z_{k}\right) \leq\left|f\left(z_{i}\right)-f\left(z_{k}\right)\right| \leq d\left(z_{i}, z_{k}\right) \leq d\left(z_{i}, x\right)+$ $d\left(x, z_{k}\right)$, therefore

$$
\begin{aligned}
& \forall_{k \in\{1, \ldots, m\}}\left(f\left(z_{i}\right)-d\left(z_{i}, x\right) \leq f\left(z_{k}\right)+d\left(z_{k}, x\right)\right) \rightarrow \\
& f\left(z_{i}\right)-d\left(z_{i}, x\right) \leq \bigwedge_{k=1}^{m}\left(f\left(z_{k}\right)+d\left(z_{k}, x\right)\right) \leftrightarrow \\
& f\left(z_{i}\right)-\bigwedge_{k=1}^{m}\left(f\left(z_{k}\right)+d\left(z_{k}, x\right)\right) \leq d\left(z_{i}, x\right) \rightarrow \\
& g\left(z_{i}\right)-g(x) \leq d\left(z_{i}, x\right) \rightarrow \\
& g\left(z_{i}\right)-g(x) \leq \frac{\epsilon}{2}
\end{aligned}
$$

From $g(x)-g\left(z_{i}\right) \leq \frac{\epsilon}{2}$ and $g\left(z_{i}\right)-g(x) \leq \frac{\epsilon}{2}$ we get $\left|g(x)-g\left(z_{i}\right)\right| \leq \frac{\epsilon}{2}$.

[^55]Lemma 4. If $f \in C_{u}(X)$ and $\epsilon>0$, there exist $\sigma>0$ and $g, g^{*} \in \operatorname{Lip}(X, \sigma)$ such that
(i) $\forall_{x \in X}\left(f(x)-\epsilon \leq g(x) \leq f(x) \leq g^{*}(x) \leq f(x)+\epsilon\right)$.
(ii) For every $e \in \operatorname{Lip}(X, \sigma)$, if $e \leq f$, then $e \leq g$.
(iii) For every $e^{*} \in \operatorname{Lip}(X, \sigma)$, if $f \leq e^{*}$, then $g^{*} \leq e^{*}$.

Proof. (i) Let $\omega_{f}$ be a modulus of continuity of $f$ and $M_{f}>0$ a bound of $f$. We define the functions $h_{x}: X \rightarrow \mathbb{R}$ and $g: X \rightarrow \mathbb{R}$ by

$$
\begin{gathered}
h_{x}:=f+\sigma d_{x}, \\
\sigma:=\frac{2 M_{f}}{\omega_{f}(\epsilon)}>0, \\
g(x):=\inf \left\{h_{x}(y) \mid y \in X\right\}=\inf \{f(y)+\sigma d(x, y) \mid y \in X\},
\end{gathered}
$$

for every $x \in X$. Note that $g(x)$ is well-defined, since $h_{x} \in C_{u}(X)$ and the infimum of $h_{x}$ exists (see [3], p. 38, 94). First we show that $g \in \operatorname{Lip}(X, \sigma)$. If $x_{1}, x_{2}, y \in X$ the inequality $d\left(x_{1}, y\right) \leq d\left(x_{2}, y\right)+d\left(x_{1}, x_{2}\right)$ implies that $f(y)+$ $\sigma d\left(x_{1}, y\right) \leq\left(f(y)+\sigma d\left(x_{2}, y\right)\right)+\sigma d\left(x_{1}, x_{2}\right)$, hence $g\left(x_{1}\right) \leq\left(f(y)+\sigma d\left(x_{2}, y\right)\right)+$ $\sigma d\left(x_{1}, x_{2}\right)$, therefore $g\left(x_{1}\right) \leq g\left(x_{2}\right)+\sigma d\left(x_{1}, x_{2}\right)$, or $g\left(x_{1}\right)-g\left(x_{2}\right) \leq \sigma d\left(x_{1}, x_{2}\right)$. Starting with the inequality $d\left(x_{2}, y\right) \leq d\left(x_{1}, y\right)+d\left(x_{1}, x_{2}\right)$ and working similarly we get that $g\left(x_{2}\right)-g\left(x_{1}\right) \leq \sigma d\left(x_{1}, x_{2}\right)$, therefore $\left|g\left(x_{1}\right)-g\left(x_{2}\right)\right| \leq \sigma d\left(x_{1}, x_{2}\right)$. Next we show that

$$
\forall_{x \in X}(f(x)-\epsilon \leq g(x) \leq f(x))
$$

Since $f(x)=f(x)+\sigma d(x, x)=h_{x}(x) \geq \inf \left\{h_{x}(y) \mid y \in X\right\}=g(x)$, for every $x \in X$, we have that $g \leq f$. Next we show that $\forall_{x \in X}(f(x)-\epsilon \leq g(x))$. For that we fix $x \in X$ and we show that $\neg(f(x)-\epsilon>g(x))$. Note that if $A \subseteq \mathbb{R}, b \in \mathbb{R}$, then $^{3} b>\inf A \rightarrow \exists_{a \in A}(a<b)$. Therefore,

$$
\begin{aligned}
& f(x)-\epsilon>g(x) \leftrightarrow \\
& f(x)-\epsilon>\inf \{f(y)+\sigma d(x, y) \mid y \in X\} \rightarrow \\
& \exists_{y \in X}(f(x)-\epsilon>f(y)+\sigma d(x, y)) \leftrightarrow \\
& \exists_{y \in X}(f(x)-f(y)>\epsilon+\sigma d(x, y)) .
\end{aligned}
$$

For this $y$ we show that $d(x, y) \leq \omega_{f}(\epsilon)$. If $d(x, y)>\omega_{f}(\epsilon)$, we have that

$$
2 M_{f} \geq f(x)+M_{f} \geq f(x)-f(y)>\epsilon+2 M_{f} \frac{d(x, y)}{\omega_{f}(\epsilon)}>\epsilon+2 M_{f}>2 M_{f}
$$

which is a contradiction. Hence, by the uniform continuity of $f$ we get that $|f(x)-f(y)| \leq \epsilon$, therefore the contradiction $\epsilon>\epsilon$ is reached, since

$$
\epsilon \geq|f(x)-f(y)| \geq f(x)-f(y)>\epsilon+\sigma d(x, y) \geq \epsilon
$$

[^56]Next we define the functions $h_{x}^{*}: X \rightarrow \mathbb{R}$ and $g^{*}: X \rightarrow \mathbb{R}$ by

$$
\begin{gathered}
h_{x}^{*}:=f-\sigma d_{x} \\
g^{*}(x):=\sup \left\{h_{x}^{*}(y) \mid y \in X\right\}=\sup \{f(y)-\sigma d(x, y) \mid y \in X\}
\end{gathered}
$$

for every $x \in X$, and $\sigma=\frac{2 M_{f}}{\omega_{f}(\epsilon)}$. Similarly ${ }^{4}$ to $g$ we get that $g^{*} \in \operatorname{Lip}(X, \sigma)$ and

$$
\forall_{x \in X}\left(f(x) \leq g^{*}(x) \leq f(x)+\epsilon\right)
$$

(ii) Let $e \in \operatorname{Lip}(X, \sigma)$ such that $e \leq f$. If we fix some $x \in X$, then for every $y \in X$ we have that $e(x)-e(y) \leq|e(x)-e(y)| \leq \sigma d(x, y)$, hence $e(x) \leq e(y)+\sigma d(x, y) \leq$ $f(y)+\sigma d(x, y)$, therefore $e(x) \leq g(x)$.
(iii) Let $e^{*} \in \operatorname{Lip}(X, \sigma)$ such that $f \leq e^{*}$. If we fix some $x \in X$, then for every $y \in X$ we have that $e^{*}(y)-e^{*}(x) \leq\left|e^{*}(y)-e^{*}(x)\right| \leq \sigma d(x, y)$, hence $f(y)-\sigma d(x, y) \leq e^{*}(y)-\sigma d(x, y) \leq e^{*}(x)$, therefore $g^{*}(x) \leq e^{*}(x)$.
Hence $g$ is the largest function in $\operatorname{Lip}(X, \sigma)$ which is smaller than $f$, and $g^{*}$ is the smallest function in $\operatorname{Lip}(X, \sigma)$ which is larger than $f$. So, if there is some $e^{\prime} \in$ $\operatorname{Lip}(X)$ such that $e^{\prime} \leq f$ and $g(x)<e^{\prime}(x)$, for some $x \in X$, then $e^{\prime} \in \operatorname{Lip}\left(X, \sigma^{\prime}\right)$, for some $\sigma^{\prime}>\sigma$. It is interesting that Lemma 4 is in complete analogy to the McShane-Kirszbraun theorem. To make this clear we include a constructive version of this theorem (for a classical presentation see [18], p. 6). Recall that $A \subseteq X$ is located, if the distance $d(x, A):=\inf \{d(x, y) \mid y \in Y\}$ exists for every $x \in X$, and that a located subset of a totally bounded metric space is totally bounded (see [3], p. 95).
Proposition 1 (McShane-Kirszbraun theorem for totally bounded metric spaces). If $\sigma>0, A \subseteq X$ is located, and $f: A \rightarrow \mathbb{R} \in \operatorname{Lip}(A, \sigma)$, then there exist $g, g^{*} \in \operatorname{Lip}(X, \sigma)$ such that $g_{\mid A}=g_{\mid A}^{*}=f$ and for every $e \in \operatorname{Lip}(X, \sigma)$ such that $e_{\mid A}=f$ we have that $g^{*} \leq e \leq g$.
Proof. The functions $g, g^{*}$ defined by $g(x):=\inf \{f(a)+\sigma d(x, a) \mid a \in A\}$, and $g^{*}(x):=\sup \{f(a)-\sigma d(x, a) \mid a \in A\}$, for every $x \in X$, are well-defined and satisfy the required properties.

Corollary 2. $\mathcal{U}(\operatorname{Lip}(X))=C_{u}(X)$.
Proof. If $\epsilon>0$, then the functions $g, g^{*} \in \operatorname{Lip}(X, \sigma)$ of Lemma 4 satisfy $U(g, f, \epsilon)$, $U\left(g^{*}, f, \epsilon\right)$, respectively.

Next follows our Stone-Weierstrass theorem for totally bounded metric spaces.
Theorem 2 (Stone-Weierstrass theorem for totally bounded metric spaces (SWtbms)). If $\Phi=\mathcal{A}\left(U_{0}^{*}(X)\right)$, then $C_{u}(X)=\mathcal{U}(\Phi)$.
Proof. First we show that $C_{u}(X) \subseteq \mathcal{U}(\Phi)$. If $f \in C_{u}(X)$ and $\epsilon>0$, then by Corollary 2 there exists $h \in \operatorname{Lip}(X)$ such that $U\left(h, f, \frac{\epsilon}{2}\right)$, while by Lemma 3 there exists $g \in \Phi$ such that $U\left(g, h, \frac{\epsilon}{2}\right)$. Consequently, $U(g, f, \epsilon)$. The converse inclusion follows from the immediate fact that all elements of $\mathcal{U}(\Phi)$ are in $C_{u}(X)$.

[^57]
## 3 Corollaries of SWtbms

## Proposition 2. SWtbms implies BSWtbms

Proof. The proof follows immediately by inspection of the proof of BSWtbms in [3], pp. 106-108. Bishop shows there that if $\Phi$ is Bishop-separating, then $\overline{1} \in \mathcal{U}(\mathcal{A}(\Phi))$, and by his Lemma 5.14 .1 one shows that $U_{0}(X) \subseteq \mathcal{U}(\mathcal{A}(\Phi))$ - this is a slight simplification of the final part of Bishop's proof that $C_{u}(X) \subseteq \mathcal{U}(\mathcal{A}(\Phi))$. Since $U_{0}^{*}(X) \subseteq \mathcal{U}(\mathcal{A}(\Phi))$, then by Remark $2 \mathcal{A}\left(U_{0}^{*}(X)\right) \subseteq \mathcal{U}(\mathcal{A}(\Phi))$, therefore $C_{u}(X)=\mathcal{U}\left(\mathcal{A}\left(U_{0}^{*}(X)\right)\right) \subseteq \mathcal{U}(\mathcal{U}(\mathcal{A}(\Phi)))=\mathcal{U}(\mathcal{A}(\Phi))$.

In the proof of Corollary 5.16 in [3], pp. 108-109, it is shown that if $(X, d)$ has positive diameter, then $\mathcal{A}\left(U_{0}(X)\right)$ is a Bishop-separating set, therefore by BSWtbms we get that $\mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right)=C_{u}(X)$. Hence $\mathbf{S W t b m s}$ is only "slightly" stronger than BSWtbms. If we use SWtbms, we get immediately the same result.

Corollary 3. If $(X, d)$ has positive diameter, then $\mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right)=C_{u}(X)$.
Proof. The hypothesis of positive diameter implies the hypothesis of Corollary 1, therefore $\overline{1} \in \mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right) \subseteq C_{u}(X)$. Hence $U_{0}^{*}(X) \subseteq \mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right)$, and by Remark 2 we get that $\mathcal{A}\left(U_{0}^{*}(X)\right) \subseteq \mathcal{A}\left(\mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right)\right)=\mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right)$. Thus $C_{u}(X)=\mathcal{U}\left(\mathcal{A}\left(U_{0}^{*}(X)\right)\right) \subseteq \mathcal{U}\left(\mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right)\right)=\mathcal{U}\left(\mathcal{A}\left(U_{0}(X)\right)\right)$.

Next we prove Corollary 5.15 in [3], p. 108 and its finite version using SWtbms. If $(X, d),(Y, \rho)$ are totally bounded, then $(X \times Y, \sigma)$ is totally bounded, where $\sigma\left(\left(x_{1}, y_{1}\right),\left(x_{2}, y_{2}\right)\right):=d\left(x_{1}, x_{2}\right)+\rho\left(y_{1}, y_{2}\right)$, for every $x_{1}, x_{2} \in X$ and $y_{1}, y_{2} \in Y$; if $A=\left\{x_{1}, \ldots, x_{n}\right\}$ is an $\frac{\epsilon}{2}$-approximation of $X$ and $B=\left\{y_{1}, \ldots, y_{m}\right\}$ is an $\frac{\epsilon}{2}$-approximation of $Y$, then $A \times B$ is an $\epsilon$-approximation of $X \times Y$. We denote by $\pi_{1}$ the projection of $X \times Y$ onto $X$ and by $\pi_{2}$ its projection onto $Y$.

Corollary 4. If $(X, d),(Y, \rho)$ are totally bounded metric spaces and

$$
\Phi:=\left\{\sum_{i=1}^{n}\left(f_{i} \circ \pi_{1}\right)\left(g_{i} \circ \pi_{2}\right) \mid f_{i} \in C_{u}(X), g_{i} \in C_{u}(Y), 1 \leq i \leq n, n \in \mathbb{N}\right\}
$$

then $\mathcal{U}(\Phi)=C_{u}(X \times Y)$.
Proof. Clearly, $\Phi \subseteq C_{u}(X \times Y), \Phi$ is an algebra (actually, $\Phi=\mathcal{A}\left(\left(C_{u}(X) \circ \pi_{1}\right) \cup\right.$ $\left.\left(C_{u}(Y) \circ \pi_{2}\right)\right)$, where $C_{u}(X) \circ \pi_{1}=\left\{f \circ \pi_{1} \mid f \in C_{u}(X)\right\}$ and $C_{u}(Y) \circ \pi_{2}=\left\{g \circ \pi_{2} \mid\right.$ $\left.\left.g \in C_{u}(Y)\right\}\right)$, and $\mathcal{U}(\Phi) \subseteq C_{u}(X \times Y)$. The constant $\overline{1}$ on $X \times Y$ is equal to ( $\overline{1} \circ$ $\left.\pi_{1}\right)\left(\overline{1} \circ \pi_{2}\right)$. If $x_{0}, x \in X$ and $y_{0}, y \in Y$, then $\sigma_{\left(x_{0}, y_{0}\right)}((x, y))=\sigma\left(\left(x_{0}, y_{0}\right),(x, y)\right)=$ $d\left(x_{0}, x\right)+\rho\left(y_{0}, y\right)=d_{x_{0}}(x)+\rho_{y_{0}}(y)=\left(d_{x_{0}} \circ \pi_{1}\right)((x, y))+\left(\rho_{y_{0}} \circ \pi_{2}\right)((x, y))$, therefore $\sigma_{\left(x_{0}, y_{0}\right)}=\left(d_{x_{0}} \circ \pi_{1}\right)+\left(\rho_{y_{0}} \circ \pi_{2}\right)=\left(d_{x_{0}} \circ \pi_{1}\right)\left(\overline{1} \circ \pi_{2}\right)+\left(\overline{1} \circ \pi_{1}\right)\left(\rho_{y_{0}} \circ \pi_{2}\right) \in \Phi$. Since $U_{0}^{*}(X \times Y) \subseteq \mathcal{U}(\Phi)$, by SWtbms we get that $C_{u}(X \times Y) \subseteq \mathcal{U}(\Phi)$.

If $\left(X_{n}, d_{n}\right)$ is totally bounded, where without loss of generality $d_{n} \leq \overline{1}$, for every $n \in \mathbb{N}$, then $\left(X, \sigma_{\infty}\right)$, where $X=\prod_{n=1}^{\infty} X_{n}$ and $\sigma_{\infty}\left(\left(x_{n}\right)_{n=1}^{\infty},\left(y_{n}\right)_{n=1}^{\infty}\right):=$ $\sum_{n=1}^{\infty} \frac{d_{n}\left(x_{n}, y_{n}\right)}{2^{n}}$, is totally bounded; if $A\left(X_{n}, \epsilon\right)$ is an $\epsilon$-approximation of $X_{n}$ and $x_{0, n}$ inhabits $X_{n}$, then $A(X, \epsilon)=\prod_{k=1}^{n_{0}} A\left(X_{k}, \frac{2^{k-1} \epsilon}{n_{0}}\right) \times \prod_{k=n_{0}+1}^{\infty}\left\{x_{0, k}\right\}$ is an $\epsilon$-approximation of $X$, where $n_{0} \in \mathbb{N}$ such that $\sum_{k=n_{0}+1}^{\infty} \frac{1}{2^{k}} \leq \frac{\epsilon}{2}$.

Corollary 5. If $\left(X, \sigma_{\infty}\right)$ is the product of a sequence $\left(X_{n}, d_{n}\right)_{n=1}^{\infty}$ of totally bounded metric spaces, then $\mathcal{U}(\Phi)=C_{u}(X)$, where

$$
\begin{gathered}
\Phi_{0}:=\left\{\prod_{i=1}^{n}\left(f_{i} \circ \pi_{i}\right) \mid f_{i} \in C_{u}\left(X_{i}\right), 1 \leq i \leq n, n \in \mathbb{N}\right\}, \\
\Phi:=\left\{\sum_{k=1}^{n} h_{k} \mid h_{k} \in \Phi_{0}, 1 \leq k \leq n, n \in \mathbb{N}\right\} .
\end{gathered}
$$

Proof. Without loss of generality let $d_{n} \leq \overline{1}$, for every $n \in \mathbb{N}$. The only difference with the proof of Corollary 4 is treated as follows. If $\left(x_{n}^{0}\right)_{n=1}^{\infty} \in X$ and $\epsilon>0$, let

$$
g:=\sum_{k=1}^{n_{0}} \frac{d_{k, x_{k}^{0}} \circ \pi_{k}}{2^{k}}=\sum_{k=1}^{n_{0}}\left(\frac{d_{k, x_{k}^{0}}}{2^{k}}\right) \circ \pi_{k} \in \Phi
$$

where $n_{0} \in \mathbb{N}$ such that $\sum_{k=n_{0}+1}^{\infty} \frac{1}{2^{k}} \leq \epsilon$. We get $U\left(g, \sigma_{\infty,\left(x_{n}^{0}\right)_{n=1}^{\infty}}, \epsilon\right)$, since

$$
\left|g\left(\left(y_{n}\right)_{n=1}^{\infty}\right)-\sigma_{\infty,\left(x_{n}^{0}\right)_{n=1}^{\infty}}\left(\left(y_{n}\right)_{n=1}^{\infty}\right)\right|=\left|\sum_{k=n_{0}+1}^{\infty} \frac{d_{k, x_{k}^{0}}\left(y_{k}\right)}{2^{k}}\right| \leq \sum_{k=n_{0}+1}^{\infty}\left|\frac{d_{k}\left(x_{k}^{0}, y_{k}\right)}{2^{k}}\right| \leq \epsilon .
$$

Recall that a totally bounded metric space is separable (see [3], p. 94). The separability of $C_{u}(X)$ follows by the next corollary.

Corollary 6. If $Q=\left\{q_{n} \mid n \in \mathbb{N}\right\}$ is dense in $(X, d), U_{0}(Q):=\left\{d_{q_{n}} \mid n \in \mathbb{N}\right\}$, and $\Phi_{0}^{*}=\mathcal{A}\left(U_{0}(Q) \cup\{\overline{1}\}\right)$, then $\mathcal{U}\left(\Phi_{0}^{*}\right)=C_{u}(X)$.

Proof. If $\left(x_{n}\right)_{n=1}^{\infty} \in X^{\mathbb{N}}$ converges pointwise to $x$, then $\left(d_{x_{n}}\right)_{n=1}^{\infty}$ converges uniformly to $d_{x}$ [i.e., if $\forall_{\epsilon>0} \exists_{n_{0}} \forall_{n \geq n_{0}}\left(d\left(x_{n}, x\right) \leq \epsilon\right)$, then $\left.\forall_{\epsilon>0} \exists_{n_{0}} \forall_{n \geq n_{0}} \forall_{y \in X}\left(\left|d\left(x_{n}, y\right)-d(x, y)\right| \leq \epsilon\right)\right]$. If $\epsilon>0$ and $n \geq n_{0}$, then $d\left(x_{n}, y\right) \leq d\left(x_{n}, x\right)+d(x, y) \rightarrow d\left(x_{n}, y\right)-d(x, y) \leq d\left(x_{n}, x\right) \leq \epsilon$, and similarly $d(x, y)-d\left(x_{n}, y\right) \leq d\left(x_{n}, x\right) \leq \epsilon$. By $\mathbf{S W t b m s}$ it suffices to show that $U_{0}(X) \subseteq \mathcal{U}\left(\mathcal{A}\left(U_{0}(Q)\right)\right)$. If $d_{x} \in U_{0}(X)$, for some $x \in X$, and $\left(q_{k_{n}}\right)_{n=1}^{\infty}$ is a subsequence of $Q$ that converges pointwise to $x$, then $\left(d_{q_{k_{n}}}\right)_{n=1}^{\infty}$ converges uniformly to $d_{x}$, therefore $d_{x} \in \mathcal{U}\left(\mathcal{A}\left(U_{0}(Q)\right)\right)$.

## 4 Concluding Comments

We presented a direct constructive proof of $\mathbf{S W}$ tbms with a clear computational content. Its translation to Type Theory and its implementation to a proof
assistant like Coq are expected to be straightforward. Although SWtbms does not look like a theorem of Stone-Weierstrass type, as BSWtbms does, it has certain advantages over it. Its proof is "natural", in comparison to Bishop's technical proof and his difficult to motivate concept of a Bishop-separating set. As we explained, SWtbms implies BSWtbms, and all applications of BSWtbms in [3] are proved directly by SWtbms. We know of no application of BSWtbms which cannot be derived by SWtbms (in [3] we found only one application of Corollary 4 and one of the Weierstrass approximation theorem ${ }^{5}$ ).

An interesting question related to Corollary 2 is if for $(X, d)$ totally bounded and $(Y, \rho)$ complete metric space, the set of Lipschitz functions $\operatorname{Lip}(X, Y)$ between them is a dense subset of the uniformly continuous functions $C_{u}(X, Y)$ between them. A similar classical result, see [14], requires a Lipschitz extension property, which indicates that the correlation of Lemma 4 to the McShaneKirszbraun theorem may not be accidental.

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${ }^{5}$ The first, in p. 414, is the uniform approximation of a test function $f(x, y)$ on $G \times G$, where $G$ is a locally compact group, by finite sums of the form $\sum_{i} f_{i}(x) g_{i}(y)$, and the second, in p. 375, is a density theorem in the theory of Hilbert spaces.
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[^0]:    ${ }^{1}$ Another strand of work focusses on theorem proving, e.g., [28], but such approaches typically require user interaction to guide the search for a proof.

[^1]:    ${ }^{2}$ There is work on model-checking infinite state transition systems, see, for example, [11], but in this paper we concentrate on the finite case.

[^2]:    ${ }^{3}$ CL is a fragment of ATL with only the next time $\langle\langle A\rangle\rangle \bigcirc$ modality.

[^3]:    ${ }^{4}$ Here, $(0,1)$ is the allocation of 0 units of $r_{1}$ and 1 unit of $r_{2}$ to agent 1 . We only show resource bound for the proponent agents, $\{1\}$ in this case. Versions of resource logic where opponents are also resource-bounded all have an undecidable model-checking problem, see [14].

[^4]:    V. Brattka-Supported by the National Research Foundation of South Africa. This article uses some historical insights that were established in [2].

[^5]:    ${ }^{1}$ See [19] for a definition of computably compact.

[^6]:    The majority of this work was done while Le Roux was at the Department of Mathematics, Technische Universität Darmstadt, Germany and Pauly was at the Computer Laboratory, University of Cambridge, United Kingdom.

    This project has been supported by the National Research Foundation of South Africa (NRF) and by the German Research Foundation (DFG) through the GermanSouth African project (DFG, 445 SUA-1 13/20/0).

[^7]:    ${ }^{1}$ Which was put to the authors by Kohlenbach.

[^8]:    ${ }^{1}$ Furthermore, in the case $k>1$, one can replace the $k$ evaluation points for the secret by a primitive element of the extension field $\mathbb{F}_{q^{k}}$, whereby one only needs $n \leq q-1$, and the privacy and reconstruction thresholds are preserved. The multiplicativity properties hold now with respect to the product in $\mathbb{F}_{q^{k}}$ (for the secrets) instead of the coordinate-wise product in $\mathbb{F}_{q}^{k}$.

[^9]:    ${ }^{2}$ Here we suppose each pair of players is connected by a secure point-to-point channel, but we do not assume the existence of a broadcast channel.

[^10]:    ${ }^{3}$ See [27] for full definitions of $\mathfrak{T}$ and $\mathfrak{R}$ and references for the simulation results just summarized.

[^11]:    ${ }^{4}$ Considerable inductive support for this claim is supplied by Knuth's [15] detailed implementations of a wide range of algorithms using a variant of $\mathfrak{R}$ which satisfies the Invariance Thesis when its multiplication and division operations are removed.
    ${ }^{5}$ I.e. for all $N_{0} \in \mathfrak{N}_{0}$ there exists $N_{1} \in \mathfrak{N}_{1}$ computing the same function such that $\operatorname{time}_{N_{1}}(|x|) \in O\left(\operatorname{time}_{N_{0}}(|x|)^{k}\right)$ and space $_{N_{1}}(|x|) \in O\left(\right.$ space $\left._{N_{0}}(|x|)\right)$ and conversely.
    ${ }^{6}$ See [27] for discussion of caveats pertaining both to how the relevant time and space measures must be formulated and the status of the requirement that the relevant overheads be achieved with respect to the same simulation.
    ${ }^{7}$ Hartmanis and Simon showed this originally with respect to a variant of the MRAM model which allows parallel bit-wise Boolean operations in the manner described above. Bertoni et al. showed that such operations were not necessary in the presence of both addition and multiplication.

[^12]:    ${ }^{8}$ For if $\mathfrak{M}$ satisfied the Invariance Thesis with respect to $\mathfrak{R}$, this would also entail that there was a time and space efficient simulation of MRAM machines by multi-tape Turing machines from which $\mathbf{P}_{\mathfrak{T}}=\mathbf{P S P A C E} \mathbf{T}_{\mathfrak{T}}$ would follow.
    ${ }^{9}$ Schönhage and Strassen [23] showed that the $O\left(n^{2}\right)$ time complexity of the "naive" carry multiplication algorithm can be improved to $O(n \log (n) \log (\log n))$, while also conjecturing a lower bound of $\Omega(n \log (n))$. The Schönhage-Strassen bound has recently been improved by Fürer [10] who presented an $n \log n 2^{O\left(\log ^{*}(n)\right)}$ algorithm.

[^13]:    ${ }^{10}$ Although I will put aside such issues here, rejoinders to these objections can be found in textbooks such as [1]. See also [9].
    ${ }^{11}$ For instance, while FP coincides with the provably recursive functions of Buss's system $S_{2}^{1}$, similar characterizations of the Linear Time Hierarchy or Polynomial Hierarchy arise as the provably recursive functions of similar theories [3]. And while $\mathbf{P}_{\mathfrak{T}}$ is characterizable as the set of languages describable in first-order logic with a least fixed point operator in the presence of a linear order, similar (and in many instances simpler) characterizations of narrower classes (like $\mathrm{AC}^{0}$ ) or broader classes (like $\mathbf{N P}_{\mathfrak{T}}$ ) are also available [14]. See [4] for function algebra descriptions of additional complexity classes such as $\mathbf{P S P A C E} \mathbf{E}_{\mathfrak{T}}$.

[^14]:    ${ }^{12}$ The graph-theoretic models of Kolomogorv and Uspensky [16] and Schönhage [22] are inspired by similar considerations. However Gandy's model differs not only defining states in defining states more abstractly - i.e. as hereditarily finite sets (which may contain unboundedly many labels which are treated as urelements) of bounded rank but also in that it allows for parallelism in the sense that machines may locally transform discontiguous parts of states in a single step.

[^15]:    ${ }^{13}$ In fact it seems likely that $\mathfrak{G}$ is a not a reasonable model. For it is easy to see that Gandy machines can simulate Schönage's storage modification machines with constant factor time overhead. As the latter model admits a linear time (i.e. $O(n)$ ) multiplication algorithm, Gandy machines do as well. And as Cook [6] (p. 403) observes, "we are forced to conclude that either multiplication is easier than we thought or that Schönhage's machines cheat". See Schöngage [22] (pp. 506-507) and van Emde Boas [26] (p. 110) for similar observations.

[^16]:    Supported in part by JSPS Kakenhi projects 24106002 and 26700001, by EU FP7 IRSES project 294962, by DFG $\mathrm{Zi} 1009 / 4-1$ and by IRTG 1529. We thank Daniel Graça and Elvira Mayordomo for inviting this extended abstract as opportunity to report on our progress since its CCA 2015 short version.

[^17]:    ${ }^{1}$ The very first use of the term "type" in programming is probably Curry's [7], to distinguish between memory words containing instructions ("orders") and those containing data ("quantities"). These reports by Curry, as reconstructed by [10], contain a surprising, non-trivial mathematical theory of programs, up to a theorem analogous to the "well-typed expressions do not go wrong" of [23]! Despite G.W. Patterson's review on JSL 22(01), 1957, 102-103, we do not know of any influence of this theory on subsequent developments of programming languages.

[^18]:    2 "A type is a class of values. Associated with each type there are a number of operations which apply to such values".
    ${ }^{3}$ "Class" is, however, the terminology of Simula 67; in Simula I they are called activities.

[^19]:    ${ }^{4}$ Under this term Priestly refers to the "coherent and comprehensive research programme within which the Algol 60 report had the status of a paradigmatic achievement, in the sense defined by the historian of science Thomas Kuhn. This research programme established the first theoretical framework for studying not only the design of programming languages, but also the process of software development." Therefore, are grouped under this broad term the developments of structured programming, of software engineering à la Dijkstra, of the formal description of programming language semantics, etc.
    ${ }^{5}$ A language provides representation independence if two correct implementations of a single specification of an ADT are observationally indistinguishable by the clients of these types.

[^20]:    ${ }^{6}$ For a pedagogical discussion, see [11], Chap. 10.

[^21]:    ${ }^{7}$ An expression of Vladimir Voevodsky.

[^22]:    ${ }^{1}$ This question is addressed in a broader scope in Mounier-Kuhn (2010a).
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[^23]:    ${ }^{2}$ Archives de l'Observatoire de Paris, Ms 1061 II-2-D.
    ${ }^{3}$ See the papers by Anne Brygoo, Jean Carteron, Colette Hoffsaës and Félix Paoletti in the proceedings of Colloque sur l'histoire de l'informatique en France, 1988.
    ${ }^{4}$ The present text is only a provisional account of a work in progress, and should not be quoted without the authors' permission.

[^24]:    ${ }^{5}$ Our choice of J. Kuntzmann as a key actor is based on his role as a science entrepreneur and the availability of his correspondence conserved at the Archives départementales de l'Isère (Grenoble), henceforth ADI.
    ${ }^{6}$ J. Kuntzmann's letters to J. Pérès and R. Darmois, May 31956 (ADI IMAG 4).

[^25]:    ${ }^{7}$ J. Kuntzmann's letters to J. Heinhold, TH München, June 51956 (ADI IMAG 4). On the Darmstadt conference, see Petzold (2004).
    ${ }^{8}$ J. Kuntzmann's letter to AIC, October 21956 (ADI IMAG 4).
    ${ }^{9}$ J. Kuntzmann's letter to É. de Lacroix de Lavalette, May 41956 (ADI IMAG 4); and É. de Lacroix de Lavalette's historical report on his Groupe de calcul numérique, 1995. Private archives of the first author.

[^26]:    ${ }^{10}$ Evry Schatzmann's interview with J.-F. Picard, February 241987.
    ${ }^{11}$ Annuaire 1963 de l'AFCALTI. Archives privées de J. Carteron.

[^27]:    ${ }^{12}$ Correspondence between I. Auerbach, A. Danjon, J. Carteron, J. Kuntzmann and J. Ville, December 7-12 1957 (ADI IMAG 5) and Archives de l'Observatoire de Paris, Ms 1061 II-2-D.
    ${ }^{13}$ The 600 pages proceedings were published by the UNESCO both in Paris (1960) and London (1960). Pierre Auger happened to coin the term Traitement de l'information, translated from "Information processing". The journal of the congress can be found at http://unesdoc.unesco.org/images/0015/001537/153718fb.pdf.

[^28]:    ${ }^{16}$ P. Mounier-Kuhn's interview with Ph. Dreyfus, September 212007.

[^29]:    R. Solomon-I would like to thank the anonymous referee for many helpful suggestions to improve the presentation of the material in this article.

[^30]:    ${ }^{1}$ One can define prefix complexity in different ways, using prefix-free decompressors (no element of the domain is a prefix of another element of the domain) or prefixstable decompressors (if $D(x)$ is defined, then $D(y)=D(x)$ for every $y$ that has prefix $x$ ). The argument works only for prefix-free decompressors; the problem with the prefix-stable ones is that computation time of a prefix-stable decompressor is not a prefix-stable function. It would be interesting to know whether the result remains true for prefix-stable decompressors.

[^31]:    ${ }^{2}$ One may also speak about semicomputability for sequences that have terms $+\infty$ and/or $-\infty$; in this case we allow the values of the function $y(\cdot, \cdot)$ to be infinite.

[^32]:    O. Bournez-This author's research was partially supported by a French National Research Agency's grant (ANR-15-CE40-0016-02).
    N. Dershowitz-This author's research benefited from a fellowship at the Institut d'Études Avancées de Paris (France), with the financial support of the French National Research Agency's "Investissements d'avenir" program (ANR-11-LABX-0027-01 Labex RFIEA+).

[^33]:    ${ }^{1}$ In particular, this implies that an effective $\Pi_{2}$-statement must be true.

[^34]:    ${ }^{2}$ For some of the following results, this assumption is actually necessary: It is e.g. not hard to check that all choice principles considered in Sect. 4 are effective (and hence trivially reducible to each other) if $V=L$.

[^35]:    ${ }^{3}$ For example, suppose there is some minimal countable $\alpha$ such that $L_{\alpha} \models$ ZFC. Then the OTM-program that writes $L$ on the tape until an $L$-level satisfying ZFC will halt in $V$, but not inside $L_{\alpha}$.
    ${ }^{4}$ We thus make the implicit assumption that the miracle tape behaves deterministically, i.e. that, whenever the miracle command is applied to some $X$, the outcome will be the same. However, this property is not used anywhere in the arguments below. One may thus drop it, at the price of some extra formal complications.

[^36]:    ${ }^{6}$ Again, some condition on the height of $M$ is required to ensure that the convergence of programs is absolute between $V$ and $M$. In particular, a parameter-free OTM can run for more than $\alpha$ many steps, where $\alpha$ is minimal such that $L_{\alpha} \models \mathrm{ZF}^{-}$.

[^37]:    ${ }^{7}$ This does not imply that ZL is effective, though, as there is no effective way to effectively choose one element from the set of all maximal elements.

[^38]:    ${ }^{8}$ See $[\mathrm{Kr}]$, where this quote is attributed to Jerry Bona.

[^39]:    ${ }^{1}$ For more discussion, see [6].
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[^40]:    ${ }^{2}$ A Dollop of Standard Terminology: $W_{i}$ is the $i$-th c.e. set, where $i$ codes a program for generating or for accepting $W_{i}[30] .\langle\cdot, \cdot\rangle$ is a pairing function, i.e., a computable isomorphism from $\mathbb{N} \times \mathbb{N}$ to $\mathbb{N}[30]$, where $\mathbb{N}=$ the natural numbers. For $A \subseteq \mathbb{N}$, the $A$ computable (respectively, partial $A$-computable) functions are the total (respectively, partial) functions over $\mathbb{N}$ that are computable relative to oracle $A . W_{i}^{A}$ is the $i$-th $A$-c.e. set, where $i$ codes a relativized program that, with oracle $A$, generates or accepts $W_{i}^{A}$ [30]. For each $k \in \mathbb{N}$, let $A^{(k)}$ be the $k$-th jump [30] of $A$, i.e.: $A^{(0)}=A$ and $A^{(i+1)}=\left\{x: x \in W_{x}^{A^{(i)}}\right\}$. Thus, $\emptyset^{(1)}=$ the halting problem, $\emptyset^{(2)}=$ the jump of $\emptyset^{(1)}$, etc. For $A \subseteq \mathbb{N}, \bar{A}=\mathbb{N}-A$. Any unexplained terminology below is from [30].

[^41]:    ${ }^{6}$ Meyer and Bagchi's conjecture had $s_{A}$ as a computable size measure and $s_{B}=\lambda p . p$.
    ${ }^{7}$ Another dollop of terminology. For partial function $\psi, \psi(x) \downarrow$ means that $\psi$ is defined on $x$ and $\psi(x) \uparrow$ means that $\psi$ is not defined on $x$. Let $\left\langle\varphi_{p}^{A}\right\rangle_{p \in \mathbb{N}}$ be an acceptable programming system for the $A$-computable partial functions over $\mathbb{N}$. For this $\varphi^{A}$ we have (i) the s-1-1 theorem for $\varphi^{A}$ : There is a computable $s: \mathbb{N}^{2} \rightarrow \mathbb{N}$ such that, for each $p, x, y: \varphi_{s(p, x)}^{A}(y)=\varphi_{p}^{A}(\langle x, y\rangle)$; and (ii) the Kleene parametric recursion theorem for $\varphi^{A}$ : There is a computable $r: \mathbb{N}^{2} \rightarrow \mathbb{N}$ such that, for each $p, x, y: \varphi_{r(p, x)}^{A}(y)=$ $\varphi_{p}^{A}(\langle r(p, x), y\rangle) .(A, B)$ double m-reduces [35] to $(C, D)\left(\right.$ written: $\left.(A, B) \leq_{m}(C, D)\right)$ if and only if there is an computable $f$ such that, for each $x,[x \in A \Leftrightarrow f(x) \in C$ and $x \in B \Leftrightarrow f(x) \in D]$; such an $f$ witnesses the reduction.

[^42]:    ${ }^{8}$ One may wonder why $\langle\langle u$ and $v \in O\rangle\rangle$ (or the same thing but with 'settled' in place of ' $\in O^{\prime}$ ') is in the conclusion of Theorem 8. The simple answer is that $\emptyset^{(2)}$ is not a strong enough oracle to remove either one from inside the $\langle\langle\cdots\rangle\rangle$ in the conclusion.

[^43]:    ${ }^{1}$ The $O^{*}$ notation suppresses polynomial factors.

[^44]:    ${ }^{1}$ In this paper we will use gray boxes for Requirements. Requirements are properties that the definition of $\mathbb{R}_{\kappa}$ will have to satisfy.

[^45]:    ${ }^{2}$ Note that in [7] Dales and Woodin do not make use of surreal numbers giving a different construction of $\mathbb{R}_{\aleph_{1}}$.

[^46]:    ${ }^{3}$ Note that this also reflects the fact that $\mathrm{No}_{<\omega}$ are the dyadic numbers and $\mathbb{R}$ is the Cauchy completion of $\mathrm{No}_{<\omega}$.

[^47]:    A. Konovalov and V. Selivanov supported by RFBR project 13-01-00015a.

[^48]:    O.V. Kudinov-Supported by RFBR projects 13-01-00015a and 14-01-00376.
    V.L. Selivanov-Supported by RFBR project 13-01-00015a.

[^49]:    F. Yang-This author is sponsored by the China Scholarship Council (CSC).

[^50]:    P. Jonsson-was partially supported by the Swedish Research Council (VR) under grant 621-2012-3239.
    B. Martin-was supported by EPSRC grant EP/L005654/1.

[^51]:    ${ }^{1}$ Weird. Thus spake Lindemann about Hilbert's non-constructive methods in the resolution of Gordon's problem (see [28]).

[^52]:    ${ }^{1}$ Another typical solution is to give definition to complexity of relationally-infinite $\Gamma$ along the lines of "easy", if it is easy for all finite subsets, and "hard", if it is hard for some finite subset.

[^53]:    R. Miller-The author was supported by Grant \# DMS - 1362206 from the N.S.F., and by several grants from the PSC-CUNY Research Award Program. This work grew out of research initiated at a workshop at the American Institute of Mathematics and continued at a workshop at the Institute for Mathematical Sciences of the National University of Singapore. Conversations with Bjorn Poonen and Alexandra Shlapentokh have been very helpful in the creation of this article.

[^54]:    ${ }^{1}$ Bishop's original term is that of a separating set, which we avoid here in the presence of the standard classical notion of a separating subset of $C(X)$.

[^55]:    ${ }^{2}$ The proof goes as follows. By the constructive trichotomy property (see [3], p. 26) either $a<c$ or $a \wedge b<a$. In the first case we get immediately what we want to show. In the second case we get that $b \leq a$, since if $b>a$, we have that $a=a \wedge b<a$, which is a contradiction. Thus $a \wedge b=b$ and the hypothesis $a \wedge b<c$ becomes $b<c$.

[^56]:    ${ }^{3}$ By the definition of $\inf A$ in [3], p. 37, we have that $\forall_{\epsilon>0} \exists_{a \in A}(a<\inf A+\epsilon)$, therefore if $b>\inf A$ and $\epsilon=b-\inf A>0$ we get that $\exists_{a \in A}(a<\inf A+(b-\inf A)=b)$.

[^57]:    ${ }^{4}$ To show that $\neg\left(g^{*}(x)>f(x)+\epsilon\right)$ we just use the fact that if $A \subseteq \mathbb{R}, b \in \mathbb{R}$, then $\sup A>b \rightarrow \exists_{a \in A}(a>b)$. The function $g^{*}$ is mentioned in [19], where nonconstructive properties of the classical $(\mathbb{R},<)$ are used.

