Charles University in Prague
Faculty of Mathematics and Physics

## DOCTORAL THESIS



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# Study of Arithmetical Structures and Theories with Regard to Representative and Descriptive Analysis 

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Specialization: Algebra, Number Theory and Mathematical Logic

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Název práce: Studium aritmetických struktur a teorií s ohledem na reprezentační a deskriptivní analýzu

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#### Abstract

Abstrakt: Jsme motivováni otázkou vztahu lokálních a globálních vlastností operace $o$ ve struktuře tvaru $\langle\mathcal{B}, o\rangle$ s ohledem na aplikaci pro studium modelů $\langle\mathcal{B}, \cdot\rangle$ Peanovy aritmetiky, kde $\mathcal{B}$ je model aritmetiky Presburgerovy. Zajímá nás zejména problém závislosti, který formulujeme jako otázku určení uzávěru závislosti $$
\operatorname{icl}^{O}(E)=\left\{\bar{d} \in B^{n} ;\left(\forall o, o^{\prime} \in O\right)\left(o \upharpoonright E=o^{\prime} \upharpoonright E \Rightarrow o(\bar{d})=o^{\prime}(\bar{d})\right)\right\}
$$ kde $\mathcal{B}$ je struktura, $O$ množina $n$-árních operací na $B$ a $E \subseteq B^{n}$. Ukážeme, že tento problém lze převést na otázku definovatelnosti v jisté expanzi $\mathcal{B}$. Speciálně, je-li $\mathcal{B}$ saturovaný model Presburgerovy aritmetiky a $O$ množina všech (saturovaných) peanovských součinů na $\mathcal{B}$, dokážeme, že pro $a \in B$ je icl ${ }^{O}(\{a\} \times B)$ nejmenší možný, tj. obsahující právě ty dvojice $\left(d_{0}, d_{1}\right) \in B^{2}$, kde jedno z $d_{i}$ je tvaru $p(a)$ pro nějaký polynom $p \in \mathbb{Q}[x]$.

Uvedená problematika úzce souvisí s deskriptivní analýzou lineárních teorií, což jsou (až na změnu jazyka) teorie jistých diskrétně uspořádaných modulů nad určitými diskrétně uspořádanými obory integrity. Dokážeme tvrzení o eliminaci kvantifikátorů v lineárních teoriích a nalezneme prvomodely jejich jednoduchých kompletních extenzí. Provedeme detailní analýzu definovatelných množin v modelu $\mathcal{A}$ lineární teorie a odvodíme, že každá definovatelná množina je sjednocením lineárních obrazů mnohostěnů v $A^{n}$ pro nějaké $n \in \mathbb{N}$.

Zvláště důležitým příkladem lineární teorie je lineární aritmetika LA (přesněji její ,,Z्Z-verze" ZLA) - aritmetická teorie s plnou indukcí rozširirující Presburgerovu aritmetiku o násobení jediným nestandardním prvkem. Jako důsledek výše uvedeného dokážeme, že LA je modelově kompletní (eliminační množina je tvořena primitivně pozitivními formulemi) a rozhodnutelná, nalezneme její jednoduché kompletní extenze a sestrojíme jejich prvomodely. Dokážeme též, že modely LA jsou až na elementární ekvivalenci právě nehlavní ultraprodukty struktur $\left\langle\mathbb{N}, 0,1,+, \leq, n \cdot{ }_{-}\right\rangle$s $n \in \mathbb{N}$.

Jako algebraickou aplikaci uvedených výsledků ukážeme, že prvomodely jednoduchých kompletních extenzí LA určují $2^{\omega}$ různých oborů integrity $R \mathrm{~s} \mathbb{Z}[x] \subseteq$ $R \subseteq \mathbb{Q}[x]$, které jsou $\omega$-stage euklidovské, ale nejsou $k$-stage euklidovské pro žádné $0<k \in \mathbb{N}$. To řeší problém položený G. E. Cookem v Coo76].


## Klíčová slova:

lineární aritmetika, eliminace kvantifikátorů, Peanova aritmetika, extenze Presburgerovy aritmetiky, kvazieuklidovské okruhy

Title: Study of arithmetical structures and theories with regard to representative and descriptive analysis

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#### Abstract

: We are motivated by a problem of understanding relations between local and global properties of an operation $o$ in a structure of the form $\langle\mathcal{B}, o\rangle$, with regard to an application for the study of models $\langle\mathcal{B}, \cdot\rangle$ of Peano arithmetic, where $\mathcal{B}$ is a model of Presburger arithmetic. We are particularly interested in a dependency problem, which we formulate as the problem of describing the dependency closure $$
\operatorname{icl}^{O}(E)=\left\{\bar{d} \in B^{n} ;\left(\forall o, o^{\prime} \in O\right)\left(o \upharpoonright E=o^{\prime} \upharpoonright E \Rightarrow o(\bar{d})=o^{\prime}(\bar{d})\right)\right\}
$$ where $\mathcal{B}$ is a structure, $O$ a set of $n$-ary operations on $B$, and $E \subseteq B^{n}$. We show, that this problem can be reduced to a definability question in certain expansion of $\mathcal{B}$. In particular, if $\mathcal{B}$ is a saturated model of Presburger arithmetic, and $O$ is the set of all (saturated) Peano products on $\mathcal{B}$, we prove that, for $a \in B, \operatorname{icl}^{O}(\{a\} \times B)$ is the smallest possible, i.e. it contains just those pairs $\left(d_{0}, d_{1}\right) \in B^{2}$ for which at least one of $d_{i}$ equals $p(a)$, for some polynomial $p \in \mathbb{Q}[x]$.

We show that the presented problematics is closely connected to the descriptive analysis of linear theories. That are theories, models of which are - up to a change of the language - certain discretely ordered modules over specific discretely ordered integral domains. We prove a quantifier elimination result in linear theories, and we find the prime models of their simple complete extensions. We perform a detailed analysis of definable sets in a model $\mathcal{A}$ of a linear theory, and show that definable sets are unions of linear images of polyhedra in $A^{n}$, with $n \in \mathbb{N}$.

A particularly important example of linear theories is the linear arithmetic LA (more precisely, its "Z-like" variant ZLA). That is an arithmetical theory with the full induction, which extends Presburger arithmetic by multiplication by a single nonstandard element. As a corollary of the results above, we show that LA is model-complete (elimination set consists of primitive positive formulas) and decidable, we find its simple complete extensions and construct their prime models. We also prove that models of LA are, up to elementary equivalence, exactly all non-principal ultraproducts of the structures $\left\langle\mathbb{N}, 0,1,+, \leq, n \cdot{ }_{-}\right\rangle$, with $n \in \mathbb{N}$.

As an algebraic application of the presented results, we show that the prime models of the simple complete extensions of LA determine $2^{\omega}$ different integral domains $R$, with $\mathbb{Z}[x] \subseteq R \subseteq \mathbb{Q}[x]$, which are $\omega$-stage Euclidean, but not $k$-stage Euclidean, for any $0<k \in \mathbb{N}$. This solves the problem posed by G. E. Cooke in Coo76.


## Keywords:

linear arithmetic, quantifier elimination, Peano arithmetic, extensions of Presburger arithmetic, quasi-Euclidean rings

## Motto

Wir müssen wissen.
Wir werden wissen.

- David Hilbert

To everybody who will read with joy.

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## Introduction

In this thesis, we are motivated by a problem of understanding relations between local and global properties of an operation $o$ in a first-order structure of the form $\langle\mathcal{B}, o\rangle$, with a particular interest in the case where $\mathcal{B}$ is a model of Presburger arithmetic $\operatorname{Pr}$ and $o$ is a "Peano product" on $\mathcal{B}$, i.e. $\langle\mathcal{B}, o\rangle$ is a model of Peano arithmetic P .

## Dependency problem

The problem above may be specified as follows: Given a "background model" $\mathcal{B}$ and a set $O$ of all $n$-ary operations on $B$ satisfying certain global property (e.g. being a Peano product), we want to describe the dependency closure

$$
\operatorname{icl}^{O}(E)=\left\{\bar{d} \in B^{n} ;\left(\forall o, o^{\prime} \in O\right)\left(o \upharpoonright E=o^{\prime} \upharpoonright E \Rightarrow o(\bar{d})=o^{\prime}(\bar{d})\right)\right\}
$$

for $E \subseteq B^{n}$ (see 2.1.1). We call this task the ( $\left.\mathcal{B}, O, E\right)$-dependency problem.
A Peano dependency problem is a $(\mathcal{B}, O, E)$-dependency problem where $\mathcal{B}$ is a (saturated) model of Pr and $O$ is the set of all (saturated) Peano products on $\mathcal{B}$. Its solution may contribute to new constructions of models of arithmetic, different from the known methods (cuts, end-extensions, ultraproducts, ...; see the historical remark at the end of the introduction).

## Reduction to definability

A $(\mathcal{B}, O, E)$-dependency problem with saturated $\mathcal{B}$ may be solved by studying a definability problem in certain expansion of $\mathcal{B}$, called a fixator. This is formulated in the $D D$-theorem 2.1:2.

In chapter 2, Proposition 2.2:1 and Corollary 2.2:2, we completely solve an important case of the Peano dependency problem, for $E=E_{a}=\{a\} \times B$, with $a$ nonstandard (an "a-slice"). We prove that, in this case, $\operatorname{icl}(E)$ is as small as possible, i.e. it contains only the trivially dependent points $\bar{d}=\left(d_{0}, d_{1}\right)$ where at least one of $d_{i}$ equals $p(a)$, for some polynomial $p \in \mathbb{Q}[x]$.

By the DD-theorem, the key for the proof is understanding definability in the respective fixator. The fixator is a model of linear arithmetid ${ }^{1}$ LA - an extension

[^0]of $\operatorname{Pr}$ by multiplication by a nonstandard scalar, with the full induction scheme. The needed definability results follow from Corollary 1.4:71) of the Main Theorem on Linear Theories 1.3:4,

## Peano products

Proposition 2.2:1, in particular, enables us to construct interesting examples of (Peano) products. The first of them are meeting pairs of Peano products, constructed in Corollary 2.3:1, Another is a construction of a Robinson product which satisfies a portion of induction (Proposition 2.3:2).

Corollary 2.4:2 solves the question of possible interpolation of a Peano product through a given point $\langle\bar{b}, d\rangle \in B^{3}$.

## Linear theories

The mentioned fundamental Main Theorem on Linear Theories $1.3: 4$ is a result of the study of chapter 1 on linear theories - a class of theories which generalize the linear arithmetic LA. Models of linear theories are (up to a change of the language) certain (integrally-divisible) discretely ordered modules over specific (regularly quasi-Euclidean and with degrees) discretely ordered integral domains.

Linear theories are of its own interest, and a large part of this thesis is devoted to the analysis of them.

Besides a quantifier elimination result for linear theories, contained in the Main Theorem on Linear Theories 1.3:4, we perform a detailed analysis of terms (the Harmonic Form Theorem 1.3:6) and definable sets and functions in their models (corollaries of the Bases Theorem 1.3:8). In particular, we prove that every definable set in a model $\mathcal{A}$ of a linear theory is a finite union of linear images of polyhedra in $A^{k}$, for some $k \in \mathbb{N}$.

The theorems 1.3:4, 1.3:6 and 1.3:8 are proven in section 1.5. The proofs are based on three fundamental Propositions - S, H and B (1.5:2, 1.5:3 and 1.5:4) and on a calculus which is a generalization of the calculus of continued fractions.

Let us note that linear theories can be understood as extending both of the following - the theory of $\mathbb{Z}$-groups (which is, in fact, the simplest linear theory) and the theory of modules over some associative ring (the first one is extended by allowing multiplication by some non-integer scalars, the second one by adding an ordering). From this point of view, our results on linear theories generalize the classical results of Mojżesz Presburger Pre29] on $\mathbb{Z}$-groups (Presburger arithmetic) and, partially, the results of Walter Baur [Bau76] and Leonard Monk [Mon75] on modules.

## Two-sorted quantifier elimination for ordered ring-modules

In section 1.6, we apply our method of proof of the Theorem 1.3:4 to generalize and strengthen known quantifier elimination results for two-sorted structures of the type "ring-module" (or "ordered ring-ordered module").

Our Corollary 1.6:2 of Theorem 1.6:1 states a quantifier elimination result for doded-modules (see 1.6) - certain two-sorted structures of the type "ordered ring-ordered module" (in fact, just two-sorted variants of models of linear theories). This is an ordered analogue of the quite well-known result by Lou van den Dries and Jan Holly in vdDH92 for two-sorted unordered modules and it strengthens the result by Volker Weispfenning in Wei97, Theorem 4.1] for twosorted discretely ordered modules over the ring $\mathbb{Z}$ of integers (more precisely for the models of a two-sorted variant of Presburger arithmetic). See section 1.6 for more details.

Let us note that in vdDH92 the problem of generalizing the results to ordered modules (even for the simplest case of the module $\mathbb{Z}$ of integers) is considered as "very interesting" but as one that "seems to be very hard".

Our proof of Corollary 1.6:2 requires substantially different methods than those used in vdDH92 or Wei97. The absence of ordering in vdDH92 is, clearly, a great simplification. In Remark 1.5:1, we point out the reasons why the general case stated in Corollary $1.6: 2$ is essentially more difficult than the Presburger case from Wei97.

## Properties of linear arithmetic

As an application of the Main Theorem on Linear Theories 1.3:4 and results from section 1.2, we find basic model-theoretic properties of LA (Corollary 1.4:7) - we prove that LA is decidable and model-complete (in fact, that every formula is in LA equivalent to a disjunction of primitive positive formulas), we describe all its simple complete extensions and construct their prime models. Models of LA are characterized as non-principal ultraproducts of definable expansions of the standard model $\langle\mathbb{N}, 0,1,+, \leq\rangle$ (Corollary $1.4: 8)$. It is also proven that LA is equivalent to a theory La (see 1.1.4.1) which arises from LA by replacing the induction scheme by the scheme of integral divisibility (see 1.1.2). All is done in section 1.4.2.

Although the properties of LA are similar to those of Pr , the proof is much more difficult. This is, as we will argue in Remark 1.5:1, mainly due to the fact that, for $n \in \mathbb{N}$, any remainder modulo $n$ may be expressed as one of finitely many constant terms, while this is no more true for $n$ non-standard.

The results on LA contribute to the ongoing research on extensions of Presburger arithmetic (see the survey paper [Bès01] for details concerning this problematic).

## An application to the theory of quasi-Euclidean integral domains

Chapter 3 contains an interesting application of results from chapter 1 to the theory of quasi-Euclidean integral domains. We find $2^{\omega}$ different integral domains $R$, with $\mathbb{Z}[x] \subseteq R \subseteq \mathbb{Q}[x]$, which are quasi-Euclidean but not $k$-stage Euclidean, for any $0<k \in \mathbb{N}$. This solves an open question of George E. Cooke from Coo76.

The domains $R$ are constructed from the prime models of simple complete extensions of LA by taking their mirror "Z-like" versions and endowing them with a natural ring structure.

Remark: Historical remark on constructions of models of Peano arithmetic
Nonstandard models of Peano arithmetic were implicitely found by Kurt Gödel as a consequence of his famous incompleteness theorems Göd31, and were explicitely constructed (using an ultraproduct) by Thoralf Skolem only a few years later ([Sko33] and [Sko34]).

However, an enormous complexity of such models was apparent already in those days. By the incompleteness theorems, no class of elementary equivalent models of P can be recursively axiomatized. Moreover, Gödel's proof showed that each model of P contains such a complicated structure as a model of the finite set theory. These results were supported in late fifties by a theorem of Stanley Tennenbaum [Ten59], which stated that in every countable nonstandard model of P both, addition and multiplication, are non-recursive.

Despite of these facts, Robert MacDowell's and Ernst Specker's work on elementary end-extensions [MS61] showed that some constructions of interesting nonstandard models are possible. The results of Jeff Paris, Laurie Kirby and Leo Harrington ([Par78], [KP82] and [PH77]), obtained by methods of cuts and indicators, provided the first examples of natural combinatorial statements, true in the standard model but unprovable in P .

## Chapter 1

## Descriptive Analysis of Linear Theories

In this chapter, we are motivated by the problem of understanding the theory of linear arithmetic (LA). LA is an arithmetical theory in the language $L^{l i n}=$ $\langle 0,1,+, \underline{a}, \leq\rangle$, where $\underline{a}$ is intended as multiplication by a single element (see 1.1.4.1 for the axiomatic). It is a theory with the full induction for its language, and as such it should be understood as standing between Presburger (Pr) and Peano (P) arithmetics.

Our study of LA is led not only by our interest in the problem itself; in chapter 2 we will use it to prove non-trivial results about the structure of Peano products on a fixed saturated model of $\operatorname{Pr}$ (a more direct approach to this can be found in [Gli09]).

Instead of the "N-like" theory LA, we are going to work in its equivalent, but technically more pleasant, "Z-like" variant, which is denoted ZLA (see 1.1.4.2). Similarly, the $\mathbb{Z}$-like variant of Presburger arithmetic (additive arithmetic AA; see 1.1.3.1) is the theory of $\mathbb{Z}$-groups ( $\mathbb{Z}$-additive arithmetic; ZAA see 1.1.3.2).

The theories ZLA and ZAA are examples of what we call linear theories. That are theories, models of which are - up to a change of language - some (expansions by constants of) discretely ordered modules over certain discretely ordered integral domains. The module is required to be integrally-divisible over the domain (i.e. integer division works), and the domain needs to be regularly quasi-Euclidean (i.e. the regular Euclidean algorithm stops in finite time) and has to have degrees (such that $\operatorname{deg}(r) \leq \operatorname{deg}(q) \Leftrightarrow|r| \leq n|q|$ for some $n \in \mathbb{N}$ ); see section 1.3.1 for the detailed definitions. We consider the expansion

$$
\mathcal{F}=\left\langle F, 0,1,+,-, \leq, r, c, q^{-1}\right\rangle_{r \in \mathrm{D}_{\mathcal{F}}, c \in \mathrm{C}_{\mathcal{F}}, q \in+\mathrm{D}_{\mathcal{F}}}
$$

of a model described above by definitions of integral division $q^{-1}$ by all positive scalars $q$ from its integral domain $\mathrm{D}_{\mathcal{F}}$ and by some definable constants $c$, such that $\mathrm{C}_{\mathcal{F}}$ is the universe of a substructure of $\mathcal{F}$. Such an expansion $\mathcal{F}$ is called a lineal (see 1.3.1.3).

The main results of this chapter are the following three theorems, formulated in section 1.3.2: the Main Theorem on Linear Theories 1.3:4, the Harmonic Form Theorem 1.3:6 and the Bases Theorem 1.3:8,

The Main Theorem on Linear Theories states that in every lineal each nonempty set definable over parameters $\bar{a}$ contains the value $t(\bar{a})$ of a term $t$. Moreover, $t$ may be chosen "almost uniformly" with respect to different lineals and different tuples $\bar{a}$; see concepts of solvability (1.2.2.1) and almost uniform solvability (1.2.2.3). As a corollary of the Main Theorem, we get a quantifier elimination result for linear theories; in particular, we will see that every lineal admits quantifier elimination. We also describe all simple complete extensions of a given linear theory $T$ and their prime models (see Corollary 1.3:5).

Working in a fixed lineal, we show that every term can be, up to a "finite noise", equivalently written in harmonic form, i.e. as a linear combination of the basic harmonic functions $r^{-1}$ (see 1.3.2.2.1), and every formula is equivalent to a harmonic one. This is the Harmonic Form Theorem 1.3:6, Moreover, we perform a detailed analysis of definable sets in a lineal $\mathcal{F}$ and give a geometric characterization of them as unions of linear images of polyhedra in $F^{k}$, for some $k \in \mathbb{N}$. A similar characterization of definable functions as "piecewise linear" is also stated (see Corollary [1.3:9) of the Bases Theorem 1.3:8). The last result justifies the name "linear theory".

As we mentioned in the Introduction chapter, the results above generalize the classical results of Presburger [Pre29] on $\mathbb{Z}$-groups and, partially, of Baur [Bau76] and Monk Mon75] on modules.

We apply the Main Theorem on Linear Theories to determine basic properties of the $\mathbb{Z}$-linear arithmetic ZLA (see Theorem 1.4:5). It is shown that ZLA is model-complete (in fact, that every formula is in LA equivalent to a disjunction of primitive positive formulas) and decidable. The simple complete extensions of ZLA correspond to elements $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$, where $\mathbb{P}$ denotes the set of prime numbers, and $\mathbb{J}_{p}$, for $p \in \mathbb{P}$, the set of all $p$-adic integers. The prime models of the complete extensions are constructed as structures $\mathcal{C}_{\tau}$, with $\mathbb{Z}[x] \subseteq \mathrm{C}_{\tau} \subseteq$ $\mathbb{Q}[x]$. As a corollary, it is shown that the models of ZLA are, up to elementary equivalence, just ultraproducts $\mathcal{Z}_{\mathcal{U}}=\left(\prod_{n \in \mathbb{N}}\left\langle\mathbb{Z}, 0,1,+,-, \underline{n} \cdot{ }_{-}, \leq\right\rangle\right) / \mathcal{U}$, where $\mathcal{U}$ is a non-principal ultrafilter on $\mathbb{N}$ (see Corollary 1.4:6).

Let us note that the sets $\mathrm{C}_{\tau}$, endowed with the structure of a ring, are examples of integral domains, which are $\omega$-stage Euclidean but not $k$-stage Euclidean for any $k \in \mathbb{N}$. This answers the open question by G. E. Cooke from Coo76; see chapter 3 for more details on this topic.

The mentioned results about ZLA can be almost automatically translated to similar statements about LA (see Corollary 1.4:7) and understood as stating that LA is a theory typologically similar to $\operatorname{Pr}$ (and far away from P). Whether the same is true also for the theory $\mathrm{LA}^{2}$ (extension of $\operatorname{Pr}$ by multiplication by two independent scalars; see 1.1.5.1), is posed as the Open question 1.

Our quantifier elimination and decidability results for LA contribute to the long and ongoing research on extensions of Presburger arithmetic. A good survey paper for this problematics is [Bès01]. 11

Our proof of the three main theorems relies on a calculus of terms in a lineal. This calculus can be seen as a generalization of the calculus of continued fractions. Basic steps of the proof are sketched in 1.5.1.

In section 1.6, we formulate an easy consequence of the proof - a quantifier elimination result (Corollary 1.6:2 of Theorem 1.6:1) for certain two-sorted structures of the type "ordered ring-ordered module" (in fact, just for two-sorted variants of models of linear theories). As we already mentioned in the Introduction, this generalizes results by Lou van den Dries and Jan Holly in vdDH92 for two-sorted unordered modules and by Volker Weispfenning in Wei97] for twosorted discretely ordered modules over the ring $\mathbb{Z}$ of integers (more precisely for the models of a two-sorted variant of Presburger arithmetic). See section 1.6 for more details.

### 1.1 Arithmetical theories

We state here a few axiomatics of theories which will play a role of important and motivating examples in our further explanation.

All the presented theories are "arithmetics" or "Z-arithmetics", i.e. extensions of the basic additive arithmetic, in the language $L^{\text {add }}=\langle 0,1,+, \leq\rangle$ or $L_{\mathbb{Z}}^{\text {add }}=\langle 0,1,+,-, \leq\rangle$ respectively, by fragments of multiplication and the full induction scheme. As we already mentioned in the prologue of this chapter, this problematics is connected with the study of expansions of the structure $\langle\mathbb{N}, 0,1,+\rangle$; we again refer to the survey article [Bès01] for more details.

All of our example theories are extensions of the additive arithmetic by multiplication by some fixed scalars (i.e. extensions by some "slices" of the full binary multiplication). They form a linearly ordered chain between Presburger (Pr) and Peano ( P ) arithmetics (which are its least and largest elements), where the ordering is given by the number of scalars.

Besides the "N-like" and "Z-like" versions of the theories, we will distinguish two equivalent but different axiomatics for each theory - the "inductive" one (based on the induction scheme for all formulas) and the "divisible" one (based on the scheme of integral-divisibility). All the axiomatics, we define, are summarized in the Table 1.1.

[^1]| Level | $\mathbb{N}$-like ind. | $\mathbb{N}$-like div. | Z-like ind. | Z-like div. |
| :---: | :---: | :---: | :---: | :---: |
| additive | Pr=AA | Aa | ZAA | ZAa |
| linear | LA | La | ZLA | ZLa |
| $\kappa$-linear | LA $^{\kappa}$ | - | ZLA $^{\kappa}$ | - |
| Peano | P | - | ZP | - |

Table 1.1: Arithmetical theories

### 1.1.1 Induction and $\mathbb{Z}$-induction

Let us remind that, for a formula $\varphi$ in a language extending $\langle 0,1,+, \leq\rangle$, the axiom of induction for $\varphi$ is the following formula:

$$
I(\varphi):(\varphi(0) \&(\varphi(x) \rightarrow \varphi(x+1))) \rightarrow(\forall x) \varphi(x) .
$$

For a formula $\varphi$ in a language extending $\langle 0,1,+, \leq\rangle$, the following is called the axiom of $\mathbb{Z}$-induction for $\varphi$ :

$$
I_{\mathbb{Z}}(\varphi): \quad((\exists z) \varphi(z) \&(\varphi(x) \leftrightarrow \varphi(x+1))) \rightarrow(\forall x) \varphi(x)
$$

For a set $\Gamma$ of formulas, $I(\Gamma)$ and $I_{\mathbb{Z}}(\Gamma)$ denote the sets of all axioms $I(\varphi)$ and $I_{\mathbb{Z}}(\varphi)$, for $\varphi \in \Gamma$, respectively. If $\Gamma$ is the set of all $L$-formulas, we write $I(L)$, $I_{\mathbb{Z}}(L)$ instead of $I(\Gamma), I_{\mathbb{Z}}(\Gamma)$.

### 1.1.2 Integral divisibility

Let $\alpha$ be an unary term in a language extending $\langle 0,1,+, \leq\rangle$. The formula

$$
i d(\alpha):(\exists y)(\alpha(y) \leq x<\alpha(y+1))
$$

is called the axiom of integral-divisibility for $\alpha$.
Let $\Lambda$ be a set of unary terms. We define $i d(\Lambda)=\{i d(\alpha) ; \alpha \in \Lambda\}$.

### 1.1.3 Additive arithmetics

The following theories are just different axiomatics of Presburger arithmetic. We define them in order to introduce a consistent and systematic notation for all presented theories. We also want to specify the axioms we use for "Presburger arithmetic" since they vary through the literature.

### 1.1.3.1 AA and Aa

Additive arithmetic (AA) is a theory in the language $L^{\text {add }}=\langle 0,1,+, \leq\rangle$. Its axioms are

$$
\begin{array}{ccc}
\text { (A1) } & 0 \neq z+1 \quad(\mathrm{~A} 2) & x+1=y+1 \rightarrow x=y \\
\text { (A3) } & x+0=x \quad(\mathrm{~A} 4) & x+(y+1)=(x+y)+1 \\
& \left(\mathrm{D}_{\leq}\right) x \leq y \leftrightarrow(\exists z)(x+z=y)
\end{array}
$$

and the scheme $I\left(L^{\text {add }}\right)$ of induction for all $L^{\text {add_}}$-formulas. AA without the induction scheme is denoted $\mathrm{AA}^{-}$.

The axiomatic Aa contains the following axioms:

$$
\begin{gather*}
0 \neq z+1 \&(x \neq 0 \rightarrow(\exists z)(x=z+1)) \quad \text { (a2) } \quad x+z=y+z \rightarrow x=y  \tag{a1}\\
x+0=x \quad \begin{array}{c}
\text { (a4) }
\end{array} x+(y+z)=(x+y)+z  \tag{a3}\\
(\mathrm{a} 5) x+y=y+x
\end{gather*}
$$

and the scheme of integral-divisibility $i d(\underline{n})$, for $0<n \in \mathbb{N}$, where $\underline{n}(x)$ denotes $x+\cdots+x$, with $n$ summands. We also write $\underline{0}(x)=0, \underline{-n}(x)=-\underline{n}(x)$ and $\mathbf{z}=\underline{z} 1$, for $z \in \mathbb{Z}$. Aa without the integral-divisibility scheme is denoted $\mathrm{Aa}^{-}$.

### 1.1.3.2 ZAA and ZAa

$\mathbb{Z}$-additive arithmetic (ZAA) is the theory in the language $L_{\mathbb{Z}}^{\text {add }}=\langle 0,1,+,-, \leq\rangle$ consisting of the axioms

$$
\text { (ZA1) } \quad(\exists z)(x=z+1) \quad(\mathrm{ZA} 2) \quad x+1=y+1 \rightarrow x=y
$$

(ZA3) $\quad x+0=x \quad(\mathrm{ZA} 4) \quad x+(y+1)=(x+y)+1$

$$
\left(\mathrm{D}_{-}\right)-x+x=0
$$

$$
\text { (O1) } \quad(x=-1 \vee 0 \leq x) \leftrightarrow 0 \leq x+1 \quad \text { (OD) } \quad x \leq y \leftrightarrow 0 \leq-x+y
$$

and the scheme $I_{\mathbb{Z}}\left(L_{\mathbb{Z}}^{\text {add }}\right)$ of $\mathbb{Z}$-induction for all $L_{\mathbb{Z}}^{\text {add }}$-formulas. ZAA without the induction scheme is denoted $\mathrm{ZAA}^{-}$. Note that $(\mathrm{ZA} i)$ is the same axiom as (A $i$ ), for $i \neq 1$.

The axiomatic ZAa is given by the axioms

$$
\begin{gather*}
\text { (Za1) } \\
\begin{array}{cc}
(\exists z)(x=z+1) & (\mathrm{Za} 2)
\end{array} \quad x+z=y+z \rightarrow x=y \\
\text { (Za3) } \\
x+0=x
\end{gather*}(\mathrm{Za}) \quad x+(y+z)=(x+y)+z, ~(\mathrm{Za} 5) x+y=y+x .
$$

$$
\begin{array}{ccc}
x \leq 0 \vee 0 \leq x & (\mathrm{o} 2) & (x \leq 0 \& 0 \leq x) \rightarrow x=0 \\
(0 \leq x \& 0 \leq y) \rightarrow 0 \leq x+y & \text { (o4) } & x \leq 0 \vee 1 \leq x \tag{o3}
\end{array}
$$

$$
(\mathrm{oD}) x \leq y \leftrightarrow 0 \leq-x+y
$$

and the scheme of integral-divisibility $i d(\underline{n})$, for $0<n \in \mathbb{N}$. ZAa without the integral-divisibility scheme is denoted $\mathrm{ZAa}^{-}$.
Remark 1.1:1. It is not difficult to see that AA is an extension of Aa, and ZAA is an extension of ZAa. Later, we show that AA, Aa and ZAA, ZAa are both even equivalent (see Proposition 1.4:1). Moreover, ZAA is equivalent to the theory of $\mathbb{Z}$-groups, i.e. $\operatorname{Th}(\langle\mathbb{Z}, 0,1,+,-, \leq\rangle)$.

Lemma 1.1:2. The following is provable in ZAa:

1) the axioms of Abelian groups, i.e.
a) $x+(y+z)=(x+y)+z$,
b) $x+y=y+x$,
c) $x+0=x$,
d) $-x+x=0$,
2) $\leq$ is discrete linear ordering, with 1 as the least positive element, and compatible with + , i.e.
a) $\leq$ is a linear ordering,
b) $0<1, x \leq 0 \vee 1 \leq x$,
c) $(u \leq x \& v \leq y) \rightarrow u+v \leq x+y$.

Proof. Easy.

### 1.1.4 Linear arithmetics

### 1.1.4.1 LA and La

Linear arithmetic (LA) is a theory in the language $L^{l i n}=\langle 0,1,+, \underline{a}, \leq\rangle$, where $\underline{a}$ is an unary functional symbol (with the intended meaning "multiplication by a scalar $a "$ ). The axioms of LA are

$$
\begin{align*}
& \text { all axioms of } \mathrm{AA}^{-} \\
& \underline{a}(x+1)=\underline{a} x+\underline{a} 1 \quad(\mathrm{~L} 2) \quad 0 \leq \underline{a} 1  \tag{L1}\\
& (\mathrm{~L} 0) \underline{a} 1 \neq n \text { for all } n \in \mathbb{N}
\end{align*}
$$

and the scheme $I\left(L^{l i n}\right)$ of induction for all $L^{l i n}$-formulas. LA without the induction scheme is denoted $\mathrm{LA}^{-}$.

Remark 1.1:3. The name "linear arithmetic" is used somewhat vaguely and/or inconsistently through the literature. It denotes more different concepts where an important role is played by inequalities of "linear" combinations of "unknowns". Therefore, we stress that, for us, linear arithmetic means always the first order theory LA above.

The axiomatic La is given by

$$
\begin{align*}
& \text { all axioms of } \mathrm{Aa}^{-} \\
& \underline{a}(x+y)=\underline{a} x+\underline{a} y \quad \text { (12) } \quad 0 \leq x \rightarrow 0 \leq \underline{a} x  \tag{11}\\
& \text { (10) } \underline{a} 1 \neq n \text { for all } n \in \mathbb{N}
\end{align*}
$$

and the scheme of integral divisibility $i d(p)$, for every polynomial $0<p \in \mathbb{Z}[a]$. Here, for $0<p=\sum_{i<m} c_{i} a^{i} \in \mathbb{Z}[a]$, with $c_{i} \in \mathbb{Z}$, we denote $\underline{p}(x)$ the term $\sum_{i<m} \underline{c}_{i}\left(\underline{a}^{i}(x)\right)$, where $\underline{a}^{i}(x)=\underline{a}(\ldots(\underline{a}(x)) \ldots)$ with $n$ occurances of $\underline{a}$. Moreover, for $r=\frac{p}{n} \in \mathbb{Q}[a]$, with $p \in \mathbb{Z}[a]$ and $0<n \in \mathbb{N}$, we define $\underline{r}(x)=y \leftrightarrow \underline{p}(x)=\underline{n}(y)$. We write $\mathbf{r}$ for the constant term $\underline{r}(1)$.

La without the integral-divisibility scheme is denoted $\mathrm{La}^{-}$.
Example 1.1:4. Let $\mathcal{A}=\langle A, 0,1,+, \cdot, \leq\rangle$ be a non-standard model of Peano arithmetic (see 1.1.6), and let $a \in A-\mathbb{N}$. Then $\mathcal{A}^{a}=\langle A, 0,1,+, \underline{a}, \leq\rangle$, where $\underline{a}(x)=a \cdot x$, is a model of LA. Moreover, if $\mathcal{A}$ is $\kappa$-saturated then $\mathcal{A}^{a}$ is as well.

### 1.1.4.2 ZLA and ZLa

$\mathbb{Z}$-linear arithmetic (ZLA) is the theory in the language $L_{\mathbb{Z}}^{l i n}=\langle 0,1,+,-, \underline{a}, \leq\rangle$ with axioms
all axioms of $\mathrm{ZAA}^{-}$
(L1) $\underline{a}(x+1)=\underline{a} x+\underline{a} 1 \quad$ (L2) $0 \leq \underline{a} 1$
$(\mathrm{~L} 0) ~ a 1 \neq n$ for all $n \in \mathbb{N}$
and the scheme $I_{\mathbb{Z}}\left(L_{\mathbb{Z}}^{l i n}\right)$ of $\mathbb{Z}$-induction for all $L_{\mathbb{Z}}^{l i n}$-formulas. ZLA without the induction scheme is denoted ZLA ${ }^{-}$.

The axiomatic ZLa consists of axioms

> all axioms of ZAa ${ }^{-}$
> (11) $\quad \underline{a}(x+y)=\underline{a} x+\underline{a} y \quad(12) \quad 0 \leq x \rightarrow 0 \leq \underline{a} x$
(10) $\underline{a} 1 \neq n$ for all $n \in \mathbb{N}$
and the scheme of integral divisibility $i d(\underline{p})$, for every polynomial $0<p \in \mathbb{Z}[a]$.
ZLa without the integral-divisibility scheme is denoted ZLa ${ }^{-}$.

### 1.1.5 $\kappa$-linear arithmetics

### 1.1.5.1 $\mathrm{LA}^{\kappa}$

For a cardinal number $\kappa$, $\kappa$-linear arithmetic $\left(\mathrm{LA}^{\kappa}\right)$ is a theory in the language $L^{\kappa-l i n}=\left\langle 0,1,+, \underline{a}_{\alpha}, \leq\right\rangle_{\alpha<\kappa}$, where $\underline{a}_{\alpha}$ are unary functional symbols (with the intended meaning "multiplication by a scalar $a_{\alpha}$ ").

The axioms of $\mathrm{LA}^{\kappa}$ are
all axioms of $\mathrm{AA}^{-}$

$$
\underline{a}_{\alpha}(x+1)=\underline{a}_{\alpha} x+\underline{a}_{\alpha} 1 \quad(\kappa \mathrm{~L} 2) \quad 0 \leq \underline{a}_{\alpha} 1 \quad(\kappa \mathrm{~L} 3) \quad \underline{a}_{\alpha}\left(\underline{a}_{\beta} x\right)=\underline{a}_{\beta}\left(\underline{a}_{\alpha} x\right)
$$ ( $\kappa \mathrm{L} 0$ ) " $\underline{a}_{\alpha}$ is not definable by any formula not containing $\underline{a}_{\alpha}$ ",

for all $\alpha, \beta<\kappa$, and the scheme $I\left(L^{\kappa-l i n}\right)$ of induction for all $L^{\kappa-l i n}$-formulas. $\mathrm{LA}^{\kappa}$ without the induction scheme is denoted $\mathrm{LA}^{\kappa-}$.

### 1.1.5.2 $\mathrm{ZLA}^{\kappa}$

For a cardinal number $\kappa, \kappa$ - $\mathbb{Z}$-linear arithmetic $\left(\mathrm{ZLA}^{\kappa}\right)$ is the theory in the language $L_{\mathbb{Z}}^{\kappa-l i n}=\left\langle 0,1,+,-, \underline{a}_{\alpha}, \leq\right\rangle_{\alpha<\kappa}$ with the axioms
all axioms of $\mathrm{ZAA}^{-}$

$$
\begin{array}{ll}
(\kappa \mathrm{L} 1) & \underline{a}_{\alpha}(x+1)=\underline{a}_{\alpha} x+\underline{a}_{\alpha} 1 \quad(\kappa \mathrm{~L} 2) \quad 0 \leq \underline{a}_{\alpha} 1 \quad(\kappa \mathrm{~L} 3) \quad \underline{a}_{\alpha}\left(\underline{a}_{\beta} x\right)=\underline{a}_{\beta}\left(\underline{a}_{\alpha} x\right) \\
& (\kappa \mathrm{L} 0) \text { " } \underline{a}_{\alpha} \text { is not definable by any formula not containing } \underline{a}_{\alpha} ",
\end{array}
$$

for all $\alpha, \beta<\kappa$, and the scheme $I_{\mathbb{Z}}\left(L_{\mathbb{Z}}^{\kappa-l i n}\right)$ of $\mathbb{Z}$-induction for all $L_{\mathbb{Z}}^{\kappa-l i n}$-formulas. ZLA $^{\kappa}$ without the induction scheme is denoted ZLA ${ }^{\kappa-}$.

Remark 1.1:5. The 0 -linear arithmetic $\mathrm{LA}^{0}$ is just the additive arithmetic AA, while the 1-linear arithmetic $\mathrm{LA}^{1}$ is equivalent to the linear arithmetic LA.

### 1.1.6 Peano arithmetics

Peano arithmetic $(\mathrm{P})$ is a theory in the language $L^{a r}=\langle 0,1,+, \cdot, \leq\rangle$. Its axioms are
all axioms of $\mathrm{AA}^{-}$

$$
\begin{equation*}
x \cdot 0=0 \quad(\mathrm{M} 2) \quad x \cdot(y+1)=x \cdot y+x \tag{M1}
\end{equation*}
$$

and the scheme $I\left(L^{a r}\right)$ of induction for all $L^{a r}$-formulas. P without the induction scheme is denoted $\mathrm{P}^{-}$(the Robinson arithmetic Q is the extension of $\mathrm{P}^{-}$by $x \neq 0 \rightarrow(\exists z)(x=z+1))$.
$\mathbb{Z}$-Peano arithmetic (ZP) is the theory in the language $L_{\mathbb{Z}}^{a r}=\langle 0,1,+,-, \cdot, \leq\rangle$ given by the axioms

$$
\begin{align*}
& \quad \text { all axioms of } \mathrm{ZAA}^{-} \\
& x \cdot 0=0 \quad(\mathrm{M} 2) \quad x \cdot(y+1)=x \cdot y+x \tag{M1}
\end{align*}
$$

and the scheme $I_{\mathbb{Z}}\left(L_{\mathbb{Z}}^{a r}\right)$ of induction for all $L_{\mathbb{Z}}^{a r}$-formulas. ZP without the induction scheme is denoted $\mathrm{ZP}^{-}$.

Observation 1.1:6. The following diagram, where $T \vdash S$ denotes that $T$ is an extension of $S$, and where $1<\kappa<\lambda$, holds:


### 1.1.7 Models of $\mathbb{N}$-like and $\mathbb{Z}$-like variants

Let $T$ be one of the theories AA, Aa, LA, La, $\mathrm{LA}^{\kappa}, \mathrm{P}$, and $Z T$ be the corresponding $\mathbb{Z}$-like variant of $T$. It can be easily shown that, for every model $\mathcal{A}=$ $\langle A, 0, S,+, \leq, \ldots\rangle \models T$, its canonical " $\mathbb{Z}$-version" $\mathcal{A}^{ \pm}=\langle A \cup-A, 0,1,+,-, \leq, \ldots\rangle$ is a model of $Z T$. On the other side, the positive part $\mathcal{B}^{+}=\left\langle B^{+}, 0,1,+, \leq, \ldots\right\rangle$ of any model $\mathcal{B} \models Z T$ is a model of $T$.

For an $L(T)$-formula $\varphi$, the formula $\varphi^{+}$, created by replacing every quantification $Q x$ in $\varphi$ by $Q x \geq 0$, satisfies

$$
\begin{equation*}
\mathcal{A} \models \varphi[\bar{a}] \Leftrightarrow \mathcal{A}^{ \pm} \models \varphi^{+}[\bar{a}], \tag{1.1}
\end{equation*}
$$

for every $\bar{a} \in A$. Similarly, it is easy (but a bit more technical), for an $L(Z T)$ formula $\psi$, to construct an $L(T)$-formula $\psi^{ \pm}$, such that it is

$$
\begin{equation*}
\mathcal{A} \models \psi^{ \pm}[\bar{a}] \Leftrightarrow \mathcal{A}^{ \pm} \models \psi[\bar{a}], \tag{1.2}
\end{equation*}
$$

for every $\bar{a} \in A$.

### 1.2 Model-theoretical background

At this place, we formulate a theoretical background for our next explanation. The concepts presented in this section are simple and the proofs mostly elementary; the reason, why we introduce them, is that they show themselves very useful for formulating and proving the presented results.

### 1.2.1 $\quad \Sigma_{1}$-separability and decidability

We prove an easy but useful equivalent for decidability of recursively axiomatizable theories (Proposition 1.2:1). This criterion is particularly useful for theories which have "well described" but uncountable (and thus not recursively enumerable) set of non-equivalent simple complete extensions.

In the following, by a theory we mean (a numeric presentation of) some axiomatic in a recursive language $L$. Then $T h(T)$ denotes the set of all $L$-sentences provable in $T$ and $C S(T)$ the set of all $L$-sentences consistent with $T$.

### 1.2.1.1 $\Gamma$-separability

Let $\Gamma \subseteq \mathcal{P}(\mathbb{N})$. A theory $T$ is $\Gamma$-separable if there is $\mathcal{S} \subseteq C S(T), \mathcal{S} \in \Gamma$ and dense in $C S(T)$, i.e. such that for $\varphi \in C S(T)$ there is $\varphi^{\prime} \in \mathcal{S}$ with $T, \varphi^{\prime} \vdash \varphi$.

Proposition 1.2:1. For a recursively axiomatizable theory $T$, the following are equivalent:

1) $T$ is decidable,
2) $T$ is $\Delta_{1}$-separable,
3) $T$ is $\Sigma_{1}$-separable.

Proof. 1) $\Rightarrow 2): C S(T)$ is $\Delta_{1}$ and dense in itself.
2) $\Rightarrow 3$ ): Clear.
3) $\Rightarrow 1$ ): Let $\mathcal{S}$ be $\Sigma_{1}$ and dense in $C S(T)$. Then $\varphi \notin T h(T) \Leftrightarrow \neg \varphi \in C S(T) \Leftrightarrow$ there is $\varphi^{\prime} \in \mathcal{S}$ such that $\neg \varphi \in \operatorname{Th}\left(T, \varphi^{\prime}\right)$; therefore $\mathbb{N}-T h(T)$ is $\Sigma_{1}$.

### 1.2.1.2 $\Gamma$-almost-completion

A binary relation $\mathcal{C} \subseteq \mathbb{N}^{2}$ is a $\Gamma$-almost-completion of a theory $T$ if $\mathcal{C} \in \Gamma$, for each $n \in \operatorname{dom}(\mathcal{C})$ the set $\mathcal{C}[n]$ is a simple complete extension of $T$, and $\mathcal{C}$ is dense in the set of all simple complete extensions of $T$, i.e., for any $\varphi \in C S(T)$, there is $n \in \operatorname{dom}(\mathcal{C})$ such that $\varphi \in \operatorname{Th}(\mathcal{C}[n])$.

Proposition 1.2:2. If a recursively axiomatizable theory $T$ has a $\Sigma_{1}$-almostcompletion then it is decidable.

Proof. If $\mathcal{C}$ is a $\Sigma_{1}$-almost-completion of $T$ then $C S(T)=\bigcup_{n \in \operatorname{dom}(\mathcal{C})} \operatorname{Th}(\mathcal{C}[n])$ is $\Sigma_{1}$.

### 1.2.2 Solvable theories

We formulate a property of solvability of a theory $T$, which states that every non-empty definable set in a model of $T$ contains a "solution" expressible as a value of a term. In Proposition 1.2:6, we show that solvability implies quantifier elimination. The property of solvability will prove itself very helpful for the detailed analysis of definable sets in models of linear theories, which we perform in section 1.3.1.

We also show that if a theory $T$ satisfies a stronger condition of almost uniform solvability then every function definable in a model of $T$ can be written as a "piecewise term".

### 1.2.2.1 Solvable and $n$-solvable theories

We define concepts of solvable and $n$-solvable theories. Although these can (and will) be defined in general, we will make use of solvability and 0 -solvability only.

We say that a theory $T$ is [ $n$-]solvable [for $n \in \mathbb{N}$ ] if, for every model $\mathcal{M} \models T$, every $L$-formula $\varphi(x, \bar{y})[$ with $l(\bar{y}) \leq n]$ and an $l(\bar{y})$-tuple $\bar{a}$ from $M$, it holds

$$
\mathcal{M} \models(\exists x) \varphi(x, \bar{a}) \Rightarrow \mathcal{M} \models \varphi(t(\bar{a}), \bar{a}), \text { for some } L_{T}-\operatorname{term} t .
$$

Remark 1.2:3. It easily follows from the proof of Lemma $1.2: 5$ that, in the definition of solvable theory, we could equivalently replace the arbitrary $\varphi$ by a quantifier-free formula. However, this is not true for $n$-solvability. Let us also note that in both cases we may equivalently replace the single variable $x$ by a tuple $\bar{x}$.

The following statement is a classical result:
Lemma 1.2:4. Let $T$ be an L-theory and $\varphi(\bar{x})$ an L-formula, such that $l(\bar{x})>0$, or $L$ contains a constant symbol. Then the following are equivalent:

1) There is a quantifier-free $\psi(\bar{x})$, such that $T \vdash \varphi \leftrightarrow \psi$.
2) For any $\mathcal{M}, \mathcal{N} \models T$ with a common substructure $\mathcal{C}$ and every $\bar{a} \in C$, it is

$$
\begin{equation*}
\mathcal{M} \models \varphi(\bar{a}) \Leftrightarrow \mathcal{N} \models \varphi(\bar{a}) . \tag{1.3}
\end{equation*}
$$

Proof. Folklore.

## Lemma 1.2:5.

1) If a theory is solvable then it has quantifier elimination.
2) Suppose that $n>0$, or $L$ contains a constant symbol. If $T$ is $n$-solvable then every $\varphi(\bar{x})$, with $l(\bar{x}) \leq n$, is equivalent to some quantifier-free $\psi(\bar{x})$.
Proof. 1): Suppose that a theory $T$ is solvable and let $\mathcal{M}, \mathcal{N}$ and $\bar{a}$ be as in Lemma 1.2:42). Clearly, it is enough to prove (1.3) for a formula $(\exists x) \psi(x, \bar{y})$ with $\psi$ quantifier-free. Let $\mathcal{M} \models(\exists x) \psi(x, \bar{a})$. Then, by solvability, $\mathcal{M} \models \psi(t(\bar{a}), \bar{a})$ for some term $t$, and since $t(\bar{a}) \in \mathcal{C} \subseteq \mathcal{N}$, we have $\mathcal{N} \models(\exists x) \psi(x, \bar{a})$.
$2)$ : By induction on the least number $m$ of quantifiers in a prenex normal form of $\varphi$. The case $m=0$ is trivial. For the induction step, we may suppose that $\varphi$ is $(\exists x) \chi(x, \bar{y})$, for some $\chi$ such that $l(\bar{y}) \leq n$. Let $\mathcal{M}, \mathcal{N}, \bar{a}$ be as in Lemma 1.2:4 2), and suppose $\mathcal{M} \models \varphi(\bar{a})$. By $n$-solvability, there is a term $t$ such that $\mathcal{M} \models \chi(t(\bar{a}), \bar{a}) . \chi(t(\bar{y}), \bar{y})$ has at most $n$ free variables, hence it is in $T$ equivalent to a quantifier-free formula, by induction assumption. Therefore $\mathcal{N} \models \chi(t(\bar{a}), \bar{a})$, and $\varphi$ is equivalent to a quantifier-free formula by Lemma $1.2: 4$,

For a structure $\mathcal{M}$ and $X \subseteq M$, we denote $\mathcal{M}\langle X\rangle$ the substructure of $\mathcal{M}$ generated by $X$ and $M\langle X\rangle$ its universe. $\mathcal{M}_{(X)}$ stands for the structure of all definable elements in $\mathcal{M}$ over $X$, and $M_{(X)}=\left\{a \in M ;\{a\} \in \operatorname{Df}^{1}(X, \mathcal{M})\right\}$ denotes its universe.

Proposition 1.2:6. For a theory $T$ in a language with a constant symbol, it is equivalent:

1) $T$ is solvable.
2) $T$ has quantifier elimination and is axiomatizable by open formulas.
3) $T$ is model-complete and is axiomatizable by open formulas.
4) $\operatorname{For} \mathcal{N} \subseteq \mathcal{M} \models T$, it is $\mathcal{N} \prec \mathcal{M}$.
5) For $\mathcal{M} \models T, X \subseteq M$, it is $M\langle X\rangle=M_{(X)}$, and it is a dense set (of all atoms) in $\operatorname{Df}^{1}(X, \mathcal{M})$.

Proof. 1) $\Rightarrow 2$ ): $T$ has quantifier elimination by Lemma 1.2:5, Let $\mathcal{M} \subseteq \mathcal{N} \models T$. Then, by the Tarski-Vaught test, it is $\mathcal{M} \prec \mathcal{N}$, and hence $\mathcal{M} \models T$. Indeed: For $\bar{a} \in M$, if $\mathcal{N} \models(\exists x) \varphi(x, \bar{a})$, there is, by solvability of $T$, a term $t(\bar{y})$ such that $\mathcal{N} \equiv \varphi(t(\bar{a}), \bar{a})$, and clearly $t(\bar{a}) \in M$.
2) $\Rightarrow 3$ ): Clear.
$3) \Rightarrow 4$ ): We get $\mathcal{N} \models T$, by open-axiomatizability of $T$, and thus $\mathcal{N} \prec \mathcal{M}$, by model-completeness.
$4) \Rightarrow 5)$ : Let $\{a\} \in \operatorname{Df}^{1}(X, \mathcal{M})$, and $\varphi(x, \bar{b})$, with $\bar{b} \in X$, define $\{a\}$ in $\mathcal{M}$. Then, by $\mathcal{M}\langle X\rangle \prec \mathcal{M}$, there is exactly one $a^{\prime} \in M\langle X\rangle$ such that $\mathcal{M}\langle X\rangle \models \varphi\left(a^{\prime}, \bar{b}\right)$, and hence also $\mathcal{M} \models \varphi\left(a^{\prime}, \bar{b}\right)$, which implies $a=a^{\prime}$. Density of $M\langle X\rangle$ : Let $\emptyset \neq D \in \operatorname{Df}^{1}(X, \mathcal{M})$. Then, by $\mathcal{M}\langle X\rangle \prec \mathcal{M}$, it is $D \cap M\langle X\rangle \neq \emptyset$.
$5) \Rightarrow 1)$ : Let $\mathcal{M} \models T, \bar{a} \in M$, and $\psi(x, \bar{y})$ be an open formula such that it is $\mathcal{M} \vDash(\exists x) \psi(x, \bar{a})$. Then, by the assumptions, there is an atom of the form $\{t(\bar{a})\}$, with $t$ a term, under $\{b ; \mathcal{M} \models \psi(b, \bar{a})\}$, and clearly $\mathcal{M} \models \psi(t(\bar{a}), \bar{a})$.

We formulate the following two strengthenings of solvability. Let us remind that, by Remark 1.2:3, in the definition of solvable theory, it suffices to verify the condition only for quantifier-free formulas.

### 1.2.2.2 Uniformly solvable theory

We say that a theory $T$ is uniformly solvable if, for each quantifier-free formula $\psi(x, \bar{y})$, there is a term $t$ such that

$$
T \vdash(\exists x) \psi(x, \bar{y}) \rightarrow \psi(t(\bar{y}), \bar{y}) .
$$

Example 1.2:7. Let $L$ be the language $\langle 0,1, P\rangle$, where 0,1 are two constant symbols, and $P$ is an unary predicate symbol. Then the theory

$$
T=\{0 \neq 1,(\forall x)(x=0 \vee x=1),(\exists!x) P(x)\}
$$

is solvable but not uniformly solvable.

### 1.2.2.3 Almost uniformly solvable theory

We say that a theory $T$ is almost uniformly solvable if, for each quantifier-free formula $\psi(x, \bar{y})$, there are finitely many terms $t_{i}, i<n$, such that

$$
T \vdash(\exists x) \psi(x, \bar{y}) \rightarrow \bigvee_{i<n} \psi\left(t_{i}(\bar{y}), \bar{y}\right)
$$

By the following lemma, almost uniform solvability may be formulated as a seemingly weaker condition.

Lemma 1.2:8. The following are equivalent:

1) $T$ is almost uniformly solvable.
2) For each quantifier-free formula $\psi(x, \bar{y})$ and $\mathcal{M} \models T$, there are finitely many terms $t_{i}, i<n$, such that $\mathcal{M} \equiv(\exists x) \psi(x, \bar{y}) \rightarrow \bigvee_{i<n} \psi\left(t_{i}(\bar{y}), \bar{y}\right)$.
Proof. 1) $\Rightarrow 2$ ) is clear.
$2) \Rightarrow 1$ ): For a model $\mathcal{M} \models T$, we get $n^{(\mathcal{M})}$ and $t_{j}^{(\mathcal{M})}$, for $j<n^{(\mathcal{M})}$. We set $X=\left\{t_{j}^{(\mathcal{M})} ; \mathcal{M} \models T, j<n^{(\mathcal{M})}\right\}$ and

$$
S=T \cup\{(\exists x) \psi(x, \bar{c})\} \cup\left\{\bigwedge_{t \in F} \neg \psi(t(\bar{c}), \bar{c}) ; F \subseteq X \text { finite }\right\}
$$

where $\bar{c}$ are new constant symbols with $l(\bar{c})=l(\bar{y})$.
We prove that $S$ is inconsistent. Otherwise it has a model $\langle\mathcal{M}, \bar{c}\rangle$ such that

$$
\mathcal{M} \vDash(\exists \bar{y})\left((\exists x) \psi(x, \bar{y}) \& \bigwedge_{j<n^{(\mathcal{M})}} \neg \psi\left(t_{j}^{(\mathcal{M})}, \bar{y}\right)\right),
$$

and $\mathcal{M} \vDash T$; this contradicts the choice of the terms $t_{j}^{(\mathcal{M})}$. Hence, for some finite $F \subseteq X$, it is $T \vdash(\exists x) \psi(x, \bar{c}) \rightarrow \bigvee_{t \in F} \psi(t(\bar{c}), \bar{c})$, and thus

$$
T \vdash(\exists x) \psi(x, \bar{y}) \leftrightarrow \bigvee_{t \in F} \psi(t, \bar{y})
$$

### 1.2.2.4 Piecewise terms

An [open] piecewise term, or shortly an [open] p-term, is a tuple $\tau=\left(t_{i}, \psi_{i}\right)_{i=0}^{n-1}$, more suggestively written as a piecewise defined function

$$
\begin{equation*}
\tau(\bar{x})=\left\{t_{i}(\bar{x}) \text { if } \psi_{i}(\bar{x}) ; i=0, \ldots, n-1\right. \tag{1.4}
\end{equation*}
$$

where $n>0, t_{i}$ are terms, and $\psi_{i}$ are [open] formulas such that, for every $\bar{a}$, exactly one of $\psi_{i}(\bar{a})$ holds. A subterm of a p-term $\tau=\left(t_{i}, \psi_{i}\right)_{i=0}^{n-1}$ is any subterm of some $t_{i}$ or of some $\psi_{i}$.

The value of a term $\tau(\bar{x})$ from (1.4) at point $\bar{a} \in M$ in structure $\mathcal{M}$ is $t_{i}(\bar{a})$, where $i$ is such that $\mathcal{M} \models \psi_{i}(\bar{a})$. We identify each term $t$ with the p-term $(t, \top)$ (here $T$ denotes "truth").

Proposition 1.2:9. Let $T$ be almost uniformly solvable, $\mathcal{M} \models T$ and $X \subseteq M$. Then every $X$-definable (in $\mathcal{M}$ ) function $f: M^{n} \rightarrow M$ is the realization of an open $p$-term $\tau(\bar{y}, \bar{a})$ with $\bar{a} \in X$.

Proof. Clearly, $T$ is solvable, hence it has quantifier elimination, by Proposition 1.2:6. Therefore we may suppose that $f$ is defined in $\mathcal{M}$ by an open formula $\chi(x, \bar{y}, \bar{a})$, with $\bar{a} \in X$. By the almost uniform solvability of $T$, there are terms $t_{i}(\bar{y}, \bar{z})$, for $i<m$, such that $\mathcal{M} \models \bigvee_{i<m} \chi\left(t_{i}(\bar{y}, \bar{a}), \bar{y}, \bar{a}\right)$.

Now, we may set $\tau=\left(t_{i}, \psi_{i}\right)_{i<m}$, where $\psi_{i}$ is the formula

$$
\chi\left(t_{i}(\bar{y}, \bar{a}), \bar{y}, \bar{a}\right) \& \bigwedge_{j<i} \neg \chi\left(t_{j}(\bar{y}, \bar{a}), \bar{y}, \bar{a}\right) .
$$

### 1.2.3 Solvable extensions by definitions

In this section, we deal with theories which have solvable extensions by definitions. Proposition 1.2:11 provides a characterization of such theories.

Example 1.2:10. All arithmetical theories with the full induction, such as Pr , LA or P (see section 1.1 for definitions), can be extended by definitions of new functions in such a way that the resulting extensions are solvable. For each formula $\varphi(x, \bar{y})$, it suffices to define a function which, for given parameters $\bar{a}$, outputs the minimal element of the set defined by $\varphi(x, \bar{a})$ (if it is non-empty).

Let $L=L^{\mathcal{F}} \cup L^{\mathcal{R}}$ be a language containing the constant symbol $0, L^{\mathcal{F}}$ be its algebraic part and $L^{\mathcal{R}}$ the relational part. For an $L$-formula $\varphi(\bar{x}, y)$, we denote
$\operatorname{cor}(\varphi)$ : the $L$-formula " $\varphi$ is a correct defining formula of a function", $\delta(\varphi)$ : the conditional definition of $\varphi$ :

$$
(\operatorname{cor}(\varphi) \& \varphi(\bar{x}, \underline{\varphi}(\bar{x}))) \vee(\neg \operatorname{cor}(\varphi) \& \underline{\varphi}(\bar{x})=0)
$$

where $\underline{\varphi}$ is a new $l(\bar{x})$-ary functional symbol. For $F \subseteq F m_{L}$, we write

$$
T^{F}=T \cup\{\delta(\varphi) ; \varphi \in F\}
$$

for the extension of $T$ by functions definable by formulas from $F$. For an $L$ structure $\mathcal{M}$, we denote $\mathcal{M}^{F}$ the corresponding expansion of $\mathcal{M}$ and $L^{F}$ the respective extension of $L$. If $s$ is a new symbol of $T^{F}$, we write $\dot{s}$ for the $L$ formula $\varphi$ which defines $s$ (i.e. such that $s=\underline{\varphi}$ ). When there is no danger of misunderstanding, we write just $s$ instead of $\dot{s}$.

We say that $T$ is $F$-solvable [ $F$ - $n$-solvable] if $T^{F}$ is solvable [ $n$-solvable]. Then $\emptyset$-solvable means just solvable and similarly for the $n$-solvability.

Proposition 1.2:11. Let $T$ be an $L$-theory and $n \in \mathbb{N}$. The following statements are equivalent:

1) $T$ is $\mathrm{Fm}_{L}-[n-]$ solvable.
2) For every $\mathcal{M} \models T$ and $X \subseteq M$ [with $|X| \leq n]$, each non-empty $X$-definable set in $\mathcal{M}$ contains an $X$-definable element.
3) For every $\mathcal{M} \vDash T$ and $X \subseteq M[$ with $|X| \leq n]$, it is $\mathcal{M}_{(X)} \prec \mathcal{M}$.
4) For every $\mathcal{M} \models T$ and $X \subseteq M$ [with $|X| \leq n]$, the structure $\mathcal{M}_{(X)}$ is a prime model over $X$ in $\mathcal{M}$.

Proof. 1) $\Rightarrow 2$ ) is immediate.
2) $\Rightarrow$ 3) follows from the Tarski-Vaught test.
$3) \Rightarrow 1$ ) is easy by observing that, for $b \in \mathcal{M}_{(\bar{a})}$, there is a $\emptyset$-definable function $f$ with $f(\bar{a})=b$.
$3) \Rightarrow 4)$ : Let $\mathcal{N}$ be an $L$-structure and $f: M \rightarrow N$ a partial elementary map, with $\operatorname{dom}(f)=X$. Then $f$ can be extended to an isomorphism between $\mathcal{M}_{(X)}$ and $\mathcal{N}_{(f[X])}$, by setting $f(a)=b$, where $b \in N$ is defined in $\mathcal{N}$ by the same $L$ formula as $a$ in $\mathcal{M}$. Then, by 3 ), $f$ is an elementary embedding.
$4) \Rightarrow 3)$ : Clearly, the only elementary embedding of $\mathcal{M}_{(X)}$ into $\mathcal{M}$ extending the identity on $X$ is the identity.

The following corollary states that the prime models and the simple complete extensions of $F$-0-solvable theories are at most as "complex" as $F$.

Corollary 1.2:12. Let $T$ be an L-theory and $F \subseteq F m_{L}$ be such that $T$ is $F$-0solvable. Then, for $\mathcal{M} \models T$, it is:

1) $\mathcal{M}_{(\emptyset)}=\mathcal{M}^{F}\langle\emptyset\rangle \mid L$ is the unique (even if $\|L\|>\omega$ ) prime model of $\operatorname{Th}(\mathcal{M})$,
2) $T h(\mathcal{M})$ is equivalent to $T \cup O T h_{L}\left(\mathcal{M}^{F}\right)$ (equivalently to $T \cup O T h_{L}\left(\mathcal{M}^{F}\langle\emptyset\rangle\right)$ ),
where $\operatorname{OTh}_{L}(\mathcal{N})$ denotes the set of canonical L-translations of open sentences true in $\mathcal{N}$.

Moreover, if $T^{F}$ is open-complete then $T$ is complete.
Proof. 1) $\mathcal{M}_{(\emptyset)}=\mathcal{M}^{F}\langle\emptyset\rangle \mid L$ is a trivial consequence of 0 -solvability of $T^{F}$. $\mathcal{M}_{(\emptyset)}$ is a prime model by Proposition 1.2:11. The uniqueness is clear. 2): Let $\mathcal{N} \models T \cup O T h_{L}\left(\mathcal{M}^{F}\right)$. Then $\mathcal{M}^{F}\langle\emptyset\rangle \cong \mathcal{N}^{F}\langle\emptyset\rangle$, and thus $\mathcal{M} \equiv \mathcal{N}$, by 1 ). "Moreover" follows from Lemma 1.2:5 2).

### 1.2.4 Syntactic presentation of prime models

In the previous section, we proved that prime models of simple complete extensions of a $F$-0-solvable $L$-theory can be seen as having universes consisting of (some) constant $L^{F}$-terms (see Corollary $1.2: 121$ )). We, informally, call this presentation of the prime models "syntactic".

The way in which syntactic presentations of the prime models of different complete extensions of a theory $T$ overlap, provides an interesting information about properties of $T$. In this section, we draft some possibilities of this method.

Let further $T$ be an $L$-theory, and $F \subseteq F m_{L}$ be such that $T$ is $F$-0-solvable. For $\mathcal{M} \vDash T$, we denote $\mathcal{M}_{\bullet, F}$ the $L^{F}$-structure of constant terms of the theory $T h\left(\mathcal{M}^{F}\right)$; the realization $r_{\mathcal{M}}$ of a symbol $r \in L^{\mathcal{R}}$ in $\mathcal{M}_{\bullet, F}$ is given by

$$
\begin{equation*}
r_{\mathcal{M}}(\bar{s}) \Leftrightarrow \mathcal{M}^{F} \models r(\bar{s}), \tag{1.5}
\end{equation*}
$$

for any tuple $\bar{s}$ of constant $L^{F}$-terms. By the formula (1.5) for $r$ equal to $=$, we also define the equivalence $=\mathcal{M}$ on the set of all constant $L^{F}$-terms.

For a constant $L^{F}$-term $s$, we write $\operatorname{cor}(s)$ instead of $\bigwedge\{\operatorname{cor}(\varphi) ; \underline{\varphi}$ occurs in $s\}$, and we denote $M_{c, F}=\{s ; \mathcal{M} \models \operatorname{cor}(s)\}$. Finally, we define the canonical structure of $\operatorname{Th}(\mathcal{M})^{F}$ as $\mathcal{M}_{*, F}=\mathcal{M}_{\bullet, F} /=_{\mathcal{M}}$.

By Corollary 1.2:12 1), $\emptyset$-definable elements in $\mathcal{M}$ are just the values of constant $L^{F}$-terms, and they form the universe of the prime-model of $\operatorname{Th}(\mathcal{M})$. Thus, we have the following:

Observation 1.2:13. $\mathcal{M}_{*, F} \mid L \cong \mathcal{M}_{(\emptyset)}$ is the unique prime model of $\operatorname{Th}(\mathcal{M})$, and thus

$$
\mathcal{M} \equiv \mathcal{N} \Leftrightarrow \mathcal{M}_{*, F}=\mathcal{N}_{*, F},
$$

for $\mathcal{M}, \mathcal{N} \models T$.
It is easy to see that every equivalence-class $[s]_{{ }_{\mathcal{M}}} \in M_{*, F}$ contains some $s^{\prime} \in M_{c, F}$. The set $M_{c, F}$ itself may carry some information about the structure $\mathcal{M}_{*, F}$. That is why we define the $F$-correctness diagram of $\mathcal{M}$ :

$$
\Delta^{c o r}(\mathcal{M}, F)=\left\{\operatorname{cor}(\varphi) ; \underline{\varphi} \in M_{c, F}\right\} \cup\left\{\neg \operatorname{cor}(\varphi) ; \underline{\varphi} \notin M_{c, F}\right\} .
$$

### 1.2.4.1 Completeness up to/in correctness

The two extremal cases, where $M_{c, F}$ carries the full and no information about $\mathcal{M}_{*, F}$, are considered in the following definitions.

A theory $T$ is said to be complete up to correctness w.r.t. $F$, or shortly $F$-cuc, if, for every $\mathcal{M}, \mathcal{N} \models T$, it is $M_{c, F}=N_{c, F} \Rightarrow \mathcal{M}_{*, F}=\mathcal{N}_{*, F}$.
$T$ is called complete in correctness w.r.t. $F$, or shortly $F$-cic, if, for every $\mathcal{M}, \mathcal{N} \models T$, it is $M_{c, F}=N_{c, F}$.

### 1.2.4.2 Compatibility

Two $L$-structures $\mathcal{M}, \mathcal{N}$ are said to be $F$-compatible if, for every $r \in L^{\mathcal{R}}$ or $r$ being $=$, it is $r_{\mathcal{M}}=r_{\mathcal{N}}$ on the set $M_{c, F} \cap N_{c, F} . T$ is $F$-compatible if every pair $\mathcal{M}, \mathcal{N} \models T$ is.

Observation 1.2:14. The following implications hold:

1) $T$ is complete $\Leftrightarrow T$ is F-cuc and F-cic.
2) $T$ is complete $\Rightarrow T$ is $F$-compatible $\Rightarrow T$ is $F$-cuc.

The following examples show that the implications in the Observation 1.2:14 2) can not be reversed.

Example 1.2:15.
a) Let $\operatorname{Pr}_{c}=\operatorname{Pr} \cup\{c>\underline{n} ; n \in \mathbb{N}\}$, with $c$ a new constant symbol, and $F$ be the set of the canonical formulas formally defining fractions of the form $\frac{m c+n}{k}$, with $m, n, k \in \mathbb{Z}, k>0$. Then $\operatorname{Pr}_{c}$ is $F$-0-solvable and $F$-compatible but not complete.
b) Let $\operatorname{Pr}_{c, d}=\operatorname{Pr} \cup\{c, d>\underline{n} ; n \in \mathbb{N}\}$, with $c, d$ new constant symbols, and $F$ be the set of the canonical definitions of fractions of the form $\frac{m c+n d+i}{k}$, with $m, n, i, k \in \mathbb{Z}, k>0$. The theory

$$
T=\operatorname{Pr}_{c, d} \cup\{(2 \mid c \& n c<d) \vee(\neg 2 \mid c \& n d<c) ; n \in \mathbb{N}\}
$$

is $F$-0-solvable and $F$-cuc but not $F$-compatible.

### 1.2.4.3 Prime-envelope

We show that all prime models of an $F$-compatible theory $T$ can be "faithfully" embedded into a single structure - a prime-envelope of $T$.

An $L^{F}$-structure $\mathcal{Q}$ is called an $F$-prime-envelope of $T$ if every $\mathcal{M}_{*, F}$, with
 and $\mathcal{Q}$ is generated by $\bigcup_{\mathcal{M} \models T} M_{*, F}$.

Proposition 1.2:16. Let $F \subseteq F m_{L}$ be such that $T$ is $F$-0-solvable. Then the following holds:

1) $T$ is $F$-cuc $\Leftrightarrow T h(\mathcal{M})$ is equivalent to $T \cup \Delta^{c o r}(\mathcal{M}, F)$, for every $\mathcal{M} \models T$.
$\Leftrightarrow \bigcup_{\mathcal{M} \models T}\left\{\bigwedge \Gamma ; \Gamma \subseteq \Delta^{c o r}(\mathcal{M}, F)\right.$ finite $\}$ is dense in the set $C S(T)$ of all sentences consistent with $T$.
2) $T$ is $F$-cic $\Leftrightarrow T \vdash \Delta^{\text {cor }}(\mathcal{M}, T)$, for every $\mathcal{M} \vDash T$.

$$
\Leftrightarrow \quad \Delta^{c o r}(\mathcal{M}, F)=\Delta^{c o r}(\mathcal{N}, F), \text { for every } \mathcal{M}, \mathcal{N} \models T
$$

3) $T$ is $F$-compatible $\Leftrightarrow T$ has an $F$-prime-envelope.
$\Leftrightarrow \quad$ The theory $T \cup\left\{\operatorname{cor}(\varphi) ; \varphi \in F^{\prime}\right\}$ decides all atomic $L^{F^{\prime}}$-sentences, for every finite $F^{\prime} \subseteq F$.

Proof. 1): The first " $\Leftrightarrow$ " is an easy consequence of Observation 1.2:13. The second " $\Leftrightarrow$ " is trivial.
2): Directly from definition.
3): The first equivalence: " $\Rightarrow$ ": We set $Q=\bigcup_{\mathcal{M} \vDash T} \mathcal{M}_{c, F} /={ }_{\mathcal{Q}}$, where $=_{\mathcal{Q}}$ is the transitive closure of $\bigcup_{\mathcal{M} \models T}\left(=_{\mathcal{M}} \upharpoonright M_{c, F}\right)$. For each symbol $r \in L^{\mathcal{R}}$, we define $r_{\mathcal{Q}}=\bigcup_{\mathcal{M} \models T} r_{\mathcal{M}} /_{=_{\mathcal{Q}}}$ and $f_{\mathcal{Q}}\left(\overline{[q]_{\mathcal{Q}_{\mathcal{Q}}}}\right)=f_{\mathcal{M}}\left(\overline{[q]_{=_{\mathcal{M}}}}\right)$, if $\bar{q} \in M_{c, F}$, and 0 otherwise. The definitions are correct, and $\mathcal{Q}=\left\langle Q, r_{\mathcal{Q}}, f_{\mathcal{Q}}\right\rangle$ is an $F$-prime-envelope of $T$. " $\Leftarrow$ " is immediate. The second equivalence is easy.

### 1.3 Analysis of lineals and linear theories

In this section, we define the key concepts of this chapter - the notions of a lineal and a linear theory - and formulate our main results concerning them; the proof of the results is postponed to section 1.5. We prove that ZAa and ZLa (defined in section 1.1) are examples of linear theories; in section 1.4, we will use the results of this section to perform a basic model-theoretic analysis of ZAa and ZLa.

### 1.3.1 Lineals and linear theories

As we allready stated in the prologue of this chapter, linear theories are, informatively (up to a change of the language), theories of some (expansions by constants of) discretely ordered modules over certain discretely ordered integral domains. More precisely, $T$ is a linear theory if every model $\mathcal{A} \models T$ is equidefinable with certain expansion of a discretely ordered module (called lineal; see 1.3.1.3) over a domain which is a doded (see 1.3.1.2). Note that, by Corollary 1.3:5, every $\mathcal{A} \models T$ has quantifier elimination in the language of the corresponding lineal.

### 1.3.1.1 Notation

For a structure $\mathcal{M}$ in a language $L=\langle 0,1,+,-, \leq, \ldots\rangle$, we denote ${ }^{+} M$ the set of all non-negative elements from $M$.

For an unary increasing function $f$ on $M, f^{-1}$ denotes the integral inverse of $f$ (if it exists), i.e. the function $f^{-1}$ such that $f^{-1}(x)$ is the largest $y$ with $f(y) \leq x$. For linear $f$, this is equivalent to

$$
\begin{equation*}
\left\langle\mathcal{M}, f, f^{-1}\right\rangle \models 0 \leq x-f\left(f^{-1}(x)\right)<f(1) . \tag{1.6}
\end{equation*}
$$

### 1.3.1.2 Doded

An ordered integral domain $\mathrm{D}=\langle\mathrm{D}, 0,1,+,-, \cdot, \leq\rangle$ is called a doded if it
(R1) is discretely ordered by $\leq$, with 1 being the least positive element, i.e.
$($ R1-a) $\leq$ is a linear ordering of D ,
(R1-b) $r \leq s \rightarrow r+t \leq s+t$,
(R1-c) $0 \leq r, s \rightarrow 0 \leq r \cdot s$,
(R1-d) $0<1 \& \neg(\exists r)(0<r<1)$,
(R2) is regularly quasi-Euclidean, i.e. the Euclidean algorithm in D (with one step given by $\left.(q, r) \mapsto\left(r, q-r r^{-1}(q)\right)\right)$ is correctly defined and always stops in finitely many steps (at some $\left(q^{\prime}, 0\right)$ ),
(R3) has degrees, i.e. there is a function deg : $\mathrm{D} \rightarrow \mathbb{N} \cup\{-\infty\}$ such that rng(deg) is a lower set in $\mathbb{N} \cup\{-\infty\}$, and $\operatorname{deg} r \leq \operatorname{deg} q \Leftrightarrow|r| \leq n|q|$, for some $n \in \mathbb{N}$.

Example 1.3:1.
a) The ordered ring of integers is a doded; the degree map is defined as $\operatorname{deg} z=0$, for $0 \neq z \in \mathbb{Z}$, and $\operatorname{deg} 0=-\infty$.
b) Let $\tau=\left\langle\tau_{p}\right\rangle \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$, where $\mathbb{P}$ is the set of all prime numbers, and $\mathbb{J}_{p}$, for $p \in \mathbb{P}$, is the ring of $p$-adic integers. Let $\mathbb{Z}[a] \subseteq \mathrm{D}_{\tau} \subseteq \mathbb{Q}[a]$ be defined as

$$
\mathrm{D}_{\tau}=\left\{\frac{r}{n} \in \mathbb{Q}[a] ; 0<n \in \mathbb{N}, r \in \mathbb{Z}[a], \text { and }(\forall p \in \mathbb{P}) \pi_{\mathrm{v}_{p}(n)}\left(r\left(\tau_{p}\right)\right)=0\right\} .
$$

Here, $\mathrm{v}_{p}$ denotes the usual $p$-valuation, $\pi_{k}$ the canonical projection of $\mathbb{J}_{p}$ to $\mathbb{Z}_{p^{k}}$. Further, $\tau_{p}$ is the $p$ th projection of $\tau$, and the substitution $r\left(\tau_{p}\right)$ is done inside $\mathbb{J}_{p}$, where $\mathbb{Z}$ is embedded via $z \mapsto\left(z \bmod p, z \bmod p^{2}, z \bmod p^{3}, \ldots\right)$. See section 3.2 for details.
Equivalently:

$$
\left.\frac{r}{n} \in \mathrm{D}_{\tau} \Leftrightarrow n \right\rvert\, r(a)
$$

where $a$ is such that $a \equiv_{p^{k}} \tau_{p}(k)$, for $p \in \mathbb{P}$ and $k \in \mathbb{N}$.
The ring $\mathrm{D}_{\tau}$ is a doded: It is discretely (linearly) ordered since no $q \in \mathbb{Q}[a]$ with $0<q<1$ is in $\mathrm{D}_{\tau}$. The degree map is the usual degree of polynomials. $\mathrm{D}_{\tau}$ is regularly quasi-Euclidean by Theorem 3.4:2,

### 1.3.1.3 Lineal

A lineal is any structure $\mathcal{F}=\left\langle F, 0,1,+,-, \leq, r, c, q^{-1}\right\rangle_{r \in \mathrm{D}_{\mathcal{F}}, c \in \mathrm{C}_{\mathcal{F}}, q \in+\mathrm{D}_{\mathcal{F}}}$ where

- $\mathrm{D}_{\mathcal{F}}$ is an universe of a doded $\mathrm{D}_{\mathcal{F}}$,
- $\mathcal{F}$ and $\mathcal{C}_{\mathcal{F}}=\mathcal{F} \upharpoonright \mathrm{C}_{\mathcal{F}}$ are expansions of discretely ordered $\mathrm{D}_{\mathcal{F}}$-modules (with the least positive element 1) by constants $c$ and integral inverses $q^{-1}$.

The definition above implicitly states that the functions $q^{-1}$, for $q \in{ }^{+} \mathrm{D}_{\mathcal{F}}$, are all correctly defined, and $\mathrm{C}_{\mathcal{F}}$ is closed on all $q^{-1}$. Moreover, since, for given $q \in{ }^{+} \mathrm{D}_{\mathcal{F}}$, the mapping $r \mapsto r 1$ is an embedding of $\left\langle\mathrm{D}_{\mathcal{F}}, 0,1,+,-, \leq, q \cdot{ }_{-}\right\rangle$into $\langle F, 0,1,+,-, \leq, q\rangle$, and $q^{-1}$ is defined in both these structures by the same formula (as in (1.6)), we get the following, for all $q, r \in \mathrm{D}_{\mathcal{F}}, q>0$ :

$$
\mathcal{F} \models\left(q^{-1} r\right) 1=q^{-1}(r 1)
$$

This observation enables us to consider $\left\langle\mathrm{D}_{\mathcal{F}}, 0,1,+,-, \leq, r{ }_{-}, q^{-1}\right\rangle_{r \in \mathrm{D}_{\mathcal{F}}, q \in{ }^{+} \mathrm{D}_{\mathcal{F}}}$ as a substructure of $\left\langle F, 0,1,+,-, \leq, r, q^{-1}\right\rangle_{r \in \mathrm{D}_{\mathcal{F}}, q \in+\mathrm{D}_{\mathcal{F}}}$ and to identify $r \in \mathrm{D}_{\mathcal{F}}$ with $r 1 \in F$.

We extend the degree map deg : $\mathrm{D}_{\mathcal{F}} \rightarrow \mathbb{N} \cup\{-\infty\}$ to the whole $F$ by

$$
\begin{equation*}
\operatorname{deg}(x)=\min \left\{\operatorname{deg}(r) ; r \in \mathrm{D}_{\mathcal{F}},|x| \leq r\right\} \tag{1.7}
\end{equation*}
$$

where $\min (\emptyset)=\infty$.
Example 1.3:2.
a) The structure $\left\langle\mathbb{Z}, 0,1,+,-, \leq, \underline{z}, \mathbf{z}, \underline{n}^{-1}\right\rangle_{z \in \mathbb{Z}, 0<n \in \mathbb{N}}$ is a lineal (the ring of integers is a doded by Example $1.3: 1$ 回)). More generally, for $\mathcal{A} \models \mathrm{ZAa}$, the structure $\mathcal{F}_{\mathcal{A}}=\left\langle A, 0,1,+,-, \leq, \underline{z}, \mathbf{z}, \underline{n}^{-1}\right\rangle_{z \in \mathbb{Z}, 0<n \in \mathbb{N}}$ is a lineal.
b) Let us recall that, for a formula $\varphi$, we denote $\operatorname{cor}(\varphi)$ the formula " $\varphi$ is correct defining formula of a function", and if $s$ is a symbol defined by $\varphi$, we write $\dot{s}$ for $\varphi$ (see section 1.2.3).
Let $\mathcal{A}=\langle A, 0,1,+,-, \underline{a}, \leq\rangle \models$ ZLa. We set

$$
\begin{aligned}
\mathrm{C}_{\mathcal{A}}=\mathrm{D}_{\mathcal{A}} & =\{r \in \mathbb{Q}[a] ; \mathcal{A} \models \operatorname{cor}(\underline{\dot{r}})\}= \\
& =\left\{r \in \mathbb{Q}[a] ; r=\frac{p}{n} \text { with } p \in \mathbb{Z}[a], n \in \mathbb{N} \text { and } \mathcal{A} \models \mathbf{n} \mid \mathbf{p}\right\}= \\
& =\mathbb{Q}[a] \cap A .
\end{aligned}
$$

Clearly, $\mathbb{Z}[a] \subseteq \mathrm{D}_{\mathcal{A}} \subseteq \mathbb{Q}[a]$, and $\mathrm{D}_{\mathcal{A}}$ is closed under operations of the polynomial ring $\mathbb{Q}[a]$; we denote $\mathrm{D}_{\mathcal{A}}$ the ordered subring of the ordered polynomial ring $\mathbb{Q}[a]$ with the universe $\mathrm{D}_{\mathcal{A}}$.
By Proposition 3.3:2 (it is easy to verify that the proof uses only properties of $\mathcal{A}$ provable in ZLa), the rings $\mathrm{D}_{\mathcal{A}}$, for $\mathcal{A} \models$ ZLa, correspond (not uniquely) to the rings $D_{\tau}$ from Example 1.3:1 b) and vice versa.

More precisely,

$$
\mathrm{D}_{\tau}=\mathrm{D}_{\mathcal{A}} \Leftrightarrow \mathcal{A} \models p^{k} \mid \mathbf{a}-\pi_{k}\left(\tau_{p}\right) \text { for all } p \in \mathbb{P}, 0<k \in \mathbb{N} .
$$

The structure $\mathcal{F}_{\mathcal{A}}=\left\langle A, 0,1,+,-, \leq, \underline{r}, \mathbf{p}, \underline{q}^{-1}\right\rangle_{r \in \mathrm{D}_{\mathcal{A}}, p \in \mathrm{C}_{\mathcal{A}}, q \in+{ }^{+}}$is a lineal.
Proof: $\mathrm{D}_{\mathcal{A}}$ is a doded by Example 1.3:1 b). $\mathcal{F}_{\mathcal{A}}$ is an expansion of an ordered $\mathrm{D}_{\mathcal{A}}$-module by Lemma 1.1:2. To show that $\mathrm{C}_{\mathcal{A}}$ is a universe of a substructure of $\mathcal{F}_{\mathcal{A}}$, it suffices to show that it is closed under the functions $q^{-1}$, for all $0<q \in \mathrm{D}_{\mathcal{A}}$; this is proved in Lemma 3.4:1 (the proof can be done in ZLa).
c) Let $\mathcal{A} \models \mathrm{ZAa}^{c}=\mathrm{ZAa} \cup\{c \geq n ; n \in \mathbb{N}\}$. We denote $\mathrm{D}_{\mathcal{A}}$ the ring of integers, and we set

$$
\mathrm{C}_{\mathcal{A}}=\mathbb{Q}\left\langle c^{A}, 1\right\rangle \cap A=\left\{\frac{i c^{A}+j}{l} ; i, j, l \in \mathbb{Z}, l>0, \mathcal{A} \models l \mid i c+j\right\} .
$$

Then the structure $\mathcal{F}_{\mathcal{A}}=\left\langle A, 0,1,+,-, \leq, \underline{z}, k, \underline{n}^{-1}\right\rangle_{z \in \mathbb{Z}, k \in \mathbb{C}_{\mathcal{A}}, 0<n \in \mathbb{N}}$ is a lineal. Indeed, it is enough to prove that $\mathrm{C}_{\mathcal{A}}$ is closed under the functions $\underline{n}^{-1}$, for $0<n \in \mathbb{N}$ : Let $k=\frac{i c+j}{l} \in \mathrm{C}_{\mathcal{A}}$ and $0<n \in \mathbb{N}$. By integral divisibility in ZAa ${ }^{c}$, there is $0 \leq m<n$ such that $n \mid k-m$. Then $\underline{n}^{-1} k=\frac{k-m}{n}=\frac{i c+(j-m l)}{n l} \in \mathrm{C}_{\mathcal{A}}$.

### 1.3.1.4 Linear theory, linealization

Let $L$ be a language extending $L^{z}=\langle 0,1,+,-, \leq\rangle$ (where - is unary), $T$ be an $L$-theory, and let $\mathrm{D}, \mathrm{C} \subseteq F m_{L}$. $\mathrm{A}(\mathrm{D}, \mathrm{C})$-linealization of $T$ is any map $\mathcal{A} \mapsto \mathcal{F}_{\mathcal{A}}$, for $\mathcal{A} \models T$, such that every $\mathcal{F}_{\mathcal{A}}=\left\langle A, 0,1,+,-, \leq, r, c, q^{-1}\right\rangle_{r \in \mathrm{D}_{\mathcal{F}_{\mathcal{A}}}, c \in \mathrm{C}_{\mathcal{F}_{\mathcal{A}}}, q \in{ }^{+} \mathrm{D}_{\mathcal{F}_{\mathcal{A}}}}$ is a lineal equidefinable with $\mathcal{A}$, and the sets $\mathrm{D}_{\mathcal{F}_{\mathcal{A}}},{\dot{\mathrm{C}_{\mathcal{F}_{\mathcal{A}}}}}^{\boldsymbol{D}^{\prime}}$ of definitions in $\mathcal{A}$ of functions from $\mathrm{D}_{\mathcal{F}_{\mathcal{A}}}$ and constants from $\mathrm{C}_{\mathcal{F}_{\mathcal{A}}}$ satisfy $\mathrm{D}_{\mathcal{F}_{\mathcal{A}}} \subseteq \mathrm{D}$ and $\mathrm{C}_{\mathcal{F}_{\mathcal{A}}} \subseteq \mathrm{C}$.

An $L$-theory $T$ is a linear theory if it has an $\left(F m_{L}, F m_{L}\right)$-linealization.

## Example 1.3:3.

a) Let $\dot{\mathbb{Z}}_{1}$ and $\dot{\mathbb{Z}}_{0}$ denote the sets of $L_{\mathbb{Z}}^{\text {add }}$-formulas $\underline{\dot{z}}$ and $\dot{\mathbf{z}}$ which define symbols $\underline{z}$ and $\mathbf{z}$, for $z \in \mathbb{Z}$, respectively. ZAa is a linear theory, and

$$
\mathcal{A} \mapsto \mathcal{F}_{\mathcal{A}}=\left\langle A, 0,1,+,-, \leq, \underline{m}, \mathbf{k}, \underline{n}^{-1}\right\rangle_{m \in \mathbb{Z}, k \in \mathbb{Z}, n \in \mathbb{N}-\{0\}},
$$

for $\mathcal{A} \models \mathrm{ZAa}$, is its $\left(\dot{\mathbb{Z}}_{1}, \dot{\mathbb{Z}}_{0}\right)$-linealization.
b) Let $\dot{\mathbb{Q}}[a]_{1}$ and $\dot{\mathbb{Q}}[a]_{0}$ be the sets of all $L_{\mathbb{Z}}^{\text {lin }}$-formulas $\dot{\underline{r}}$ and $\dot{\mathbf{r}}$ which define symbols $\underline{r}$ or $\mathbf{r}$ respectively, for $r \in \mathbb{Q}[a]$. ZLa is a linear theory, and

$$
\mathcal{A} \mapsto \mathcal{F}_{\mathcal{A}}=\left\langle A, 0,1,+,-, \leq, \underline{r}, \mathbf{p}, \underline{q}^{-1}\right\rangle_{r \in \mathrm{D}_{\mathcal{A}}, p \in \mathrm{C}_{\mathcal{A}}, q \in{ }^{+} \mathrm{D}_{\mathcal{A}}},
$$

for $\mathcal{A}=\mathrm{ZAa}$, is its $\left(\dot{\mathbb{Q}}[a]_{1}, \dot{\mathbb{Q}}[a]_{0}\right)$-linealization.
c) Let $\dot{\mathbb{Q}}\langle c, 1\rangle_{0}$ denotes the set of all formulas $\dot{k}$ which define constants $k \in \mathbb{Q}\langle c, 1\rangle$ (see Example $1.3: 2$ (c)). The theory $\mathrm{ZAa}^{c}$ from Example $1.3: 2$ (c) is a linear theory, and

$$
\mathcal{A} \mapsto \mathcal{F}_{\mathcal{A}}=\left\langle A, 0,1,+,-, \leq, \underline{m}, k, \underline{n}^{-1}\right\rangle_{m \in \mathbb{Z}, k \in \mathrm{C}_{\mathcal{A}}, n \in \mathbb{N}-\{0\}},
$$

for $\mathcal{A} \models \mathrm{ZAa}^{c}$, is its $\left(\dot{\mathbb{Z}}_{1}, \dot{\mathbb{Q}}\langle c, 1\rangle_{0}\right)$-linealization.
We further identify $\mathrm{D}_{\mathcal{F}_{\mathcal{A}}}$ with the set $\dot{\mathrm{D}}_{\mathcal{F}_{\mathcal{A}}} \subseteq F m_{L}$ of $L$-definitions of functions $r \in \mathrm{D}_{\mathcal{F}_{\mathcal{A}}}$ and similarly for $\mathrm{C}_{\mathcal{F}_{\mathcal{A}}}$. When a linealization is fixed, we often write $\mathrm{D}_{\mathcal{A}}$, $\mathrm{C}_{\mathcal{A}}$ instead of $\mathrm{D}_{\mathcal{F}_{\mathcal{A}}}, \mathrm{C}_{\mathcal{F}_{\mathcal{A}}}$.

Also, for a set $D \subseteq F m_{L}$ of definitions of unary functions, by $D^{-1}$ we denote the set of definitions of their integral inverses (see section 1.3.1.1).

### 1.3.2 Main results

At this place, we state our three main theorems on linear theories and lineals (1.3:4, 1.3:6 and $1.3: 8$ ) and their important corollaries. The Main Theorem on Linear Theories 1.3:4 is essential for our descriptive analysis of linear theories, the other two theorems are its refinements, which help us to provide a detailed characterization of definable functions and sets in models of linear theories.

### 1.3.2.1 Solvability and quantifier elimination

The following is a fundamental statement concerning descriptive complexity of (models of) linear theories. The concept of solvability is defined in section 1.2.2, see also section 1.2 .3 for the explanation of the notation $T^{F}$.

Theorem 1.3:4 (Main Theorem on Linear Theories). Let $T$ be a linear theory in a language $L, \mathcal{A} \mapsto \mathcal{F}_{\mathcal{A}}$ be a $(\mathrm{D}, \mathrm{C})$-linealization, and $E=\mathrm{D} \cup \mathrm{C} \cup \mathrm{D}^{-1}$. Then

1) $T^{E}$ is almost uniformly solvable.
2) $T^{\mathrm{C}}$ is 0-solvable.

From Theorem $1.3: 42$ ), it follows that, for every $\mathcal{A} \models T$, the set $\mathrm{C}_{\mathcal{A}}$ is the universe of a substructure of $\mathcal{A}$; we set $\mathcal{C}_{\mathcal{A}}=\mathcal{A} \upharpoonright \mathrm{C}_{\mathcal{A}}$.

## Corollary 1.3:5.

1) $T^{E}$ admits quantifier elimination and is axiomatizable by open formulas.
2) For $\mathcal{A} \models T$, the structure $\mathcal{C}_{\mathcal{A}}$ is the unique prime model of $T h(\mathcal{A})$.
3) For $\mathcal{A} \models T$, the theory $T h(\mathcal{A})$ is equivalent to $T \cup O T h_{L}\left(\mathcal{A}^{\mathrm{C}}\right)$ (or equivalently to $\left.T \cup O T h_{L}\left(\mathcal{C}_{\mathcal{A}}{ }^{\mathrm{C}}\right)\right)$,
where $\operatorname{OTh}_{L}(\mathcal{N})$ is the canonical L-translation of the set of all open sentences true in $\mathcal{N}$.
Proof. 1) follows from Proposition 1.2:6, 2) and 3) from Corollary 1.2:12.

### 1.3.2.2 Harmonic forms

We prove that every term or formula can be, in a given lineal, equivalently written in "harmonic form", i.e. as composed solely of linear combinations of expressions of the form $r^{-1} x$, where $x$ is a variable. Let, further, $\mathcal{F}$ be a fixed lineal. We write D and C instead of $\mathrm{D}_{\mathcal{F}}$ and $\mathrm{C}_{\mathcal{F}}$.
1.3.2.2.1 Harmonic term We say that a term $t(\bar{x})$ is harmonic (or equivalently in harmonic form) if

$$
t(\bar{x})=\sum_{i=0}^{N-1} \underline{q}_{i} \underline{r}^{-1}\left(x_{f(i)}\right)+\underline{c},
$$

for some $q_{i}, r_{i} \in \mathrm{D}, c \in \mathrm{C}$ and $f: N \rightarrow l(\bar{x})$. A formula or a p-term (see 1.2.2.4) is harmonic if all its maximal subterms are.
1.3.2.2.2 Almost-term The following special case of a p-term is worth to be named. A p-term $\tau$ is called an almost-term if it is of the form

$$
\tau(\bar{x})=\left\{s(\bar{x})+c_{i} \text { if } \psi_{i}(\bar{x}), i<n\right.
$$

where $s(\bar{x})$ is a term, and $c_{i} \in \mathrm{C}$, for $i<n$. We write $\operatorname{core}(\tau)$ for $s$ and $\operatorname{cond}(\tau)$ for the set of all "conditions" $\psi_{i}, i<n$.

Theorem 1.3:6 (Harmonic Form Theorem). Let $\mathcal{F}$ be a lineal.

1) For every term $t(\bar{x})$, there is an open harmonic almost-term $\tau(\bar{x})$ such that $\mathcal{F} \models t(\bar{x})=\tau(\bar{x})$.
2) For every formula $\varphi(\bar{x})$, there is an open harmonic formula $\psi(\bar{x})$ such that $\mathcal{F} \models \varphi(\bar{x}) \leftrightarrow \psi(\bar{x})$.

Remark 1.3:7. Our proof of Theorem 1.3:6, in fact, proves more than stated - the equivalent harmonic forms can be found not only for each lineal $\mathcal{F}$ separately but at once for a given linear theory. However, proving this explicitely would cause that all the statements and subproofs of our proof would be recognizably longer and more complicated.

### 1.3.2.3 Bases and definable functions and sets in lineals

The machinery, we are going to develop for the proof of the Main Theorem 1.3:4, enables us to perform a detailed analysis of definable functions and sets in a lineal $\mathcal{F}$. In particular, we prove that every definable set $D \subseteq F^{n}$ is a union of linear images of polyhedra in $F^{m}$, for some $m \in \mathbb{N}$ (see Corollary 1.3:9]2)).
1.3.2.3.1 Divisor Let $\alpha$ be a formula or a p-term in harmonic form. A scalar $r \in \mathrm{D}$ is called an $x$-divisor [divisor] of $\alpha$ if $\alpha$ contains a subterm of the form $r^{-1} x$ [for some variable $x$ ]. The set of all $x$-divisors [divisors] in $\alpha$ is denoted $\operatorname{Div}_{x}(\alpha)$ $[\operatorname{Div}(\alpha)] . \alpha$ is said to be over a set $S \subseteq \mathrm{D}[\operatorname{in} x]$ if $\operatorname{Div}(\alpha) \subseteq S\left[\operatorname{Div}_{x}(\alpha) \subseteq S\right]$. If $\operatorname{Div}_{[x]}(\alpha)=\emptyset$, we call $\alpha$ linear $[$ in $x]$.
1.3.2.3.2 Basis Let $d \leq e \leq \omega$ and $B \subseteq{ }^{+} \mathrm{D}$. We say that $B$ is a $[d, e]$-basis if there is an enumeration $B=\left\langle b_{i}\right\rangle_{d \leq i<e}$ such that $\operatorname{deg} b_{i}=i$. [0, $\omega$ ]-basis is often called just basis.

Theorem 1.3:8 (Bases Theorem). Let $\bar{\delta} \in F^{n}, C_{p}(\bar{\delta})=\prod_{i<n}\left[\delta_{i}, \delta_{i}+p-1\right] \subseteq F^{n}$ be a cube with edges of scalar length $p \in \mathrm{D}, e=\operatorname{deg}(p)$, and $B$ be $a[0, e]$-basis. Let $l(\bar{x})=n$. Then the following holds:

1) Every term $t(\bar{x})$ is on $C_{p}(\bar{\delta})$ equal to a harmonic almost-term $\tau(\bar{x})$ which is over $m B$ for some $m \in \mathbb{N}$.
2) Every formula $\varphi(\bar{x})$ is on $C_{p}(\bar{\delta})$ equivalent to an open harmonic formula $\psi(\bar{x})$ which is over $m B$ for some $m \in \mathbb{N}$.

Moreover:

- $m$ can be chosen as any number sufficiently large with respect to divisibility.
- If $t$ or $\varphi$ contain parameters from a set $X$ then $\tau$ and $\psi$ contain only parameters from $X \cup \bar{\delta}$.
1.3.2.3.3 Box, polyhedron For $0<\bar{a} \in F^{n}$, we denote

$$
K(\bar{a})=\prod_{i=0}^{n-1}\left[0, a_{i}-1\right]
$$

the $n$-dimensional box with edges $\bar{a}$. Let $Y \subseteq F^{n}$ and $\bar{\beta} \in F$. A set $P \subseteq Y \subseteq F^{n}$ is called a polyhedron in $Y$ over parameters from $X \subseteq F$ if $P$ is the set of all solutions $\bar{x} \in Y$ of a system of inequalities of the form $L(\bar{x}) \leq s(\bar{\beta})$, where $s$ is a term, $\bar{\beta} \in X$ are parameters, and $L$ is a linear form (in $\mathcal{F}$, i.e. with coefficients from D$)$.
1.3.2.3.4 Linear coordination of $F^{n}$ Let $\bar{\delta}=\left(\delta_{0}, \ldots, \delta_{n-1}\right) \in F^{n}, m=$ $a_{0} \in \mathbb{N}$, and let $\bar{a}=\left(a_{0}, a_{1}, \ldots, a_{N}\right)$ and $p$ be scalars such that $\operatorname{deg}\left(a_{i}\right)=1$, for $1 \leq i \leq N, \operatorname{deg}(p)=N$, and $m b_{N} \geq p$, where $b_{0}=1, b_{j}=\prod_{1 \leq i \leq j} a_{i}$ (then $\left\langle b_{j}\right\rangle_{0 \leq j \leq N}$ is a $[0, N+1]$-basis).

We define $g^{\prime}: K(\bar{a}) \times F \rightarrow F$ as

$$
\begin{equation*}
g^{\prime}(\bar{\alpha}, u)=\alpha_{0}+\sum_{j=0}^{N-1} m b_{j} \alpha_{j+1}+p u \tag{1.8}
\end{equation*}
$$

and $g:(K(\bar{a}) \times F)^{n} \rightarrow F^{n}$ as $g=\left(g^{\prime}+\delta_{0}, \ldots, g^{\prime}+\delta_{n-1}\right)$, i.e.

$$
\begin{equation*}
g\left(\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n}\right)=\left(g^{\prime}\left(\overline{\alpha_{i}}, u_{i}\right)+\delta_{i}\right)_{i<n} \tag{1.9}
\end{equation*}
$$

We call such a $g$ the (linear) $(\bar{\delta}, \bar{a}, p)$-coordination of $F^{n}$.
Obviously, $g$ is surjective, thanks to $m b_{N} \geq p$. If $m b_{N}=p$ then $g$ is a bijection.
We set

$$
P_{g}=\left\{\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n} ; \bigwedge_{i<n} \alpha_{i, 0}+\sum_{j=0}^{N-1} m b_{j} \alpha_{i, j+1}<p\right\} .
$$

It is easy to see that $g \upharpoonright P_{g}: P_{g} \rightarrow F^{n}$ is a bijection.
An important property of a coordination is that, for $x=g^{\prime}(\bar{\alpha}, 0),\left(m b_{i}\right)^{-1} x$ is a linear combination of $\bar{\alpha}$ :

$$
\begin{equation*}
\left(m b_{i}\right)^{-1} x=\sum_{j=i}^{N-1} \frac{b_{j}}{b_{i}} \alpha_{j+1} \tag{1.10}
\end{equation*}
$$

where $b_{i} \mid b_{j}$, for $i \leq j$.
Corollary 1.3:9. Let $\mathcal{F}$ be a lineal, $\bar{\delta}=\left(\delta_{0}, \ldots, \delta_{n-1}\right) \in F^{n},\left\langle a_{i}\right\rangle_{i=1}^{\infty}$ be scalars with $\operatorname{deg}\left(a_{i}\right)=1$, for all $i$, and let $X \subseteq F$ be a set of parameters. Then the following holds:

1) Every $X$-definable (in $\mathcal{F}$ ) function $f: F^{n} \rightarrow F$ is given by a formula

$$
f \circ g=\lambda \text { on } P_{g}
$$

i.e. $f\left(g\left(\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n}\right)\right)=\lambda\left(\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n}\right)$ for $\left(\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n}\right) \in P_{g}$, where the map $g:(K(\bar{a}) \times F)^{n} \rightarrow F^{n}$ is a $(\bar{\delta}, \bar{a}, p)$-coordination of $F^{n}$, with $a_{0}=m \in \mathbb{N}$, $\bar{a}=\left(a_{0}, \ldots, a_{N}\right)$ and $p$ a scalar, and $\lambda\left(\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n}\right)$ is a linear $p$-term over parameters from $X \cup \bar{\delta}$.
In a particular case when $f$ is a term, $\lambda$ can be chosen as a linear almost-term.
2) Every set $D \subseteq F^{n} X$-definable in $\mathcal{F}$ can be written as

$$
D=\bigcup_{i<k} g\left[P_{i}\right]
$$

where $g:(K(\bar{a}) \times F)^{n} \rightarrow F^{n}$ is a $(\bar{\delta}, \bar{a}, p)$-coordination of $F^{n}$, with $a_{0}=m \in \mathbb{N}$, $\bar{a}=\left(a_{0}, \ldots, a_{N}\right)$ and $p$ a scalar, and $P_{i} \subseteq P_{g}$, for $i<k$, are finitely many polyhedra in $(K(\bar{a}) \times F)^{n}$ over parameters from $X \cup \bar{\delta}$.

Moreover, $a_{0}=m$ and $p$ may be chosen as any elements sufficiently large with respect to divisibility (the choice of $m$ depends on $p$ ).

Proof. 1): By Proposition 1.2:9, $f$ is the realization of a p-term $\tau$ with parameters from $X$. Therefore it is enough to prove the statement for $f=t(\bar{x})$ where $t$ is a term (with parameters from $X$ ). We may also suppose $\bar{\delta}=\overline{0}$ (the general result may be then obtained by setting $t^{\prime}(\bar{y})=t(\bar{y}+\bar{\delta})$ ).

Let $t(\bar{x})$ be a term and $\bar{\delta}=\overline{0}$. There is a scalar $p$ which is a linear period of $t$, i.e. there are scalars $\bar{\gamma}$ such that $t(\bar{x}+p \bar{u})=t(\bar{x})+\overline{\gamma u}$ holds for all $\bar{x}, \bar{u}$. We set $N=\operatorname{deg}(p), b_{0}=1, b_{j}=\prod_{1 \leq i \leq j} a_{i}$ (then $\left\langle b_{j}\right\rangle_{0 \leq j \leq N}$ is a $[0, N+1]$-basis) and take $a_{0}=m \in \mathbb{N}$ such that $m b_{N} \geq p$ and such that there is an almost-term $\tau(\bar{x})$ over $\left\langle m b_{j}\right\rangle_{j<N}$ with $\tau(\bar{x})=t(\bar{x})$, for $0 \leq \bar{x}<p$ (this can be achieved by Theorem 1.3:8, $\tau$ is with parameters from $X$ ). Let $g$ be the ( $\overline{0}, \bar{a}, p)$-coordination of $F^{n}$.

Every $\bar{x} \in F^{n}$ can be uniquely written in the form $\bar{x}=g\left(\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n}\right)$, with $\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n} \in P_{g}$. Set $h\left(\overline{\alpha_{i}}\right)=\alpha_{i, 0}+\sum_{j=0}^{N-1} m b_{j} \alpha_{i, j+1}$. Then it is

$$
t\left(g\left(\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n}\right)\right)=t\left(\left(h\left(\overline{\alpha_{i}}\right)\right)_{i<n}\right)+\overline{\gamma u}=\tau\left(\left(h\left(\overline{\alpha_{i}}\right)\right)_{i<n}\right)+\overline{\gamma u} .
$$

By (1.10), it is $\tau\left(\left(h\left(\overline{\alpha_{i}}\right)\right)_{i<n}\right)+\overline{\gamma u}=\lambda\left(\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n}\right)$, for some linear almostterm $\lambda$.

The "moreover" part of the statement is clear from our choice of $p$ and $m$.
2): Let $D$ be defined by a formula $\varphi(\bar{x})$ with parameters from $X$. By Corollary 1.3:5 (1), we may suppose that $\varphi$ is open.

For every atomic subformula $t(\bar{x}) \leq 0$ of $\varphi$, let $\lambda_{t}$ be a linear almost-term and $g_{t}$ a $(\delta, \bar{a}, p)$-coordination such that $t \circ g_{t}=\lambda_{t}$ (this is possible by 1$)$ ). By the "moreover" part of the statement, we may suppose that all $g_{t}$ are mutually equal and denote them just $g$. Every $\bar{x} \in F^{n}$ can be uniquely written in the form $\bar{x}=$ $g\left(\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n}\right)$, with $\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n} \in P_{g}$. Then $t(\bar{x}) \leq 0 \Leftrightarrow \lambda_{t}\left(\left(\overline{\alpha_{i}}, u_{i}\right)_{i<n}\right) \leq 0$. The last inequality defines a finite union of polyhedra in $(K(\bar{a} \times F))^{n}$ with parameters $X \cup \bar{\delta}$. This proves the theorem.

Remark 1.3:10. The statement of the Corollary 1.3:9 2) can be understood as stating that the Boolean algebra of definable sets (over parameters $X$ ) in $\mathcal{F}$ is isomorphic to the algebra generated by polyhedra over $X$ in $K(\bar{a})$.

The similar statement for the Lindenbaum algebra of a linear theory can be proven as well. However, we do not do that for the same reasons which we already explained in the Remark 1.3:7,

Remark 1.3:11. Let us note that, alternatively, linear theories may be defined as two sorted ("ordered ring-ordered module") theories. In that case the proof of the Main Theorem on Linear Theories 1.3:4 yields a quantifier elimination statement for ordered modules with scalar variables - see section 1.6 for details. This problem has been studied in vdDH92 for unordered modules and in Wei97 for discretely ordered modules over the ring $\mathbb{Z}$ of integers (more precisely for the two-sorted variant of Presburger arithmetic).

### 1.4 Application of the main results

We use the Main Theorem on Linear Theories $1.3: 4$ to examine basic modeltheoretic properties of linear theories ZAa and ZLa (and consequently of their $\mathbb{N}$-like versions Aa and La).

The results are stated in Propositions 1.4:1 and Theorem 1.4:5, respectively, and in their corollaries.

### 1.4.1 Properties of ZAa and Aa

The theory ZAa (which is only a $\mathbb{Z}$-like version of Presburger arithmetic Pr) is well-explored. In this section, we mostly reprove long-known results. We do that in order to show the possibilities of our method and to give the reader an opportunity to become more familiar with concepts we defined. If the reader feels confident in understanding our previous definitions, he or she may safely skip this section.

In Example $1.3: 3$ 回), we showed that ZAa is a linear theory and we described its linealization. Let us remind that $\dot{\mathbb{Z}}_{1}$ and $\dot{\mathbb{Z}}_{0}$ stand for the sets of $L_{\mathbb{Z}}^{\text {add }}$-formulas $\underline{\dot{z}}$ and $\dot{\mathbf{z}}$ which define symbols $\underline{z}$ and $\mathbf{z}$, for $z \in \mathbb{Z}$, respectively. We denote $\dot{\dot{Z}}=\dot{\mathbb{Z}}_{1} \cup \dot{\mathbb{Z}}_{0} \cup+\dot{\mathbb{Z}}_{1}^{-1}$, where ${ }^{+} \dot{\mathbb{Z}}_{1}^{-1}$ denotes the set of definitions of integral inverses of positive scalars $\underline{z}$. It is easy to see that all formulas from $\dot{\mathbb{Z}}$ are in ZAa equivalent to existential ones.

The Main Theorem on Linear theories $1.3: 4$ and the results from section 1.2 make it easy to prove the following properties of the theory ZAa:

Proposition 1.4:1 (Properties of ZAa).

1) $\mathrm{ZAa}^{\dot{\mathbb{Z}}}$ is almost uniformly solvable, $\mathrm{ZAa}^{\dot{Z}_{0}}$ is 0 -solvable.

Hence: ZAa is model-complete.
Moreover: Every formula is in ZAa equivalent to a disjunction of primitive positive formulas, i.e. to a formula of the form $\bigvee_{i<n}(\exists \bar{z}) \psi_{i}$, where each $\psi_{i}$ is a system of linear inequalities.
2) ZAa is decidable, complete, and $\langle\mathbb{Z}, 0,1,+,-, \leq\rangle$ is its prime model.
3) Theories $\mathrm{ZAa}, \mathrm{ZAA}$ and $\operatorname{Th}(\langle\mathbb{Z}, 0,1,+,-, \leq\rangle)$ are equivalent.

Proof. 1) is a corollary of Theorem $1.3: 4$ and the fact that each formula from $\dot{\mathbb{Z}}$ is equivalent to a disjunction of primitive positive formulas.
$2):\langle\mathbb{Z}, 0,1,+,-, \leq\rangle$ is the prime model of every $\operatorname{Th}(\mathcal{A})$ with $\mathcal{A} \models$ ZAa, by Corollary $1.3: 5$ 2), hence it is the prime model of ZAa. Completeness and decidability are immediate consequences.
$3)$ : Clearly, ZAA and $T h(\langle\mathbb{Z}, 0,1,+,-, \leq\rangle)$ are extensions of ZAa; the statement then follows from 2).

Let $\dot{\mathbb{N}}_{1}$ and $\dot{\mathbb{N}}_{0}$ be the sets of $L^{\text {add }}$-formulas $\underline{\dot{n}}$ and $\dot{\mathbf{n}}$ which define symbols $\underline{n}$ and $\mathbf{n}$, for $n \in \mathbb{N}$, respectively. We denote $\dot{\mathbb{N}}=\dot{\mathbb{N}}_{1} \cup \dot{\mathbb{N}}_{0} \cup{ }^{+} \dot{\mathbb{N}}_{1}^{-1}$, where ${ }^{+} \dot{\mathbb{N}}_{1}^{-1}$ denotes the set of definitions of integral inverses of positive scalars $\underline{n}$.

Corollary 1.4:2 (Properties of Aa).

1) $\mathrm{Aa}^{\dot{\mathbb{N}}}$ is almost uniformly solvable, $\mathrm{Aa}^{\dot{\mathbb{N}}_{0}}$ is 0 -solvable.

Hence: Aa is model-complete.
Moreover: Every formula is in Aa equivalent to a disjunction of primitive positive formulas, i.e. to a formula of the form $\bigvee_{i<n}(\exists \bar{z}) \psi_{i}$, where each $\psi_{i}$ is a system of linear inequalities.
2) Aa is decidable, complete, and $\langle\mathbb{N}, 0,1,+, \leq\rangle$ is its prime model.
3) Theories Aa, AA and $\operatorname{Th}(\langle\mathbb{N}, 0,1,+, \leq\rangle)$ are equivalent.

Proof. The statements follow easily from Proposition 1.4:1, by using relations (1.1) and (1.2) from 1.1.7.

### 1.4.2 Properties of ZLa and La

We state here basic properties of ZLa - we show its elimination set of formulas, describe its simple complete extensions including their prime models and prove its decidability (see Theorem 1.4:5). Moreover, we provide a characterization of models of ZLa as non-principal ultraproducts of definable expansions of the standard model $\langle\mathbb{Z}, 0,1,+, \leq\rangle$ (see Corollary 1.4:6). As a corollary, we get similar results also for La (Corollaries 1.4:7 and 1.4:8).

The results can be interpreted as stating that LA is model-theoretically very similar to $\operatorname{Pr}$ and far away from P (although the proof of the properties for LA is much more difficult than the same for $\operatorname{Pr}$; we will discuss that in section (1.5). Whether this is true also for $\mathrm{LA}_{\kappa}$ with $\kappa \geq 2$, is posed as the Open question 1 .

Let us remind that $\dot{\mathbb{Q}}[a]_{1}$ and $\dot{\mathbb{Q}}[a]_{0}$ are the sets of all $L_{\mathbb{Z}}^{l i n}$-formulas $\underline{\dot{r}}$ and $\dot{\mathbf{r}}$ which define symbols $\underline{r}$ or $\mathbf{r}$ respectively, for $r \in \mathbb{Q}[a]$. The following lemma states that ZLa proves the "full" scheme of integral-divisibility (see 1.1.2 for definition).

Lemma 1.4:3. ZLa $\vdash \operatorname{cor}(\underline{\underline{r}}) \rightarrow i d(\underline{r})$, for $0<r \in \mathbb{Q}[a]$.
Proof. Let $r=\frac{p}{n}$, with $p \in \mathbb{Z}[a], 0<n \in \mathbb{N}$. Further, we work in a fixed $\mathcal{A} \equiv$ ZLa.

Suppose that $\mathcal{A} \models \operatorname{cor}(\underline{\underline{r}})$, i.e. $\mathbf{n} \mid \mathbf{p}$ in $\mathcal{A}$. By $\operatorname{id}(\underline{p})$, there is $y$ such that $0 \leq \underline{n} x-\underline{p} y<\underline{p} 1 ;$ then $0 \leq x-\underline{r} y<\underline{r} 1$.

In Example 1.3:3 b), we showed that ZLa is a linear theory, and we described its linealization. In order to formulate the consequences of the Main Theorem on Linear Theories $1.3: 4$ for ZLa, we need to introduce some notation.

### 1.4.2.1 Extensions ZLa $_{\tau}$, structures $\mathcal{C}_{\tau}$

Motivated by the Example $1.3: 2$ b), we define $\mathrm{ZLa}_{\tau}$, for $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$, to be the extension of ZLa by axioms expressing $\mathbf{p}^{\mathbf{k}} \mid \mathbf{a}-\pi_{k}\left(\tau_{p}\right)$, for all $p \in \mathbb{P}, 0<k \in \mathbb{N}$.

We also set $\mathcal{C}_{\tau}=\left\langle\mathrm{D}_{\tau}, 0,1,+,-, \leq, \underline{a}\right\rangle$, where $\underline{a}$ is the unary function of multiplication by the variable $a$ (i.e. $\mathcal{C}_{\tau}$ is an $L_{\mathbb{Z}}^{\text {lin }}$-structure, which is a restriction of $\mathrm{D}_{\tau}$ as a module over itself).

### 1.4.2.2 Syntactic presentation of $\mathrm{D}_{\mathcal{A}}$

For $\mathcal{A} \models$ ZLa, we identify the set $\mathrm{D}_{\mathcal{A}}\left(\right.$ see Example $1.3: 2[\mathrm{~b})$ ) with the set $A_{*, \dot{\mathbb{Q}}[a]_{0}}$ of all equivalence-classes of correct constant terms of $\mathcal{A}^{\dot{\mathbb{Q}}[a]_{0}}\left(A_{*, \dot{\mathbb{Q}}[a]_{0}}\right.$ is the universe of the canonical structure of $\operatorname{Th}(\mathcal{A})^{\dot{\mathbb{Q}}[a]_{0}}$; see section 1.2 .4 for details). This is possible since $\mathcal{A} \models \mathbf{r} \neq \mathbf{r}^{\prime}$, for two different elements $r, r^{\prime} \in \mathrm{D}_{\mathcal{A}}$.

Lemma 1.4:4. The naturally defined $L^{\dot{\mathbb{Q}}[a]_{0}}$-structure $\mathcal{Q}$ with the universe $\mathbb{Q}[a]$ is a $\dot{\mathbb{Q}}[a]_{0}$-prime-envelope of ZLa. Therefore, ZLa is $\dot{\mathbb{Q}}[a]_{0}$-compatible and $\dot{\mathbb{Q}}[a]_{0^{\prime}}$ cuc.

Proof. It is $A_{* \dot{\mathbb{Q}}[a]_{0}}=\mathrm{D}_{\mathcal{A}} \subseteq \mathbb{Q}[a]$, for every $\mathcal{A} \models$ ZLa. Hence, $\mathcal{Q}$ is a $\dot{\mathbb{Q}}[a]_{0}$-primeenvelope of ZLa. The rest of the statement follows from Proposition 1.2:16 and Observation 1.2:14.

### 1.4.2.3 Properties of ZLa

Denote $\dot{\mathbb{Q}}[a]=\dot{\mathbb{Q}}[a]_{1} \cup \dot{\mathbb{Q}}[a]_{0} \cup^{+} \dot{\mathbb{Q}}[a]_{1}^{-1}$, where ${ }^{+} \dot{\mathbb{Q}}[a]_{1}^{-1}$ denotes the set of definitions of integral inverses of positive scalars $\underline{q}$. It is easy to see that all formulas from $\dot{\mathbb{Q}}[a]$ are in ZLa equivalent to existential formulas.

Theorem 1.4:5 (Properties of ZLa).

1) $\mathrm{ZLa}^{\dot{\mathbb{Q}}[a]}$ is almost uniformly solvable, $\mathrm{ZLa}^{\dot{\mathbb{Q}}[a]_{0}}$ is 0 -solvable.

Hence: ZLa is model-complete.
Moreover: Every formula is in ZLa equivalent to a disjunction of primitive positive formulas, i.e. to a formula of the form $\bigvee_{i<n}(\exists \bar{z}) \psi_{i}$, where each $\psi_{i}$ is a system of linear inequalities.
2) $\mathrm{ZLa}_{\tau}$, for $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$, are all simple complete extensions of ZLa . For $\mathcal{A}, \mathcal{B} \models \mathrm{ZLa}$, it is $\mathcal{A} \equiv \mathcal{B} \Leftrightarrow \mathbf{a}^{\mathcal{A}} \equiv \mathbf{a}^{\mathcal{B}} \bmod n$, for all $0<n \in \mathbb{N}$.
3) $\mathcal{C}_{\tau}$ is the unique prime model of $\mathrm{ZLa}_{\tau}$, for $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$.
4) ZLa is decidable.
$\mathrm{ZLa}_{\tau}$ is decidable if and only if $\tau$ is recursive.
5) Theories ZLa and ZLA are equivalent.

Proof. 1) is a corollary of Theorem 1.3:4 and the fact that each formula from $\dot{\mathbb{Q}}[a]$ is equivalent to a disjunction of primitive positive formulas.
2): By Example $1.3: 2$ b), every model $\mathcal{A} \models \mathrm{ZLa}$ is a model of some $\mathrm{ZLa}_{\tau}$, with $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$. ZLa is $\dot{\mathbb{Q}}[a]_{0}$-cuc, by Lemma 1.4:4. Now, by Proposition 1.2:16, it is enough to prove that for $\mathcal{A}, \mathcal{B} \models \mathrm{ZLa}_{\tau}$ it is $\mathrm{D}_{\mathcal{A}}=\mathrm{D}_{\mathcal{B}}$. This is true since $\left.\frac{r}{n} \in \mathrm{D}_{\mathcal{A}} \Leftrightarrow \mathcal{A} \models \mathbf{n} \right\rvert\, \mathbf{r}$, and the last is decided by the new axioms of $\mathrm{ZLa}_{\tau}$. The characterization of models up to elementary equivalence is an immediate consequence.
3) follows from 2) and Corollary $1.3: 5$ (2).
4): The set $\left\{\bigwedge_{p_{0}>p \in \mathbb{P}, k<k_{0}} \mathbf{a} \equiv_{p^{k}} \tau(p, k) ; p_{0} \in \mathbb{P}, k_{0} \in \mathbb{N}, \tau:\left(\mathbb{P} \cap p_{0}\right) \times k_{0} \rightarrow \mathbb{N}\right.$ such that $\tau\left(p, k^{\prime}\right) \equiv_{p^{k}} \tau(p, k)<p^{k}$, for all $\left.k \leq k^{\prime}<k_{0}, p_{0}>p \in \mathbb{P}\right\}$ is dense in $C S(\mathrm{ZLa})$, and it is easy to verify that it is $\Sigma_{1}$. Therefore, ZLa is $\Sigma_{1}$-separable and hence decidable, by Proposition 1.2:1,
$\mathrm{ZLa}_{\tau}$ is decidable if and only if it is recursively axiomatizable.
5): For each $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$, the theory $\mathrm{ZLA}_{\tau}$ is a simple extension of $\mathrm{ZLa}_{\tau}$, and therefore these theories are equivalent, by 2). Then ZLA and ZLa have the same simple complete extensions, thus are equivalent.

Corollary 1.4:6. Up to elementary equivalence, models of ZLa are exactly all ultraproducts

$$
\mathcal{Z}_{\mathcal{U}}=\left(\prod_{n \in \mathbb{N}}\langle\mathbb{Z}, 0,1,+,-, \underline{n}, \leq\rangle\right) / \mathcal{U}
$$

where $\mathcal{U}$ is a non-principal ultrafilter on $\mathbb{N}$, i.e. $\mathcal{U} \in \beta \mathbb{N}-\mathbb{N}$.
Proof. 1) Let $\mathcal{U} \in \beta \mathbb{N}-\mathbb{N}$. We show that $\mathcal{Z}_{\mathcal{U}} \models$ ZLa.
All axioms of ZLa, except the axioms $\underline{a} 1 \neq m$, are true in all structures $\langle\mathbb{Z}, 0,1,+,-, \underline{n}, \leq\rangle$. The axiom $\underline{a} 1 \neq m$ holds in all $\langle\mathbb{Z}, 0,1,+,-, \underline{n}, \leq\rangle$ with $n>m$, and $\{n ; n>m\} \in \mathcal{U}$ since $\mathcal{U}$ is non-principal. Therefore, $\mathcal{Z}_{\mathcal{U}}=$ ZLa.
2) Let $\mathrm{ZLa}_{\tau}$, with $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$, be a simple complete extension of ZLa. We find $\mathcal{U} \in \beta \mathbb{N}-\mathbb{N}$ such that $\mathcal{Z}_{\mathcal{U}} \models$ ZLa $_{\tau}$.

Let $\mathcal{S}_{\tau}=\{[m, \infty) ; m \in \mathbb{N}\} \cup\left\{\pi_{m}\left(\tau_{p}\right)+p^{m} \cdot \mathbb{N} ; p \in \mathbb{P}, 0<m \in \mathbb{N}\right\}$. By the Chinese Remainder Theorem, finite intersections of elements from $\mathcal{S}$ are nonempty, hence there is an ultrafilter $\mathcal{U} \supseteq \mathcal{S}$. Clearly, $\mathcal{U}$ is non-principal and $\mathcal{Z}_{\mathcal{U}} \models \mathrm{ZLa}_{\tau}$.

### 1.4.2.4 Properties of La

Let ${ }^{+} \dot{\mathbb{Q}}[a]_{1}$ and ${ }^{+} \dot{\mathbb{Q}}[a]_{0}$ be the sets of all $L^{l i n}$-formulas $\underline{\dot{r}}$ and $\dot{\mathbf{r}}$ which define symbols $\underline{r}$ or $\mathbf{r}$ respectively, for $0 \leq r \in \mathbb{Q}[a]$. Denote ${ }^{+} \dot{\mathbb{Q}}[a]={ }^{+} \dot{\mathbb{Q}}[a]_{1} \cup^{+} \dot{\mathbb{Q}}[a]_{0} \cup^{+} \dot{\mathbb{Q}}[a]_{1}{ }^{-1}$, where ${ }^{+} \dot{\mathbb{Q}}[a]_{1}{ }^{-1}$ denotes the set of definitions of integral inverses of positive scalars $\underline{q}$.

We define $\mathrm{La}_{\tau}$, for $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$, to be the extension of La by axioms expressing $\mathbf{p}^{\mathbf{k}} \mid \mathbf{a}-\pi_{k}\left(\tau_{p}\right)$, for all $p \in \mathbb{P}, 0<k \in \mathbb{N}$. We also set $\mathcal{C}_{\tau}^{+}=\left\langle{ }^{+} \mathrm{D}_{\tau}, 0,1,+,-, \leq, \underline{a}\right\rangle$, where $\underline{a}$ is the unary function of multiplication by the variable $a$.

Corollary 1.4:7 (Properties of La).

1) $\mathrm{La}^{+\dot{\mathbb{Q}}[a]}$ is almost uniformly solvable, $\mathrm{La}^{+\dot{\mathbb{Q}}[a]_{0}}$ is 0 -solvable.

Hence: La is model-complete.
Moreover: Every formula is in La equivalent to a disjunction of primitive positive formulas, i.e. to a formula of the form $\bigvee_{i<n}(\exists \bar{z}) \psi_{i}$, where each $\psi_{i}$ is a system of linear inequalities.
2) $\mathrm{La}_{\tau}$, for $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$, are all simple complete extensions of La .

For $\mathcal{A}, \mathcal{B} \models \mathrm{La}$, it is $\mathcal{A} \equiv \mathcal{B} \Leftrightarrow \mathbf{a}^{\mathcal{A}} \equiv \mathbf{a}^{\mathcal{B}} \bmod n$, for all $0<n \in \mathbb{N}$.
3) $\mathcal{C}_{\tau}^{+}$is the unique prime model of $\mathrm{La}_{\tau}$, for $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$.
4) La is decidable.
$\mathrm{La}_{\tau}$ is decidable if and only if $\tau$ is recursive.
5) Theories La and LA are equivalent.

Proof. Follows easily from Theorem 1.4:5, by using relations (1.1) and (1.2) from 1.1.7.

Corollary 1.4:8. Up to elementary equivalence, models of La are exactly all ultraproducts

$$
\mathcal{N}_{\mathcal{U}}=\left(\prod_{n \in \mathbb{N}}\langle\mathbb{N}, 0,1,+,-, \underline{n}, \leq\rangle\right) / \mathcal{U}
$$

where $\mathcal{U}$ is a non-principal ultrafilter on $\mathbb{N}$, i.e. $\mathcal{U} \in \beta \mathbb{N}-\mathbb{N}$.
Proof. Similarly, as for Corollary 1.4:2,
As we have already noted, the theories $\mathrm{LA}^{\kappa}$, with $\kappa$ cardinal, form an ascending chain of theories between Pr and P . We have also remarked that Corollary 1.4:7 can be understood as stating that $\mathrm{LA}=\mathrm{LA}^{1}$ is model-theoretically similar to $\operatorname{Pr}=L^{0}$ and different from P . In particular, no model of P is definable in a model of LA. Therefore, it is natural to ask the following:

## Open question 1.

a) Are model-theoretical properties of $\mathrm{LA}^{\kappa}$, with $\kappa \geq 2$, still similar to those of $\operatorname{Pr}$ ? In particular, are theories $\mathrm{LA}^{\kappa}$, with $\kappa \geq 2$, model-complete and decidable?
b) Could be some model of P definable in a model of $\mathrm{LA}^{\kappa}$ ?

### 1.5 Proofs

In this section, we prove Theorems 1.3:4, 1.3:6 and 1.3:8, For reader's convenience, we sketch the key steps of the proofs here.

### 1.5.1 Proof prologue

In section 1.5.2, we derive the theorems from three crucial propositions, denoted $\mathrm{S}, \mathrm{H}$ and B . All these propositions are statements considering a single lineal. Therefore we fix a lineal $\mathcal{F}$ with universe $F$ and denote the sets $\mathrm{D}_{\mathcal{F}}$ and $\mathrm{C}_{\mathcal{F}}$ of scalars and constants of $\mathcal{F}$ shortly as D and C .

We prove the Propositions $\mathrm{S}, \mathrm{H}$, and B in the following steps. First, in Proposition 1.5:12, we manage to decompose every "basic non-harmonic" term $\left[\begin{array}{l}q \\ r\end{array}\right](x)=r^{-1} \underline{q} x$ into a sum of simpler terms. We use this result to show that we can get rid of non-harmonic terms completely (Proposition H, 1.5:3). This will enable us to prove Proposition B 1.5:4 and finally use its special case for a base $\left\langle b_{i}\right\rangle_{i<\omega}$ with $b_{i} \mid b_{i+1}$, to deduce Proposition S $1.5: 2$, too.

Our method of proof relies on a calculus of terms in $\mathcal{F}$, which is a generalization of the calculus of continued fractions.
Remark 1.5:1. The problem of descriptive analysis for linear theories (such as ZLA) turns out to be considerably harder than the same task for similar theories, e.g. the theory of $\mathbb{Z}$-groups (ZAA) - which is the simplest linear theory - or the theory of modules over an associative ring $R$. In fact, linear theories can be seen as generalizing both these cases.

The reason of greater complexity of definable sets in ZLA, compared to the theory of modules, lies, of course, in the presence of the ordering. The difference between ZLA and ZAA can be better understood by considering the following example:

Let $\varphi(x)$ be the formula $r^{-1}(q x)-\frac{q}{r} x \leq c$, where $q, r$ are scalars and $c$ is a constant.

The set $D$ defined by $\varphi$ is, clearly, $r$-periodical. In ZAA, it is $r \in \mathbb{Z}$, therefore $D$ can be written as a union of finitely many arithmetical progressions. Nevertheless, in ZLA, $r$ may be non-standard. That is why the finite decomposition trick does not work and $D$ needs to be examined in detail.

### 1.5.2 Main propositions

Here, we state three propositions, denoted S, H and B, which form three pillars of our proof of the theorems $1.3: 4,1.3: 6$ and $1.3: 8$. We prove the propositions in the following sections. In this section, we derive the theorems from them.

The following proposition is the crucial step in the proof of the Main Theorem on Linear Theories 1.3:4.

Proposition 1.5:2 (Proposition S). Let $\psi(x, \bar{y})$ be an open formula. There are finitely many terms $t_{i}(\bar{y}), i<n$, such that

$$
(\exists x) \psi \leftrightarrow \bigvee_{i<n} \psi\left(t_{i}, \bar{y}\right)
$$

The proposition below proves the Harmonic Forms Theorem 1.3:6.
Proposition 1.5:3 (Proposition H). Every term $t(\bar{x})$ is equivalent to a harmonic almost-term $\tau(\bar{x})$.

Moreover, if a variable $x$ does not occur in any subterm of the form $r^{-1}$ s (where $s$ is a term) in then the same is true in $\tau$.

The following proposition states the key step for the proof of the Bases Theorem 1.3:8.

Proposition 1.5:4 (Proposition B). Let $\delta \in F, p, r \in{ }^{+}$D be scalars, $e=\operatorname{deg}(p)$, $d=\operatorname{deg}(r)$, and $B=\left\langle b_{i}\right\rangle_{d \leq i<e}$ be a $[d, e]$-basis. Then $r^{-1} x$ is on $[\delta, \delta+p-1]$ equal to a harmonic almost-term $\tau(x)$ (possibly with parameter $\delta$ ) which is over $m B=\left\langle m b_{i}\right\rangle_{d \leq i<e}$, for some $m \in \mathbb{N}$.

Moreover, $m$ can be chosen as any number sufficiently large with respect to divisibility.

The proof of the Propositions S, H and B occupies a significant portion of this text. At this place, we derive the theorems $1.3: 4, ~ 1.3: 6$ and $1.3: 8$ from them.

### 1.5.2.1 Proof of the Main Theorem on Linear Theories

Theorem 1.3:4 (Main Theorem on Linear Theories). Let $T$ be a linear theory in a language $L, \mathcal{A} \mapsto \mathcal{F}_{\mathcal{A}}$ be a $(\mathrm{D}, \mathrm{C})$-linealization, and $E=\mathrm{D} \cup \mathrm{C} \cup \mathrm{D}^{-1}$. Then

1) $T^{E}$ is almost uniformly solvable.
2) $T^{\mathrm{C}}$ is 0-solvable.

Proof. 1): By Lemma 1.2:8, it is enough to prove that

$$
\begin{equation*}
\mathcal{A}^{E} \models(\exists x) \psi(x, \bar{y}) \rightarrow \bigvee_{i<n} \psi\left(t_{i}(\bar{y}), \bar{y}\right), \tag{1.11}
\end{equation*}
$$

for every $\mathcal{A} \models T$, quantifier-free $L^{E}$-formula $\psi$ and some $L^{E}$-terms $t_{i}, i<n$, depending on $\mathcal{A}, \psi$. In fact, we prove (1.11) even for arbitrary $L^{E}$-formula $\psi$ and with $t_{i}, i<n, L\left(\mathcal{F}_{\mathcal{A}}\right)$-terms.

Note that $\mathcal{A}^{E}$ is an expansion of the lineal $\mathcal{F}_{\mathcal{A}}$. Let $\psi$ be an $L^{E}$-formula. By easy translation, $\psi$ is in $\mathcal{A}^{E}$ equivalent to an $L\left(\mathcal{F}_{\mathcal{A}}\right)$-formula $\varphi$. By Proposition S
(1.5:2), $\operatorname{Th}\left(\mathcal{F}_{\mathcal{A}}\right)$ is almost uniformly solvable, hence solvable and admits quantifier elimination, by Proposition 1.2:6. Therefore,

$$
\mathcal{F}_{\mathcal{A}} \models(\exists x) \varphi(x, \bar{y}) \rightarrow \bigvee_{i<n} \varphi\left(t_{i}(\bar{y}), \bar{y}\right)
$$

for some $L\left(\mathcal{F}_{\mathcal{A}}\right)$-terms $t_{i}, i<n$. Then, clearly, (1.11) holds for these $t_{i}, i<n$. 2): Let $\psi$ be an $L^{\mathrm{C}}$-sentence. In 1), we proved that there are constant $L\left(\mathcal{F}_{\mathcal{A}}\right)$ terms $t_{i}, i<n$, such that (1.11) holds. By the definition of lineal (see 1.3.1.3), every $t_{i}$ is equal in $\mathcal{F}_{\mathcal{A}}$ to some constant $c_{i} \in \mathrm{C}_{\mathcal{A}}$.

### 1.5.2.2 Proof of the Harmonic Form Theorem

Theorem 1.3:6 (Harmonic Form Theorem). Let $\mathcal{F}$ be a lineal.

1) For every term $t(\bar{x})$, there is an open harmonic almost-term $\tau(\bar{x})$ such that $\mathcal{F} \models t(\bar{x})=\tau(\bar{x})$.
2) For every formula $\varphi(\bar{x})$, there is an open harmonic formula $\psi(\bar{x})$ such that $\mathcal{F} \models \varphi(\bar{x}) \leftrightarrow \psi(\bar{x})$.
Proof. 1) is an immediate consequence of the Proposition H 1.5:3. We may suppose that $\tau$ is open, thanks to Corollary 1.3:5 1) of Theorem 1.3:4.
3) follows from 1), by replacing all maximal subterms in $\varphi$ by their harmonic equivalents.

### 1.5.2.3 Proof of the Bases Theorem

Theorem 1.3:8 (Bases Theorem). Let $\bar{\delta} \in F^{n}, C_{p}(\bar{\delta})=\prod_{i<n}\left[\delta_{i}, \delta_{i}+p-1\right] \subseteq F^{n}$ be a cube with edges of scalar length $p \in \mathrm{D}, e=\operatorname{deg}(p)$, and $B$ be $a[0, e]$-basis. Let $l(\bar{x})=n$. Then the following holds:

1) Every term $t(\bar{x})$ is on $C_{p}(\bar{\delta})$ equal to a harmonic almost-term $\tau(\bar{x})$ which is over $m B$ for some $m \in \mathbb{N}$.
2) Every formula $\varphi(\bar{x})$ is on $C_{p}(\bar{\delta})$ equivalent to an open harmonic formula $\psi(\bar{x})$ which is over $m B$ for some $m \in \mathbb{N}$.

## Moreover:

- $m$ can be chosen as any number sufficiently large with respect to divisibility.
- If $t$ or $\varphi$ contain parameters from a set $X$ then $\tau$ and $\psi$ contain only parameters from $X \cup \bar{\delta}$.
Proof. 1): By Proposition H 1.5:3, $t(\bar{x})$ is equivalent to a harmonic almost-term $\sigma(\bar{x})$. Now it suffices to replace every subterm $r^{-1}\left(x_{i}\right)$ of $\sigma$ by its equivalent from Proposition B 1.5:4,

2) follows directly from 1), by replacing all maximal subterms of $\varphi$ by their equivalents.

### 1.5.3 Preliminaries of the proof

From now on, we are heading towards the proof of the Propositions S, H and B (see section 1.5.2).

We consider all the new symbols defined in the rest of this section just as abbreviations, i.e. we do not add them formally to the language. Elements from D (and their realizations in $\mathcal{F}$ ) are further often called just scalars, elements from C are referred to as constants. For a scalar $r$, we denote the constant $r 1$ also as $r$.

For better clarity of our formulas, we also freely use fractions with denominators from $\mathrm{D}-\{0\}$; expressions as $\frac{x}{r}$ or $x / r$ always denote fractions, while the integer division is strictly denoted as $r^{-1} x$.

The following Lemma is easy:
Lemma 1.5:5. Let $q, r, r^{\prime} \in \mathrm{D}$. Then the following holds:
a) $\operatorname{deg}(q)<\operatorname{deg}(r) \Rightarrow|q|<|r|$
b) $\operatorname{deg}(r)=0 \Leftrightarrow r \in \mathbb{Z}-\{0\}$
c) For $q \neq 0$ it is $\operatorname{deg}\left(r^{\prime}\right)<\operatorname{deg}(r) \Leftrightarrow \operatorname{deg}\left(q r^{\prime}\right)<\operatorname{deg}(q r)$
d) $\operatorname{deg}(q+r) \leq \max (\operatorname{deg}(q), \operatorname{deg}(r))$
e) $\operatorname{deg}(q r) \geq \operatorname{deg}(q)+\operatorname{deg}(r)$

Proof. We prove only 回) ; the other statements are trivial. We show that it is $\operatorname{deg}(q r) \geq \operatorname{deg}(q)+\operatorname{deg}(r)$, for a fixed $q$, by induction on $\operatorname{deg}(r)$. The case $\operatorname{deg}(r)=0$ follows from (b). For the induction step, we have, for all $r^{\prime}$ with $\operatorname{deg}\left(r^{\prime}\right)<\operatorname{deg}(r)$, the following: $\operatorname{deg}(q r)>\operatorname{deg}\left(q r^{\prime}\right) \geq \operatorname{deg}(q)+\operatorname{deg}\left(r^{\prime}\right)$. Hence, $\operatorname{deg}(q r) \geq \operatorname{deg}(q)+\operatorname{deg}(r)$.

For $0<q \in \mathrm{D}$, we define the remainder function for division by $q$ :

$$
\mu_{q}(x)=x-q q^{-1} x
$$

Moreover, we write shortly $\mu_{r_{1}, \ldots, r_{n}}(x)$ instead of $\mu_{r_{1}}\left(\mu_{r_{2}}\left(\ldots\left(\mu_{r_{n}}(x)\right) \ldots\right)\right)$.
The following is easy to prove:
Lemma 1.5:6. For all scalars $r, q>0$ and $x \in F$, it is:
a) $r^{-1} q^{-1} x=(r q)^{-1} x=q^{-1} r^{-1} x$,
b) $(q r)^{-1}(q x)=r^{-1} x$,
c) $0 \leq \mu_{r}(x)<r$,
d) $r^{-1} x=\frac{x-\mu_{r}(x)}{r}$,
e) $r^{-1}(x+y)=r^{-1} x+r^{-1} y+i_{r, x, y}$, where $i_{r, x, y}=0$ if $\mu_{r}(x)+\mu_{r}(y)<r$, and $i_{r, x, y}=1$, otherwise.

### 1.5.4 Continued fractions

We start to build the calculus of terms in $\mathcal{F}$. It is based on and generalizes the calculus of continued fractions.

For scalars $a_{1}, \ldots, a_{n} \in \mathrm{D}$, where $a_{i}>0$ for $i>1$, we denote

$$
\left[a_{1}, \ldots, a_{n}\right]=a_{1}+\frac{1}{a_{2}+\frac{1}{\ddots+\frac{1}{a_{n}}}}
$$

the continued fraction with coefficients $a_{1}, \ldots, a_{n}$.
The axiom (R2) from the definition of doded (see 1.3.1.2) ensures that for every $q, r \in \mathrm{D}$, with $r>0$, there are $n \in \mathbb{N}$ and $a_{1}, \ldots, a_{n} \in \mathrm{D}$ such that $q / r=\left[a_{1}, \ldots, a_{n}\right]$. Indeed, we may define $a_{i}=t_{i}{ }^{-1} s_{i}$, where $\left(s_{i}, t_{i}\right)$ is the partial result after the $(i-1)$-th step of the Euclidean algorithm, starting from $(q, r)$, i.e.

$$
\begin{array}{ll}
s_{1}=q, & t_{1}=r \\
s_{i+1}=t_{i}, & t_{i+1}=\mu_{t_{i}}\left(s_{i}\right)
\end{array}
$$

By (R2), the algorithm stops after $n+1$ steps, for some $n \in \mathbb{N}$. We define the nominators $q_{i}$ and denominators $r_{i}$ of the partial continued fractions $\left[a_{1}, \ldots, a_{i}\right]$, for $1 \leq i \leq n$. For technical purposes, we also set $q_{-1}=r_{0}=0, q_{0}=r_{-1}=1$. Further, we fix this notation, i.e. unless stated otherwise, given the pair $q, r \in \mathrm{D}$, $r>0$, the symbols $n, a_{i}, q_{i}$ and $r_{i}$ are defined for $q, r$ as above.

Lemma 1.5:7. Let $q, r \in \mathrm{D}$, and $1 \leq j \leq n$. For $i<j$, we set $x_{i}=\left[a_{i+1}, \ldots, a_{j}\right]$. Then
a) $q_{j}=a_{j} q_{j-1}+q_{j-2}, r_{j}=a_{j} r_{j-1}+r_{j-2}$,
b) $q_{j-1} r_{j}-q_{j} r_{j-1}=(-1)^{j+1}$,
c) If $i+1<j$ then $x_{i}=a_{i+1}+\frac{1}{x_{i+1}}$,
d) $\frac{q_{j}}{r_{j}}=\frac{q_{i} x_{i}+q_{i-1}}{r_{i} x_{i}+r_{i-1}}$,
e) $q / r=\left[a_{1}, \ldots, a_{n}\right]=q_{n} / r_{n}$, and $q_{n}, r_{n}$ are co-prime.

Proof. Easy.

### 1.5.5 Bracket $\left[\begin{array}{l}q \\ r\end{array}\right]$ and its decomposition

We introduce an useful concept of a bracket function $\left[\begin{array}{l}q \\ r\end{array}\right](x)$, for scalars $q, r$, $r>0$, as a cornerstone for our calculus of terms.

Brackets $\left[\begin{array}{l}1 \\ r\end{array}\right]=r^{-1}$ should be understood as "basic harmonic functions", while $\left[\begin{array}{l}q \\ r\end{array}\right]$, with $q \neq 1$, as non-harmonic ones, where the rate of non-harmonicity roughly corresponds to the complexity of the continued fraction of $q / r$. In this section, we prove that the graph of $\left[\begin{array}{l}q \\ r\end{array}\right]$ is a curve parametrized by a pair of linear forms defined on a "spiraloid" (see Proposition 1.5:12 1). As a consequence, every bracket $\left[\begin{array}{l}q \\ r\end{array}\right]$ can be decomposed into a sum of "more harmonic" terms (Proposition 1.5:12[2)).

### 1.5.5.1 Bracket $\left[\begin{array}{l}q \\ r\end{array}\right]$

Let $q, r$ be two scalars, $r>0$. We define the bracket

$$
\left[\begin{array}{l}
q \\
r
\end{array}\right](x)=r^{-1} q x
$$

If $q, r$ are not coprime, and $q^{\prime}, r^{\prime}$ are coprime and such that $\frac{q}{r}=\frac{q^{\prime}}{r^{\prime}}$ then, clearly, $\left[\begin{array}{l}q \\ r\end{array}\right](x)=\left[\begin{array}{l}q^{\prime} \\ r^{\prime}\end{array}\right](x)$, for all $x$, by Lemma 1.5:6 burther, let $q, r$ be be fixed coprime scalars, $r>0,\left[a_{1}, \ldots, a_{n}\right]$ (where $a_{1}<0$ if $q<0$ ) be the continued fraction of $\frac{q}{r}$ and $q_{i}, r_{i}$ the nominators and denominators of the partial continued fractions (for instance, $q_{n}=q$ and $r_{n}=r$ ).

For the proof of Proposition 1.5:12, we will need the following lemmas.
Lemma 1.5:8. For $i=0, \ldots, n-1$, let $m_{i}=r_{i} \cdot\left[a_{i+1}, \ldots, a_{n}\right]+r_{i-1}$. Then the following holds for any $z_{i} \in F$ :

$$
\frac{q_{n}}{r_{n}} \cdot r_{i} z_{i}=q_{i} z_{i}+\frac{(-1)^{i+1}}{m_{i}} z_{i}
$$

Proof. Denote $x_{i}=\left[a_{i+1}, \ldots, a_{n}\right]$. Then, by Lemma 1.5:7,

$$
\begin{aligned}
\frac{q_{n}}{r_{n}} \cdot r_{i} z_{i} & =\frac{q_{i} x_{i} r_{i} z_{i}+q_{i-1} r_{i} z_{i}}{r_{i} x_{i}+r_{i-1}}= \\
& =\frac{\left(q_{i} x_{i} r_{i} z_{i}+q_{i} z_{i} r_{i-1}\right)+\left(q_{i-1} r_{i} z_{i}-q_{i} z_{i} r_{i-1}\right)}{r_{i} x_{i}+r_{i-1}}= \\
& =q_{i} z_{i}+\frac{q_{i-1} r_{i}-q_{i} r_{i-1}}{r_{i} x_{i}+r_{i-1}} \cdot z_{i}= \\
& =q_{i} z_{i}+\frac{(-1)^{i+1}}{r_{i} x_{i}+r_{i-1}} z_{i} .
\end{aligned}
$$

Lemma 1.5:9. Let $\bar{z}=\left(z_{1}, \ldots, z_{n-1}\right) \in\left[0 ; a_{2}-1\right] \times \prod_{i=2}^{n-1}\left[0 ; a_{i+1}\right]$, and let $m_{i}$ denotes $r_{i} \cdot\left[a_{i+1}, \ldots, a_{n}\right]+r_{i-1}$. Then

$$
\sum_{i=1}^{n-1} \frac{z_{i}}{m_{i}} \leq 1-\frac{1}{r_{n}}
$$

Moreover, the bound is tight.
Proof. Clearly, it is enough to prove

$$
\begin{equation*}
\sum_{i=1}^{n-1} \frac{a_{i+1}}{m_{i}}-\frac{1}{m_{1}}=1-\frac{1}{r_{n}} \tag{1.12}
\end{equation*}
$$

Denote $x_{i}=\left[a_{i+1}, \ldots, a_{n}\right]$ as in the previous proof. At first, we express all $m_{i}$ 's in terms of $x_{j}$ 's. For $i<m-1$, it is, by Lemma 1.5:7 c),

$$
\begin{gathered}
m_{i}=r_{i} x_{i}+r_{i-1}=\left(r_{i} a_{i+1}+r_{i-1}\right)+\frac{r_{i}}{x_{i+1}}=r_{i+1}+\frac{r_{i}}{x_{i+1}}=\frac{m_{i+1}}{x_{i+1}} \\
m_{0}=r_{0} x_{0}+r_{-1}=0 \cdot x_{0}+1=1
\end{gathered}
$$

Hence,

$$
\begin{equation*}
m_{i}=\prod_{j=1}^{i} x_{j} \tag{1.13}
\end{equation*}
$$

Let $x_{i}=\frac{s_{i}}{t_{i}}$, where $s_{i}, t_{i}$ are relatively prime. For $i=0, \ldots, n-2$, using Lemma $1.5: 7 \mathrm{c}$ ), we can get $\frac{s_{i}}{t_{i}}=\frac{a_{i+1} s_{i+1}+t_{i+1}}{s_{i+1}}$ and thus (since $a_{i+1} s_{i+1}+t_{i+1}$ and $s_{i+1}$ are, clearly, relatively prime)

$$
\begin{equation*}
s_{i+1}=t_{i} \tag{1.14}
\end{equation*}
$$

and

$$
\begin{align*}
a_{i+1} s_{i+1} & =s_{i}-t_{i+1} \\
a_{i+1} t_{i} & =s_{i}-t_{i+1} \tag{1.15}
\end{align*}
$$

For $i=n-1$, we have $t_{n-1} a_{n}=1 \cdot a_{n}=s_{n-1}$.
From (1.14), we get

$$
\begin{equation*}
m_{i} \stackrel{1.13}{=} \prod_{j=1}^{i}=\frac{s_{1}}{t_{1}} \cdot \frac{s_{2}}{t_{2}} \cdots \cdots \frac{s_{i}}{t_{i}}=\frac{s_{1}}{t_{i}} \tag{1.16}
\end{equation*}
$$

and further,

$$
\begin{equation*}
\sum_{i=1}^{n-1} \frac{a_{i+1}}{m_{i}} \stackrel{1.16}{=} \sum_{i=1}^{n-1} \frac{t_{i} a_{i+1}}{s_{1}} \stackrel{1.15}{=} \frac{\sum_{i=1}^{n-2}\left(s_{i}-t_{i+1}\right)+s_{n-1}}{s_{1}} \stackrel{\sqrt{1.14}}{=} \frac{s_{1}+s_{2}-t_{n-1}}{s_{1}} \tag{1.17}
\end{equation*}
$$

Since, obviously, $s_{1}=r_{n}, s_{2}=t_{1}, t_{n-1}=1, m_{1}=\frac{s_{1}}{t_{1}}$, we have

$$
\sum_{i=1}^{n-1} \frac{a_{i+1}}{m_{i}}-\frac{1}{m_{1}} \stackrel{1.17}{=} \frac{r_{n}+t_{1}-1}{r_{n}}-\frac{t_{1}}{r_{n}}=1-\frac{1}{r_{n}}
$$

### 1.5.5.2 $\mathcal{C}_{r}^{q}$, forms $f_{r}, f_{q}$

We define the "cuboid" $\mathcal{C}_{r}^{q}$ and linear forms $f_{r}, f_{q}: \mathcal{C}_{r}^{q} \rightarrow F$ as

$$
\begin{gathered}
\mathcal{C}_{r}^{q}=\left[0 ; a_{2}-1\right] \times \prod_{i=2}^{n-1}\left[0 ; a_{i+1}\right] \times(\leftarrow ; \rightarrow), \\
f_{r}(\bar{z})=\sum_{i=1}^{n}(-1)^{i+1} r_{i} z_{i}, \quad f_{q}(\bar{z})=\sum_{i=1}^{n}(-1)^{i+1} q_{i} z_{i} .
\end{gathered}
$$

The following lemma states that the form $f_{r}$ gives a "cuboid parametrization" of $F$.

Lemma 1.5:10. The form $f_{r}$ is surjective and $(\leq 2)$-to-1.
Proof. Easy.
Remark 1.5:11. Let $\mathcal{S}_{r}^{q} \subseteq \mathcal{C}_{r}^{q}$ be the set of all $\leq_{\text {Lex }}$-maximal elemets of $\sim_{q, r}$-factor classes (where $\leq_{L e x}$ is the lexicographical order of $\mathcal{C}_{r}^{q}$, and the equivalence $\sim_{q, r}$ is defined by $\left.\bar{z} \sim_{q, r} \overline{z^{\prime}} \Leftrightarrow f_{r}(\bar{z})=f_{r}\left(\overline{z^{\prime}}\right)\right)$.

It is easy to see that the form $f_{r} \upharpoonright \mathcal{S}_{r}^{q}$ is an isomorphism of $\left\langle\mathcal{S}_{r}^{q}, \leq_{\text {Lex }}\right\rangle$ and $\left\langle F, \leq^{F}\right\rangle$. Hence, $\left\langle F, \leq^{F}\right\rangle$ can be imagined as a spiral "wrapped around" $\mathcal{S}_{r}^{q}$, that is why we call $\mathcal{S}_{r}^{q}$ a "spiraloid".

Proposition 1.5:12 (cuboid decomposition of $\left[\begin{array}{l}q \\ r\end{array}\right]$ ). Let $q, r$ be scalars, $r>0$. Then the following holds:

1) $\left[\begin{array}{l}q \\ r\end{array}\right] \circ f_{r}=f_{q}$ on $\mathcal{C}_{r}^{q}$.
2) Let $s_{q, r}=\sum_{i=1}^{\left\lfloor\frac{n-1}{2}\right\rfloor} r_{2 i} a_{2 i+1}, t_{q, r}=\sum_{i=1}^{\left\lfloor\frac{n-1}{2}\right\rfloor} q_{2 i} a_{2 i+1}$. Then, for all $x \in F$,

$$
\left[\begin{array}{l}
q  \tag{1.18}\\
r
\end{array}\right](x)=\sum_{i=1}^{n} q_{i} \cdot r_{i}^{-1} \mu_{r_{i+1}, \ldots, r_{n}}\left(x+s_{q, r}\right)-t_{q, r} .
$$

We call the expression on the right side of (1.18) the cuboid decomposition of the bracket $\left[\begin{array}{l}q \\ r\end{array}\right](x)$.

Proof. (of Proposition 1.5:12)
11) Denote $x=f_{r}(\bar{z})=\sum_{i=1}^{n}(-1)^{i+1} r_{i} z_{i}$, for fixed $\bar{z} \in \mathcal{C}_{r}^{q}$. Then we have

$$
\begin{aligned}
& {\left[\begin{array}{l}
q \\
r
\end{array}\right](x) }=\left\lfloor\frac{q_{n}}{r_{n}} x\right\rfloor \text { Lemm\& } 1.5: 8 \\
&= \\
& i=1 \\
& n \\
&\left.(-1)^{i+1} q_{i} z_{i}+\sum_{i=1}^{n-1} \frac{z_{i}}{m_{i}}\right\rfloor= \\
&=\sum_{i=1}^{n}(-1)^{i+1} q_{i} z_{i}+\left\lfloor\sum_{i=1}^{n-1} \frac{z_{i}}{m_{i}}\right\rfloor \stackrel{\text { Lemmd }}{=} 1.5: 9 \\
& \sum_{i=1}^{n}(-1)^{i+1} q_{i} z_{i}=f_{q}(\bar{z}) .
\end{aligned}
$$

(2) is a corollary of (1) via a change of coordinates. Let $x \in F$ be given. By Lemma 1.5:10, we can choose coordinates $\bar{z} \in \mathcal{C}_{r}^{q}$ such that $x=\sum_{i=1}^{n}(-1)^{i+1} r_{i} z_{i}$. Set

$$
z_{i}^{\prime}=\left\{\begin{array}{rl}
a_{i+1}-z_{i} & i<n \text { even } \\
z_{i} & i<n \text { odd } \\
(-1)^{n+1} z_{n} & i=n
\end{array}\right.
$$

Then

$$
\begin{equation*}
x+s_{q, r}=\sum_{i=1}^{n} r_{i} z_{i}^{\prime}=\sum_{i=1}^{n} r_{i} z_{i}^{\prime \prime} \tag{1.19}
\end{equation*}
$$

where $\overline{z^{\prime \prime}} \in \mathcal{S}_{r}^{q}$. By (1), we get also

$$
\begin{equation*}
\sum_{i=1}^{n} q_{i} z_{i}^{\prime}=\sum_{i=1}^{n} q_{i} z_{i}^{\prime \prime} \tag{1.20}
\end{equation*}
$$

and, by maximality of $\overline{z^{\prime \prime}}$ and (1.19),

$$
\begin{equation*}
z_{i}^{\prime \prime}=r_{i}^{-1} \mu_{r_{i+1}, \ldots, r_{n}}\left(x+s_{q, r}\right) \tag{1.21}
\end{equation*}
$$

Finaly, we can compute

$$
\begin{aligned}
{\left[\begin{array}{l}
q \\
r
\end{array}\right](x) } & \stackrel{11}{=} \sum_{i=1}^{n}(-1)^{i+1} q_{i} z_{i}=\sum_{i=1}^{n} q_{i} z_{i}^{\prime}-t_{q, r} \stackrel{1.20}{=} \sum_{i=1}^{n} q_{i} z_{i}^{\prime \prime}-t_{q, r} \stackrel{1.211}{=} \\
& \stackrel{1.21}{=} \sum_{i=1}^{n} q_{i} \cdot r_{i}^{-1} \mu_{r_{i+1}, \ldots, r_{n}}\left(x+s_{q, r}\right)-t_{q, r}
\end{aligned}
$$

### 1.5.6 Harmonization

In this subsection, we prove the Proposition H 1.5:3. That is, we show that every term in $\mathcal{F}$ can be (up to a "noise") expressed as a linear combination of basic "harmonic" functions $r^{-1}$, and, consequently, any open formula is equivalent to one with all terms harmonic (see 1.3.2.2.1).

We are going to develop a calculus of brackets $\left[\begin{array}{l}q \\ r\end{array}\right]$ (see 1.5.5.1 for definition). For the reason of simplicity, we will write

$$
\left[\begin{array}{lll}
a_{1} & \ldots & a_{n} \\
b_{1} & \ldots & b_{n}
\end{array}\right]
$$

instead of

$$
\left[\begin{array}{l}
a_{1} \\
b_{1}
\end{array}\right] \circ \ldots \circ\left[\begin{array}{l}
a_{n} \\
b_{n}
\end{array}\right] .
$$

### 1.5.6.1 Harmonic terms

We rewrite here the definition of harmonic term (see 1.3.2.2.1) in the bracket notation: A term $t(\bar{x})$ is harmonic (or equivalently in harmonic form) if it is a sum of expressions of the form $\left[\begin{array}{ll}A & 1 \\ 1 & B\end{array}\right]\left(x_{i}\right)$, with $A, B \in \mathrm{D}$, and possibly a scalar $r \in \mathrm{D}$.

### 1.5.6.2 Convention

The rest of this subsection is devoted to the proof of Proposition H 1.5:3. For the purpose of this proof, we further consider (the so far abbreviations) binary minus (-) and $\mu_{r}()_{-}$to be symbols in our language. On the other hand, we forbid the unary minus (it can be replaced using multiplication by the scalar -1 ). (Formally, we get a modification $L^{\prime}$ of our original language $L(\mathcal{F})$. But the difference between $L(\mathcal{F})$ and $L^{\prime}$ is of purely technical character, since there is a simple translation between these two languages.) In the rest of this subsection, the word "term" means $L^{\prime}$-term and similarly for almost-term and formula.

The idea of the proof is to lower the "non-harmonicity" of a term (almostterm) by decomposing its "strings" $\left[\begin{array}{lll}a_{1} & \ldots & a_{n} \\ b_{1} & \ldots & b_{n}\end{array}\right]$, using Proposition 1.5:12 and "almost-distributivity" of strings over addition (see Lemma 1.5:16 below). The non-harmonicity of newly created strings needs to be controled during the proces. For these purposes, we introduce a few new concepts.

### 1.5.6.3 Strings

Let $t$ be a term. In the tree of subterms of $t$, vertices correspond to subterms (root to $t$, leafs to variables and constants) and edges to symbols.

Example 1.5:13. Let $t=q\left(r_{0} s^{-1} x-r_{1}\left(y+t^{-1} r_{2} c\right)\right)$ be a term, with $x, y$ variables, $c$ a constant and $q, r_{0}, r_{1}, r_{2}, s, t$ scalars. The corresponding tree of subterms of $t$ is


Any sequence of symbols from edges and the leaf of a branch in the tree of subterms of $t$, in the ascending order from the root to the leaf, in which we omit all edges corresponding to symbols,+- , is called a string of $t$. The string is a + -string $[-$-string $]$ if the number of omitted symbols - was even [odd]. The set of all strings [+-strings, - -strings] of $t$ is denoted $\operatorname{str}(t)\left[\operatorname{str}^{+}(t), \operatorname{str}^{-}(t)\right]$. The set of strings of an p-term $\tau$ [formula $\varphi$ ] (denoted $\operatorname{str}(\tau)[\operatorname{str}(\varphi)]$ ) is the union of the sets $\operatorname{str}(t)$, over all maximal subterms $t$ of $\tau[\varphi]$.
Example 1.5:14. Let $t=q\left(r_{0} s^{-1} x-r_{1}\left(y+t^{-1} r_{2} c\right)\right)$ be a term, as in Example 1.5:13. There are three strings of $t: \alpha_{0}=\left\langle q, r_{0}, s^{-1}, x\right\rangle, \alpha_{1}=\left\langle q, r_{1}, y\right\rangle$ and $\alpha_{2}=\left\langle q, r_{1}, t^{-1}, r_{2}, c\right\rangle$. Two of them, $\alpha_{1}$ and $\alpha_{2}$, are --strings, while $\alpha_{0}$ is a + -string.

The reduced string $\tilde{\alpha}$ arises from a string $\alpha$ by removing the maximal initial segment of $\alpha$ consisting only of scalar multiplications. We denote the set of reduced strings of $t$ as $\operatorname{s\tilde {tr}}(t)=\{\tilde{\alpha} ; \alpha \in \operatorname{str}(t)\}$ and similarly for $\operatorname{str}(\tau)$ and $\tilde{\operatorname{str}}(\varphi)$.

Example 1.5:15. Let $\alpha_{0}=\left\langle q, r_{0}, s^{-1}, x\right\rangle, \alpha_{1}=\left\langle q, r_{1}, y\right\rangle$ and $\alpha_{2}=\left\langle q, r_{1}, t^{-1}, r_{2}, c\right\rangle$ be the strings from Example 1.5:14. Their reduced versions are $\tilde{\alpha_{0}}=\left\langle s^{-1}, x\right\rangle$, $\tilde{\alpha_{1}}=\langle y\rangle$ and $\tilde{\alpha_{2}}=\left\langle t^{-1}, r_{2}, c\right\rangle$.

### 1.5.6.4 Notation $T \rightarrow T^{\prime}, \llbracket T \rrbracket_{S}, \alpha \sqsubseteq_{n} \beta$ and $\alpha \sqsubseteq \beta$

The following notation will help us to keep the complexity of almost-terms under control, during the proof of Proposition H .

For two strings $\alpha, \beta$ and $n \in \omega$, we write $\alpha \sqsubseteq_{n} \beta$ if there are at most $n$ symbols $f_{0}, \ldots, f_{i-1}$ such that $\beta=\left\langle f_{0}, \ldots, f_{i-1}\right\rangle \smile \alpha$. We write $\alpha \sqsubseteq \beta$ if $\alpha \sqsubseteq_{n} \beta$, for some $n \in \omega$.

For two sets of terms $T, S$, we denote $\llbracket T \rrbracket_{S}$ the set of all almost-terms $\tau$ with $\operatorname{core}(\tau) \in T$ and such that all maximal subterms in $\operatorname{cond}(\tau)$ are of the form $\sum_{i} q_{i} \cdot s_{i}+r$, where $r, q_{i} \in \mathrm{D}$ and $s_{i} \in S$. We will often write $\llbracket t \rrbracket_{S}$ instead of $\llbracket\{t\} \rrbracket_{S}$. We will also write $\llbracket t \rrbracket_{\square}$ instead of $\llbracket t \rrbracket_{\{\alpha ; \exists \beta \in \operatorname{str}(t)(\alpha \square \beta)\}}$, where $\square$ stands for any symbol $\sqsubseteq, \sqsubseteq_{n}$. We will sometimes write $\llbracket t \rrbracket$ instead of $\llbracket t \rrbracket_{\sqsubseteq_{0}}$ and $\llbracket t \rrbracket_{=}$instead of $\llbracket t \rrbracket_{\{t\}}$. We may also use abbreviations with obvious meaning such as $\llbracket t \rrbracket_{\sqsubseteq_{i} \cup=}$.

Let $T, T^{\prime}$ be two sets of almost-terms. We say that $T$ reduces to $T^{\prime}$ and denote it $T \rightarrow T^{\prime}$ if every almost-term $\tau \in T$ is equivalent to some $\tau^{\prime} \in T^{\prime}$. We will often write $\tau \rightarrow T^{\prime}$ instead of $\{\tau\} \rightarrow T^{\prime}$ and similarly on the right side.

We say that a term $t$ is distributed if it is a sum of strings. A p-term [a formula] is distributed if all its maximal subterms are.

Lemma 1.5:16. Let $t, s$ be terms and $r \in \mathrm{D}, r>0$. Then
a) $r^{-1}(t \pm s) \rightarrow \llbracket r^{-1} t \pm r^{-1} s \rrbracket_{\left\{t, s, r^{-1} t, r^{-1} s\right\}}$
b) $\mu_{r}(t \pm s) \rightarrow \llbracket \mu_{r}(t) \pm \mu_{r}(s) \rrbracket_{\left\{\mu_{r}(t), \mu_{r}(s)\right\}}$
c) $t \rightarrow \llbracket \sum \operatorname{str}^{+}(t)-\sum \operatorname{str}^{-}(t) \rrbracket_{\sqsubseteq}$
d) Every almost-term is equivalent to a distributed one.

Proof.
a) It is easy to see that

$$
r^{-1} x+y= \begin{cases}r^{-1} x+r^{-1} y & \text { if } \mu_{r}(x)+\mu_{r}(y)<r \\ r^{-1} x+r^{-1} y+1 & \text { if } \mu_{r}(x)+\mu_{r}(y) \geq r\end{cases}
$$

and similarly for $x-y$.
b) Similarly as a).
c) Follows from a) and b), by induction on the maximal depth of occurrence of ,+- in $t$.
d) Easy consequence of a) and b).

The following lemma lists basic techniques of complexity reduction, which we are going to use throughout our proof.

Here and further on, we use notation $s(\bar{T})$, where $s$ is a term, and $T_{i}$ are sets of terms, to denote the set of all terms $s(\bar{t})$, where $t_{i} \in T_{i}$. Let $S$ be a set of terms. Then $\operatorname{sTm}(S)$ denotes the set of all subterms of terms from $S$.

Lemma 1.5:17. Let $t_{i}, t, s$ be terms, $T, S, U, V, W, T_{i}, S_{i}$ sets of terms and $X_{i}, Y_{i}$ sets of almost-terms.

1) a) Relation $\rightarrow$ is a preorder on $\mathcal{P}(a T m)$, where aTm stands for the set of all almost-terms.
b) If $X_{i} \rightarrow Y_{i}$ then $s\left(X_{0}, \ldots, X_{n-1}\right) \rightarrow s\left(Y_{0}, \ldots, Y_{n-1}\right)$.
c) If $t \rightarrow \llbracket s \rrbracket_{S}$ then $\llbracket t \rrbracket_{T} \rightarrow \llbracket s \rrbracket_{T \cup S}$.
d) If $T \rightarrow \llbracket S \rrbracket_{V}$ and $U \rightarrow \llbracket W \rrbracket_{W}$ then $\llbracket T \rrbracket_{U} \rightarrow \llbracket S \rrbracket_{V \cup W}$.
2) a) The following holds

$$
s\left(\llbracket t_{0} \rrbracket_{T_{0}}, \ldots, \llbracket t_{n-1} \rrbracket_{T_{n-1}}\right) \rightarrow \llbracket s\left(t_{0}, \ldots, t_{n-1}\right) \rrbracket_{\bigcup_{i=0}^{n-1} T_{i} \cup \mathrm{STm}(s)(\bar{t})}
$$

Moreover, if $s$ does not contain divs nor mods then

$$
s\left(\llbracket t_{0} \rrbracket_{T_{0}}, \ldots, \llbracket t_{n-1} \rrbracket_{T_{n-1}}\right) \rightarrow \llbracket s\left(t_{0}, \ldots, t_{n-1}\right) \rrbracket_{\bigcup_{i=0}^{n-1} T_{i}}
$$

b) $\llbracket t \rrbracket_{T} \circ\left(\llbracket s_{0} \rrbracket_{S_{0}}, \ldots, \llbracket s_{n} \rrbracket_{S_{n}}\right) \rightarrow \llbracket t \circ \bar{s} \rrbracket_{\bigcup_{i=0}^{n-1} S_{i} \cup s T \mathrm{~T}(t, T) \mathrm{os}}$.

Proof. 1): Easy verification.
2) a): By induction on complexity of $s$. In the induction steps for $r^{-1}$ and $\mu_{r}$, Lemma $1.5: 16$ a) and b) is used.
2) b): Follows directly from 2) a).

By the Proposition (1.5:12(2), every bracket $\left[\begin{array}{l}q \\ r\end{array}\right]$ can be expressed in a cuboid form

$$
\left[\begin{array}{l}
q  \tag{1.22}\\
r
\end{array}\right](x)=\sum_{i=1}^{n} q_{i} \cdot\left(\left[\begin{array}{l}
1 \\
r_{i}
\end{array}\right] \mu_{r_{i+1}, \ldots, r_{n}}(x+s)\right)-t
$$

where $q_{i}, r_{i} \in \mathrm{D}$ are the nominator and the denominator of the $i$-th partial continued fraction of $\frac{q}{r}$, and $s, t \in \mathrm{C}$. In the following series of lemmas, we will reduce $\left[\begin{array}{l}q \\ r\end{array}\right]$ to a simpler form. For the reason of simplicity, we often use functional notation for terms, i.e. we write, for example, $\mu_{r}$ instead of $\mu_{r}(x)$ or $i d$ instead of $x$.

Lemma 1.5:18. Let $b_{0} \leq \ldots \leq b_{k-1}$, with $k>1$, be scalars having the same degree (in the doded D). Then
a) $\mu_{b_{1}, \ldots, b_{k-1}} \rightarrow \llbracket \mu_{b_{k-1}} \rrbracket$,
b) $\left[\begin{array}{c}1 \\ b_{0}\end{array}\right] \mu_{b_{1}, \ldots, b_{k-1}} \rightarrow \llbracket 0 \rrbracket_{\mu_{b_{k-1}}}$.

Proof. We will prove both statements, (a) and (b), of the Lemma simultaneously by induction on $k$.

For $k=2$, (可) is trivial. To prove (B), consider $\left[\begin{array}{c}1 \\ b_{0}\end{array}\right] \mu_{b_{1}}(x)$. We have

$$
\mu_{b_{1}}(x)<b_{1} \leq m \cdot b_{0}
$$

for some $m \in \omega\left(\right.$ since $\left.\operatorname{deg}\left(b_{0}\right)=\operatorname{deg}\left(b_{1}\right)\right)$. Then

$$
\left[\begin{array}{c}
1 \\
b_{0}
\end{array}\right] \mu_{b_{1}}(x)=\left\{i \text { if } \psi_{i} ; i<m+1,\right.
$$

where $\psi_{i}(x)=i b_{0} \leq \mu_{b_{1}}(x)<(i+1) \cdot b_{0}$. Clearly, the later is an almost-term in $\llbracket 0 \rrbracket_{\mu_{b_{1}}}$.

For the induction step in (a), we have

$$
\begin{aligned}
\mu_{b_{1}, \ldots, b_{k-1}} & =\mu_{b_{1}} \circ \mu_{b_{2}, \ldots, b_{k-1}}= \\
& =\mu_{b_{2}, \ldots, b_{k-1}}-\left[\begin{array}{cc}
b_{1} & 1 \\
1 & b_{1}
\end{array}\right] \circ \mu_{b_{2}, \ldots, b_{k-1}} \rightarrow \\
& \rightarrow \llbracket \mu_{b_{k-1}} \rrbracket-\llbracket 0 \rrbracket_{\mu_{b_{k-1}}} \rightarrow \llbracket \mu_{b_{k-1}} \rrbracket,
\end{aligned}
$$

where the first arrow follows from the induction assumptions and Lemma 1.5:17 1b), 2a) and the second from 2a). The induction step for (B):

$$
\left[\begin{array}{c}
1 \\
b_{0}
\end{array}\right] \mu_{b_{1}, \ldots, b_{k-1}} \rightarrow\left[\begin{array}{c}
1 \\
b_{0}
\end{array}\right] \llbracket \mu_{b_{k-1}} \rrbracket \rightarrow \llbracket\left[\begin{array}{c}
1 \\
b_{0}
\end{array}\right] \mu_{b_{k-1}} \rrbracket_{\sqsubseteq} \rightarrow \llbracket 0 \rrbracket_{\mu_{b_{k-1}}},
$$

where the first arrow follows from (四), the second one from Lemma 1.5:17 2a), and the third one from the induction assumption for $k=2$ and Lemma 1.5:17 1d).

Lemma 1.5:19. Let $b_{0} \leq \ldots \leq b_{k-1}, k>0$, be scalars. For any finite set $F \subseteq \omega$, we take an enumeration $F=\left\{f_{0}, \ldots, f_{|F|-1}\right\}$ such that $f_{0}<\ldots<f_{|F|-1}$.
a) Then

$$
\mu_{b_{1}, \ldots, b_{k-1}} \rightarrow \llbracket \sum_{F \subseteq\{1, \ldots, k-1\}}(-1)^{|F|}\left[\begin{array}{ccccc}
b_{f_{0}} & b_{f_{1}} & \ldots & b_{f_{|F|-1}} & 1 \\
1 & b_{f_{0}} & \ldots & b_{f_{|F|-2}} & b_{f_{|F|-1}}
\end{array}\right] \rrbracket
$$

b) and

$$
\left[\begin{array}{c}
1 \\
b_{0}
\end{array}\right] \mu_{b_{1}, \ldots, b_{k-1}} \rightarrow \llbracket \sum_{F \subseteq\{1, \ldots, k-1\}}(-1)^{|F|}\left[\begin{array}{ccccc}
b_{f_{0}} & b_{f_{1}} & \ldots & b_{f_{|F|-1}} & 1 \\
b_{0} & b_{f_{0}} & \ldots & b_{f_{|F|-2}} & b_{f_{|F|-1}}
\end{array}\right] \rrbracket_{\sqsubseteq_{1}}
$$

Proof. By simultaneous induction on $k$. The case $k=1$ is trivial. Let $k=2$.
a): $\mu_{b_{1}}=i d-\left[\begin{array}{cc}b_{1} & 1 \\ 1 & b_{1}\end{array}\right] \in \llbracket \sum_{F \subseteq\{1\}}(-1)^{|F|}\left[\begin{array}{cc}b_{f_{0}} & 1 \\ 1 & b_{f_{0}}\end{array}\right] \rrbracket$.
b): $\left[\begin{array}{c}1 \\ b_{0}\end{array}\right] \mu_{b_{1}}=\left[\begin{array}{c}1 \\ b_{0}\end{array}\right]\left(i d-\left[\begin{array}{ll}b_{1} & 1 \\ 1 & b_{1}\end{array}\right]\right) \rightarrow \llbracket\left[\begin{array}{c}1 \\ b_{0}\end{array}\right]-\left[\begin{array}{ll}b_{1} & 1 \\ b_{0} & b_{1}\end{array}\right] \rrbracket_{\sqsubseteq_{1}}$, where the arrow follows from Lemma 1.5:16 a).

Induction step: a):

$$
\mu_{b_{1}, \ldots, b_{k-1}}=\mu_{b_{2}, \ldots, b_{k-1}}-\left[\begin{array}{cc}
b_{1} & 1  \tag{1.23}\\
1 & b_{1}
\end{array}\right] \mu_{b_{2}, \ldots, b_{k-1}}
$$

By induction assumptions on a) and b), we get

$$
\begin{aligned}
\mu_{b_{2}, \ldots, b_{k-1}} & \rightarrow\left\|\sum_{F \subseteq\{2, \ldots, k-1\}}(-1)^{|F|}\left[\begin{array}{ccccc}
b_{f_{0}} & b_{f_{1}} & \ldots & b_{f_{|F|-1}} & 1 \\
1 & b_{f_{0}} & \ldots & b_{f_{|F|-2}} & b_{f_{|F|-1}}
\end{array}\right]\right\|, \\
{\left[\begin{array}{cc}
b_{1} & 1 \\
1 & b_{1}
\end{array}\right] \mu_{b_{2}, \ldots, b_{k-1}} } & \left.\rightarrow \| \sum_{F \subseteq\{2, \ldots, k-1\}}(-1)^{|F|}\left[\begin{array}{cccccc}
b_{1} & b_{f_{0}} & b_{f_{1}} & \ldots & b_{f_{|F|-1}} & 1 \\
1 & b_{1} & b_{f_{0}} & \ldots & b_{f_{|F|-2}} & b_{f_{|F|-1}}
\end{array}\right]\right]_{\sqsubseteq_{1}} .
\end{aligned}
$$

For the string $\beta=\left[\begin{array}{cccccc}b_{1} & b_{f_{0}} & b_{f_{1}} & \ldots & b_{f_{|F|-1}} & 1 \\ 1 & b_{1} & b_{f_{0}} & \ldots & b_{f_{|F|-2}} & b_{f_{|F|-1}}\end{array}\right]$, there are only two strings $\alpha \sqsubseteq_{1} \tilde{\beta}$, namely $\beta$ itself and $\left[\begin{array}{ccccc}b_{f_{0}} & b_{f_{1}} & \ldots & b_{f_{|F|-1}} & 1 \\ 1 & b_{f_{0}} & \ldots & b_{f_{|F|-2}} & b_{f_{|F|-1}}\end{array}\right]$. That is why we get from (1.23) and Lemma 1.5:17 2a) the following:

$$
\mu_{b_{1}, \ldots, b_{k-1}} \rightarrow \llbracket \sum_{F \subseteq\{1, \ldots, k-1\}}(-1)^{|F|}\left[\begin{array}{ccccc}
b_{f_{0}} & b_{f_{1}} & \ldots & b_{f_{|F|-1}} & 1 \\
1 & b_{f_{0}} & \ldots & b_{f_{|F|-2}} & b_{f_{|F|-1}}
\end{array}\right] \rrbracket
$$

To prove b), consider

$$
\begin{aligned}
{\left[\begin{array}{c}
1 \\
b_{0}
\end{array}\right] \mu_{b_{1}, \ldots, b_{k-1}} } & \rightarrow \llbracket\left[\begin{array}{c}
1 \\
b_{0}
\end{array}\right] \sum_{F}(-1)^{|F|}\left[\begin{array}{ccccc}
b_{f_{0}} & b_{f_{1}} & \ldots & b_{f_{|F|-1}} & 1 \\
1 & b_{f_{0}} & \ldots & b_{f_{|F|-2}} & b_{f_{|F|-1}}
\end{array}\right] \|_{\left(\sqsubseteq_{1} \cup=\right)} \rightarrow \\
& \rightarrow \llbracket \sum_{F}(-1)^{|F|}\left[\begin{array}{ccccc}
b_{f_{0}} & b_{f_{1}} & \ldots & b_{f_{|F|-1}} & 1 \\
b_{0} & b_{f_{0}} & \ldots & b_{f_{|F|-2}} & b_{f_{|F|-1}}
\end{array}\right] \rrbracket_{\sqsubseteq_{1}}
\end{aligned}
$$

where the first arrow follows from a) and Lemma 1.5:17 2a), and the second one from Lemma 1.5:16 a) and Lemma 1.5:17 1d).

Lemma 1.5:20. Let $\frac{q}{r}=\left[a_{1}, \ldots, a_{n}\right]$ and $i$ is maximal such that $\operatorname{deg}\left(a_{i}\right)>0$ ( $i=0$ if all $a_{i} \in \mathbb{Z}$ ). Then

$$
\left[\begin{array}{c}
q \\
r
\end{array}\right] \rightarrow \llbracket \sum_{j=1}^{i-1} \sum_{F \subseteq\{j+1, \ldots, i-1\} \cup\{n\}}(-1)^{|F|}\left[\begin{array}{cccccc}
q_{j} & r_{f_{0}} & r_{f_{1}} & \ldots & r_{f_{|F|-1}} & 1 \\
1 & r_{j} & r_{f_{0}} & \ldots & r_{f_{|F|-2}} & r_{f_{|F|-1}}
\end{array}\right]+\left[\begin{array}{cc}
q_{n} & 1 \\
1 & r_{n}
\end{array}\right] \rrbracket .
$$

In particular,

$$
\left[\begin{array}{l}
q \\
r
\end{array}\right] \rightarrow \llbracket \sum_{j} s_{j} \rrbracket_{S_{r}}
$$

where $s_{j} \in S_{r}$, and $S_{r}$ is the set of all strings $\left[\begin{array}{lllll}a_{1} & a_{2} & \ldots & a_{l-1} & 1 \\ 1 & b_{2} & \ldots & b_{l-1} & b_{l}\end{array}\right]$ such that $\operatorname{deg}\left(b_{i}\right)<\operatorname{deg}(r)$, for $i<l$, and $\operatorname{deg}\left(b_{l}\right) \leq \operatorname{deg}(r)$.

Proof. At first, observe that

$$
\begin{equation*}
0=\operatorname{deg} r_{1} \leq \ldots \leq \operatorname{deg}\left(r_{i-1}\right)<\operatorname{deg}\left(r_{i}\right)=\ldots=\operatorname{deg}\left(r_{n}\right)=\operatorname{deg}(r) \tag{1.24}
\end{equation*}
$$

In particular, if $i=0,1$ then $r_{j} \in \mathbb{N}$, for all $j$. Indeed, by Lemma 1.5:7, it is $r_{j}=a_{j} \cdot r_{j-1}+r_{j-2}$. Since $\operatorname{deg}\left(a_{i}\right)>0$, we have $\operatorname{deg}\left(r_{i}\right)>\operatorname{deg}\left(r_{i-1}\right)$. Due to maximality of $i$, we get $\operatorname{deg}\left(r_{j}\right)=\operatorname{deg}\left(r_{i}\right)$, for $j \geq i$.

By substitution $y=x+s$ into the cuboid form (1.22) of $\left[\begin{array}{l}q \\ r\end{array}\right]$, we get

$$
\left[\begin{array}{l}
q  \tag{1.25}\\
r
\end{array}\right](x)=\sum_{j=1}^{n} q_{j} \cdot\left(\left[\begin{array}{c}
1 \\
r_{j}
\end{array}\right] \mu_{r_{j+1}, \ldots, r_{n}}(y)\right)-t=\sum_{j=1}^{n} q_{j} \cdot t_{j}(y)-t
$$

where $t \in \mathrm{D}$, and $t_{j}=\left[\begin{array}{c}1 \\ r_{j}\end{array}\right] \mu_{r_{j+1}, \ldots, r_{n}}$.
For $1 \leq j \leq i-1$, we have

$$
t_{j}=\left[\begin{array}{c}
1 \\
r_{j}
\end{array}\right] \mu_{r_{j+1}, \ldots, r_{i-1}} \circ \mu_{r_{i}, \ldots, r_{n}}
$$

By (1.24) and Lemmas $1.5: 18$ and 1.5:19, it is

$$
\begin{aligned}
\mu_{r_{i}, \ldots, r_{n}} & \rightarrow \llbracket \mu_{r_{n}} \rrbracket \rightarrow \llbracket i d-\left[\begin{array}{cc}
r_{n} & 1 \\
1 & r_{n}
\end{array}\right] \rrbracket, \\
{\left[\begin{array}{c}
1 \\
r_{j}
\end{array}\right] \mu_{r_{j+1}, \ldots, r_{i-1}} } & \rightarrow \llbracket \sum_{F \subseteq\{j+1, \ldots, i-1\}}(-1)^{|F|}\left[\begin{array}{ccccc}
r_{f_{0}} & r_{f_{1}} & \ldots & r_{f_{|F|-1}} & 1 \\
r_{j} & r_{f_{0}} & \ldots & r_{f_{|F|-2}} & r_{f_{|F|-1}}
\end{array}\right] \rrbracket_{\sqsubseteq_{1}} .
\end{aligned}
$$

Then, by Lemma 1.5:17 2 b ), we get

$$
\begin{align*}
& t_{j} \rightarrow \llbracket \sum_{F \subseteq\{j+1, \ldots, i-1\}}(-1)^{|F|}\left[\begin{array}{ccccc}
r_{f_{0}} & r_{f_{1}} & \ldots & r_{f_{|F|-1}} & 1 \\
r_{j} & r_{f_{0}} & \ldots & r_{f_{|F|-2}} & r_{f_{|F|-1}}
\end{array}\right]\left(i d-\left[\begin{array}{c}
r_{n} \\
1 \\
1
\end{array} r_{n}\right]\right) \rrbracket_{\sqsubseteq} \rightarrow \\
& \rightarrow \llbracket \sum_{F \subseteq\{j+1, \ldots, i-1\} \cup\{n\}}(-1)^{|F|}\left[\begin{array}{ccccc}
r_{f_{0}} & r_{f_{1}} & \ldots & r_{f_{|F|-1}} & 1 \\
r_{j} & r_{f_{0}} & \ldots & r_{f_{|F|-2}} & r_{f_{|F|-1}}
\end{array}\right] \rrbracket_{\sqsubseteq} \text {. } \tag{1.26}
\end{align*}
$$

For $i \leq j<n$, it is

Finally, for $j=n$, we get

$$
t_{j}=\left[\begin{array}{c}
1  \tag{1.28}\\
r_{n}
\end{array}\right]
$$

Combining (1.26) - (1.28) with (1.25), we obtain

$$
\begin{aligned}
& {\left[\begin{array}{l}
q \\
r
\end{array}\right](x) \rightarrow\left(\sum_{j=1}^{i-1} q_{j} \cdot \llbracket \sum_{F \subseteq\{j+1, \ldots, i-1\} \cup\{n\}}(-1)^{|F|}\left[\begin{array}{rrrl}
r_{f_{0}} & r_{f_{1}} & \ldots & r_{f_{|F|-1}} \\
r_{j} & r_{f_{0}} & \ldots & r_{f_{|F|-2}} \\
r_{f_{|F|-1}}
\end{array}\right]\right]_{\sqsubseteq}+} \\
& \left.+\sum_{j=i}^{n-1} q_{j} \cdot \llbracket 0 \rrbracket\left\{{ }_{i d,}\left[\begin{array}{c}
1 \\
r_{n}
\end{array}\right]\right\}+\left[\begin{array}{cc}
q_{n} & 1 \\
1 & r_{n}
\end{array}\right]\right)(y) \rightarrow \\
& \rightarrow \llbracket\left(\sum_{j=1}^{i-1} \sum_{F \subseteq\{j+1, \ldots, i-1\} \cup\{n\}}(-1)^{|F|}\left[\begin{array}{rrrc}
q_{j} & r_{f_{0}} & \ldots & 1 \\
1 & r_{j} & \ldots & r_{f_{|F|-1}}
\end{array}\right]+\left[\begin{array}{cc}
q_{n} & 1 \\
1 & r_{n}
\end{array}\right]\right)(y) \rrbracket_{\sqsubseteq} \rightarrow \\
& \rightarrow \llbracket\left(\sum_{j=1}^{i-1} \sum_{F \subseteq\{j+1, \ldots, i-1\} \cup\{n\}}(-1)^{|F|}\left[\begin{array}{rrrc}
q_{j} & r_{f_{0}} & \ldots & 1 \\
1 & r_{j} & \ldots & r_{f_{|F|-1}}
\end{array}\right]+\left[\begin{array}{cc}
q_{n} & 1 \\
1 & r_{n}
\end{array}\right]\right)(x) \rrbracket_{\sqsubseteq},
\end{aligned}
$$

where the last arrow is by substitution $x+s$ for $y$ and Lemma 1.5:16 c).
Finally, it is easy to see that if $\alpha \sqsubseteq\left[\begin{array}{cccc}q_{j} & r_{f_{0}} & \ldots & 1 \\ 1 & r_{j} & \ldots & r_{f_{|F|-1}}\end{array}\right]$, for some $j$, then $\tilde{\alpha} \sqsubseteq_{0} \tilde{\beta}$, where $\beta=\left[\begin{array}{cccc}q_{j^{\prime}} & r_{f_{0}} & \ldots & 1 \\ 1 & r_{j^{\prime}} & \ldots & r_{f_{|F|-1}}\end{array}\right]$, for some $j^{\prime}$. Therefore, in the last expression, the symbol $\sqsubseteq$ can be replaced by $\sqsubseteq_{0}$.

The "in particular" follows directly from (1.24).

The following lemma is the first step in the inductive proof of Proposition H 1.5:3.

Lemma 1.5:21. Let $a_{i}, b_{i}, i=1, \ldots, n$ be scalars such that $\operatorname{deg}\left(b_{i}\right)=0$, for all $i$. Then

$$
\left[\begin{array}{ccc}
a_{1} & \ldots & a_{n} \\
b_{1} & \ldots & b_{n}
\end{array}\right] \rightarrow \llbracket\left[\begin{array}{cc}
\prod_{i=1}^{n} a_{i} & 1 \\
1 & \prod_{i=1}^{n} b_{i}
\end{array}\right] \rrbracket_{\sqsubseteq_{1}} .
$$

Proof. Set $p_{j}=\prod_{i=1}^{j} b_{i} \in \mathbb{N}, g_{j}=\prod_{i=1}^{j} a_{i}$. Denote $f_{n}=\left[\begin{array}{ccc}a_{1} & \ldots & a_{n} \\ b_{1} & \ldots & b_{n}\end{array}\right]$. We prove that

$$
f_{n}\left(x+p_{n} k\right)=f_{n}(x)+g_{n} k
$$

by induction on $n$. If $n=1$, it is clear.
For the induction step, we have

$$
\begin{aligned}
f_{n}\left(x+p_{n} k\right) & =f_{n-1}\left(\left[\begin{array}{c}
a_{n} \\
b_{n}
\end{array}\right](x)+p_{n-1} a_{n} \cdot k\right)= \\
& =f_{n-1}\left(\left[\begin{array}{c}
a_{n} \\
b_{n}
\end{array}\right](x)\right)+g_{n-1} \cdot a_{n} \cdot k= \\
& =f_{n}(x)+g_{n} k
\end{aligned}
$$

where the first equality is due to $b_{n} \mid p_{n}$, and the second one is by the induction assumption.

Now,

$$
\begin{aligned}
f_{n}(x) & =f_{n}\left(\mu_{p_{n}}(x)+p_{n} \cdot\left[\begin{array}{c}
1 \\
p_{n}
\end{array}\right](x)\right)=f_{n}\left(\mu_{p_{n}}(x)\right)+g_{n} \cdot\left[\begin{array}{c}
1 \\
p_{n}
\end{array}\right](x)= \\
& =\left\{\left[\begin{array}{ll}
g_{n} & 1 \\
1 & p_{n}
\end{array}\right](x)+f_{n}(i) \text { if } \mu_{p_{n}}(x)=i ; i=0, \ldots, p_{n}-1\right.
\end{aligned}
$$

Now, we are ready for our proof of Proposition H 1.5:3.
Proposition 1.5:3 (Proposition H). Every term $t(\bar{x})$ is equivalent to a harmonic almost-term $\tau(\bar{x})$.

Moreover, if a variable $x$ does not occur in any subterm of the form $r^{-1}$ s (where $s$ is a term) in $t$ then the same is true in $\tau$.

Proof. Let $t$ be a term. By Lemma 1.5:16 d), $t$ is equivalent to a distributed almost-term $\sigma$. We may assume that $\sigma$ does not contain binary minus nor any symbol $\mu_{r}$. Then it is, clearly, enough to prove that every string

$$
\alpha=\left[\begin{array}{lll}
a_{1} & \ldots & a_{n}  \tag{1.29}\\
b_{1} & \ldots & b_{n}
\end{array}\right]
$$

is equivalent to an almost-term in harmonic form.
Let assign to any string $\alpha$ of the form (1.29) the triple

$$
I_{\alpha}=\left(d_{\alpha}, N_{\alpha}, K_{\alpha}\right)
$$

where $d_{\alpha}=\max \operatorname{deg} b_{i}, N_{\alpha}=\left|\left\{i ; \operatorname{deg} b_{i}=d_{\alpha}\right\}\right|$, and $K_{\alpha}=\min \left\{n-i ; \operatorname{deg} b_{i}=d_{\alpha}\right\}$. We prove the previous statement by induction on $I_{\alpha} \in\left\langle\mathbb{N}^{3}, \leq_{\text {Lex }}\right\rangle$. We will denote the only free variable in $\alpha$ as $x$, but we will often omit writing it.

If $d_{\alpha}=0$, the statement follows from Lemma 1.5:21. Otherwise, let $\alpha$ be a string as in (1.29). Denote $d_{\alpha}, N_{\alpha}, K_{\alpha}$ just $d, N, K$, and set $J=n-K$ (then $b_{J}$ is the rightmost $b_{i}$ with $\operatorname{deg} b_{i}=d$ ). Then

$$
\left[\begin{array}{c}
a_{J} \\
b_{J}
\end{array}\right] \rightarrow \llbracket \sum_{j} s_{j} \rrbracket_{S_{b_{J}}}
$$

with $s_{j} \in S_{b_{J}}$, according to Lemma 1.5:20, By Lemmas 1.5:16 and 1.5:17, then

$$
\alpha \rightarrow \llbracket \sum_{j}\left[\begin{array}{ccc}
a_{1} & \ldots & a_{J-1} \\
b_{1} & \ldots & b_{J-1}
\end{array}\right] \circ s_{j} \circ\left[\begin{array}{ccc}
a_{J+1} & \ldots & a_{n} \\
b_{J+1} & \ldots & b_{n}
\end{array}\right] \rrbracket_{\sqsubseteq} .
$$

Denote the $j$-th summand in the previous expression as $\alpha_{j}$, and let $\beta \sqsubseteq \alpha_{j}$. We complete the proof by showing that $\beta$ is equivalent to a string $\beta^{\prime}$ with $I_{\beta^{\prime}}<I_{\alpha}$.

By the definition of $S_{b_{J}}$, it is $d_{s_{j}} \leq \operatorname{deg} b_{J}=d_{\alpha}$. If $d_{s_{j}}<d_{\alpha}$ then, clearly, also $I_{\beta} \leq I_{\alpha_{j}}<I_{\alpha}$. Otherwise, $s_{j}=\left[\begin{array}{ccc}u_{1} & \ldots & 1 \\ 1 & \ldots & v_{l}\end{array}\right]$, where $\operatorname{deg} v_{i}<d_{\alpha}$, for $i<l$, and $\operatorname{deg} v_{l}=d_{\alpha}$. Suppose that $J<n$. Then

$$
\alpha_{j}=\left[\begin{array}{ccccccccc}
a_{1} & \ldots & a_{J-1} & u_{1} & \ldots & 1 & a_{J+1} & \ldots & a_{n} \\
b_{1} & \ldots & b_{J-1} & 1 & \ldots & v_{l} & b_{J+1} & \ldots & b_{n}
\end{array}\right]
$$

and, since $\left[\begin{array}{cc}1 & a_{J+1} \\ v_{l} & b_{J+1}\end{array}\right]=\left[\begin{array}{c}a_{J+1} \\ v_{l} \cdot b_{J+1}\end{array}\right]=\left[\begin{array}{cc}1 & a_{J+1} \\ b_{J+1} & v_{l}\end{array}\right]$, any $\beta \sqsubseteq \alpha_{j}$ is equivalent to some $\beta^{\prime}$ such that

$$
\beta^{\prime} \sqsubseteq \alpha_{j}^{\prime}=\left[\begin{array}{ccccccccc}
a_{1} & \ldots & a_{J-1} & u_{1} & \ldots & 1 & a_{J+1} & \ldots & a_{n} \\
b_{1} & \ldots & b_{J-1} & 1 & \ldots & b_{J+1} & v_{l} & \ldots & b_{n}
\end{array}\right],
$$

or

$$
\beta^{\prime} \sqsubseteq\left[\begin{array}{ccc}
a_{J+1} & \ldots & a_{n} \\
b_{J+1} & \ldots & b_{n}
\end{array}\right] .
$$

Again, it is easy to verify that $I_{\beta^{\prime}} \leq I_{\alpha_{j}^{\prime}}<I_{\alpha}$. Finally, if $J=n$ then

$$
\alpha_{j}=\left[\begin{array}{cccccc}
a_{1} & \ldots & a_{J-1} & u_{1} & \ldots & 1 \\
b_{1} & \ldots & b_{J-1} & 1 & \ldots & v_{l}
\end{array}\right]=\left[\begin{array}{cccccc}
a_{1} & \ldots & a_{J-1} & u_{1} & \ldots & u_{l-1} \\
b_{1} & \ldots & b_{J-1} & 1 & \ldots & v_{l-1}
\end{array}\right](y),
$$

where $y=y^{(j)}=\left[\begin{array}{c}1 \\ v_{l}\end{array}\right]=\left[\begin{array}{c}1 \\ v_{l}^{(j)}\end{array}\right]$, and any $\beta \sqsubseteq \alpha_{j}$ is equivalent to some $\beta^{\prime}(y)$ such that

$$
\beta^{\prime} \sqsubseteq \alpha_{j}^{\prime}=\left[\begin{array}{cccccc}
a_{1} & \ldots & a_{J-1} & u_{1} & \ldots & u_{l-1} \\
b_{1} & \ldots & b_{J-1} & 1 & \ldots & v_{l-1}
\end{array}\right] .
$$

Clearly, $I_{\beta^{\prime}} \leq I_{\alpha_{j}^{\prime}}<I_{\alpha}$.
By the induction assumption, $\alpha$ is equivalent to an almost-term $\tau(x, \bar{y})$ in harmonic form. After substituting $y^{(j)}=\left[\begin{array}{c}1 \\ v_{l}^{(j)}\end{array}\right]$, we get a term $\tau^{\prime}(x)$ in harmonic form, equivalent to $\alpha$.

### 1.5.7 Bases

In this subsection, we prove Proposition B 1.5:4. The proposition implies that every p-term or open formula can be on any interval $[Q, R]$, with $Q, R \in \mathrm{D}$, equivalently written over a multiple of a given basis. Bases $\left\langle b_{i}\right\rangle_{i<\omega}$ with $b_{i} \mid b_{i+1}$ will be of special importance, in particular, for the proof of Proposition S 1.5:2,

We will need the following lemmas.
Lemma 1.5:22. Let $0<R<r \in \mathrm{D}, 0 \neq n \in \mathbb{N}$. Then
a) $r^{-1} x=\left\{n \cdot(n r)^{-1}(x)+i\right.$ if $i r \leq \mu_{n r}(x)<(i+1) r ; i=0, \ldots, n-1$,
b) $r^{-1} x=\left\{\begin{array}{cl}{\left[\begin{array}{ll}R & 1 \\ r & R\end{array}\right](x)} & \text { if } r\left(\left[\begin{array}{ll}R & 1 \\ r & R\end{array}\right](x)+1\right) \geq x, \\ {\left[\begin{array}{ll}R & 1 \\ r & R\end{array}\right](x)+1} & \text { otherwise. }\end{array}\right.$

Proof. Direct computation.
Lemma 1.5:23. Let $0<R, r \in \mathrm{D}$ such that $\operatorname{deg} R \leq \operatorname{deg} r$, and $0 \neq n \in \mathbb{N}$.
a) There is a harmonic almost-term $\tau(x)$ over $\{n r\}$, equivalent to $r^{-1} x$.
b) There is a harmonic almost-term $\tau(x)$ over the set $\left\{r^{\prime} ; R \mid r^{\prime}\right\}$, equivalent to $r^{-1} x$.

Proof. a) Directly from Lemma 1.5:22 a).
b) By a), we may suppose that $R<r$ (there is $n \in \mathbb{N}$ such that $n r>R$ ). Denote $\beta=R^{-1} x$. Then, by Lemma 1.5:22 a), we have:

$$
r^{-1} x=\left\{\begin{array}{cl}
{\left[\begin{array}{l}
R \\
r
\end{array}\right](\beta)} & \text { if } r \cdot\left(\left[\begin{array}{l}
R \\
r
\end{array}\right](\beta)+1\right) \geq x \\
{\left[\begin{array}{l}
R \\
r
\end{array}\right](\beta)+1} & \text { otherwise }
\end{array}\right.
$$

Consider now $\beta$ as a variable. By Proposition H 1.5:3, there is an almost-term $\tau^{\prime}(\beta, x)$ in harmonic form, equivalent to $r^{-1} x$ and such that $\operatorname{Div}_{x}\left(\tau^{\prime}\right)=\emptyset$. After substituting $\beta=R^{-1} x$ into $\tau^{\prime}$, we get an almost-term $\tau(x)$ in harmonic form such that each $r^{\prime} \in \operatorname{Div}_{x}(\tau)$ is a multiple of $R$.

Now, we are ready for a proof of Proposition B 1.5:4.
Proposition 1.5:4 (Proposition B). Let $\delta \in F, p, r \in{ }^{+}$D be scalars, $e=\operatorname{deg}(p)$, $d=\operatorname{deg}(r)$, and $B=\left\langle b_{i}\right\rangle_{d \leq i<e}$ be a $[d, e]$-basis. Then $r^{-1} x$ is on $[\delta, \delta+p-1]$ equal to a harmonic almost-term $\tau(x)$ (possibly with parameter $\delta$ ) which is over $m B=\left\langle m b_{i}\right\rangle_{d \leq i<e}$, for some $m \in \mathbb{N}$.

Moreover, $m$ can be chosen as any number sufficiently large with respect to divisibility.

Proof. Observe, first, that it is enough to prove the statement for $\delta=0$. Indeed, $r^{-1}(y+\delta)$ is, by Lemma 1.5:6 e), equivalent to an almost-term $\sigma(y)$ with $\operatorname{Div}(\sigma)=$ $\{r\}$, and hence, by the proposition's case $\delta=0$, on $[0, p-1]$ equivalent to a harmonic almost-term $\tau^{\prime}(y)$, with $\operatorname{Div}_{y}\left(\tau^{\prime}\right) \subseteq\left\{m \cdot b_{i} ; d \leq i<e\right\}$, for a given $m$. By substitution $y=x-\delta$, we then have $r^{-1} x$ equivalent to $\tau^{\prime}(x-\delta)$ on $[\delta, \delta+p-1]$ and $\tau^{\prime}(x-\delta)$ (again by Lemma 1.5:6 e)) equivalent to an almost-term $\tau(x)$ in harmonic form, with $\operatorname{Div}_{x}(\tau)=\operatorname{Div}_{y}\left(\tau^{\prime}\right) \subseteq\left\{m \cdot b_{i} ; \operatorname{deg} r \leq i<e\right\}$.

Further, suppose $\delta=0$. It is enough to prove that $r^{-1} x$ is on $[0, p-1]$ equivalent to a harmonic almost-term $\tau_{r}^{\prime}(x)$ with

$$
\begin{equation*}
\operatorname{Div}\left(\tau_{r}^{\prime}\right) \subseteq\left\{m^{\prime} \cdot b_{i} ; 0<m^{\prime} \in \mathbb{N}, \operatorname{deg} r \leq i<e\right\} \tag{1.30}
\end{equation*}
$$

Then we are done almost immediately: We choose $m_{S}$ to be the least common multiple of all such $m^{\prime} \in \mathbb{N}$ that $m^{\prime} b_{i} \in \operatorname{Div}\left(\tau_{r}^{\prime}\right)$, for some $i$. Now, if $0<m \in \mathbb{N}$ is a multiple of $m_{S}$ then $\tau_{r}^{\prime}(x)$ is equivalent to a harmonic almost-term $\tau_{r}(x)$ with $\operatorname{Div}\left(\tau_{r}\right) \subseteq\left\{m \cdot b_{i} ; \operatorname{deg} r \leq i<e\right\}$, according to Lemma 1.5:23 a).

Now, we find $\tau_{r}^{\prime}$ such that (1.30) holds. We proceed by backwards induction on $d$. When $d \geq e$, there is $n \in \mathbb{N}$ such that $n r>p$. Then $r^{-1} x$ is on $[0, p-1]$ equivalent to

$$
\tau_{r}^{\prime}(x)=\{l \text { if } l r \leq x<(l+1) r ; l=0, \ldots, n-1
$$

Suppose that $d<e$ and that the statement holds for all $r^{\prime}$ with $\operatorname{deg} r^{\prime}>d$. Then $r^{-1} x$ is, by Lemma $1.5: 23 \mathrm{~b}$ ), equivalent to an almost-term $\sigma(x)$ with all $r^{\prime} \in \operatorname{Div}(\sigma)$ divisible by $b_{d} \in B$. For $r^{\prime} \in \operatorname{Div}(\sigma)$, it is either $r^{\prime}=m^{\prime} b_{d}$, for some $m^{\prime} \in \mathbb{N}$, or $\operatorname{deg} r^{\prime}>d$. Denote $\tau_{r}^{\prime}$ the almost-term created from $\sigma$ by replacing all $r^{\prime-1} x$ with $\operatorname{deg} r^{\prime}>d$ by the respective almost-terms $\tau_{r^{\prime}}^{\prime}$ from the induction assumption; $\tau_{r}^{\prime}$ has, clearly, the demanded properties.

### 1.5.8 Solvability

Using the Proposition B 1.5:4 for $B=\left\langle b_{i}\right\rangle_{i<\omega}$ with $b_{i} \mid b_{i+1}$, we may now prove the last of the three main propositions - Proposition S 1.5:2,

### 1.5.8.1 Linear period and linear growth

A scalar $p \in \mathrm{D}$ is called a linear period of a term $t(x, \bar{y})$ in (a variable) $x$ if $t(p u+v, \bar{y})$ is affine in $u$, for every $v, \bar{y} \in F$, i.e. for every $v, \bar{y} \in F$, there is $\gamma_{x, p}=\gamma_{x, p}(v, \bar{y}) \in \mathrm{D}$ such that

$$
\begin{equation*}
t(p u+v, \bar{y})=\gamma_{x, p} \cdot u+t(v, \bar{y}) \tag{1.31}
\end{equation*}
$$

holds for all $u \in F$. It is not hard to see that the linear growth $\gamma_{x, p}(v, \bar{y})$ does not depend on $v$ nor $\bar{y}$; we denote it $\gamma_{x, p}(t)$ (this follows from the easy observation that if $P$ is the product of all occurances of $x$-divisors in $s(x, v, \bar{y})$ then $s(P u, v, \bar{y})=$ $s^{\prime}(P u)+s^{\prime \prime}(v, \bar{y})$, for some terms $\left.s^{\prime}, s^{\prime \prime}\right)$.

We say that $p \in \mathrm{D}$ is a linear period of an open formula $\psi[\mathrm{p}$-term $\tau]$ if it is a linear period of every maximal subterm of $\psi[\tau]$.

### 1.5.8.2 Balanced form

We say that an open harmonic formula $\psi$ is balanced [in $x$ ] (in a balanced form $[$ in $x]$ ) if there is an enumeration $\left\langle r_{i}\right\rangle_{i<n}$ of $\operatorname{Div}(\psi)\left[\operatorname{Div}_{x}(\psi)\right]$ such that $r_{i} \mid r_{i+1}$. Similarly, we define balanced form for an open harmonic p-term. If $\psi$ is harmonic and balanced in $x$ then, clearly, the maximal $x$-divisor in $\psi$ is a linear period of $\psi$ in $x$.

Lemma 1.5:24. Let $\psi(v, \bar{y})$ be an open formula, and $0<P \in \mathrm{D}$. Then there is an open harmonic $\chi(v, \bar{y})$, having a linear period $p$ in $v$ with $\operatorname{deg}(p)<\operatorname{deg}(P)$, and such that

$$
(\forall \bar{y})(\forall 0 \leq v<P)(\psi(v, \bar{y}) \leftrightarrow \chi(v, \bar{y})) .
$$

Proof. We apply the Proposition B $1.5: 4$ to (every maximal subterm of) $\psi$, the interval $[0, P]$ and a $[0, \operatorname{deg}(P)]$-basis $B=\left\langle b_{i}\right\rangle_{i<\operatorname{deg}(P)}$ such that $b_{i} \mid b_{i+1}$. The resulting formula $\chi$ is harmonic and over $m B$ in $v$, for some $m \in \mathbb{N}$, hence balanced in $v$, and therefore its maximal $v$-divisor $p(p=1$ if there are no $v$ divisors) is a linear period of $\chi$ in $v$. It is $p=m b_{i}$, for some $i<\operatorname{deg}(P)$, and thus $\operatorname{deg}(p)=i<\operatorname{deg}(P)$.

The following lemma proves Proposition $\mathrm{S} 1.5: 2$ for the case $\operatorname{Div}_{x}(\psi)=\emptyset$. The algorithm it contains is known as the Fourier-Motzkin elimination.

Lemma 1.5:25 ("Fourier-Motzkin elimination"). Suppose that $\psi(x, \bar{y})$ is open and such that $\operatorname{Div}_{x}(\psi)=\emptyset$. Then there are finitely many terms $t_{j}(\bar{y}), j<n \in \mathbb{N}$, such that

$$
(\exists x) \psi(x, \bar{y}) \leftrightarrow \bigvee_{j<n} \psi\left(t_{j}, \bar{y}\right)
$$

Proof. Without loss of generality, we can assume that $\psi$ is a system of linear inequalities of the form $a_{i} x-s_{i} \leq 0$, where $a_{i} \in \mathrm{D}$ and $s_{i}$ are terms with all their variables among $\bar{y}$. Denote $I^{+}\left[I^{-}, I^{0}\right]$ the set of all indices $i$ for witch $a_{i}>0$ $\left[a_{i}<0, a_{i}=0\right]$. Then

$$
\psi \leftrightarrow \bigwedge_{i \in I^{0}} s_{i} \leq 0 \& \bigwedge_{i \in I^{-}} x \geq \frac{s_{i}}{a_{i}} \& \bigwedge_{j \in I^{+}} x \leq \frac{s_{j}}{a_{j}}
$$

If both $I^{+}$and $I^{-}$are empty then we can set $n=1$ and $t_{0}=0$. Suppose that $I^{+} \neq \emptyset$ (the other case is symetric). Then

$$
(\exists x) \psi(x, \bar{y}) \leftrightarrow \bigvee_{j \in I^{+}} \psi\left(a_{j}^{-1} s_{j}, \bar{y}\right)
$$

Now, we are going to prove Proposition S 1.5:2,
Proposition 1.5:2 (Proposition S). Let $\psi(x, \bar{y})$ be an open formula. There are finitely many terms $t_{i}(\bar{y}), i<n$, such that

$$
(\exists x) \psi \leftrightarrow \bigvee_{i<n} \psi\left(t_{i}, \bar{y}\right)
$$

Proof. The case $\operatorname{Div}_{x}(\psi)=\emptyset$ follows immediately from Lemma 1.5:25; assume further $\operatorname{Div}_{x}(\psi) \neq \emptyset$. We may also suppose that $\psi$ is harmonic (thanks to Proposition H 1.5:3). Let $P$ be a linear period of $\psi$ in $x$ of the least degree; we prove the statement by induction on $\operatorname{deg}(P)$.

We denote $\tilde{\psi}(u, v, \bar{y})$ the formula created from $\psi$ by replacing its each maximal subterm $t(x, \bar{y})$ with $\gamma_{x, P}(t) \cdot u+t(v, \bar{y})$. By (1.31) we have

$$
\begin{equation*}
\psi(P u+v, \bar{y}) \leftrightarrow \tilde{\psi}(u, v, \bar{y}) . \tag{1.32}
\end{equation*}
$$

Since $\operatorname{Div}_{u}(\tilde{\psi})=\emptyset$, by Lemma 1.5:25, there are terms $t_{j}^{\prime}(v, \bar{y}), j<n^{\prime}$, such that

$$
(\exists x) \psi(x, \bar{y}) \leftrightarrow(\exists 0 \leq v<P) \bigvee_{j<n^{\prime}} \tilde{\psi}\left(t_{j}^{\prime}, v, \bar{y}\right) \leftrightarrow(\exists 0 \leq v<P) \bigvee_{j<n^{\prime}} \psi\left(t_{j}^{\prime \prime}, \bar{y}\right)
$$

where $t_{j}^{\prime \prime}(v, \bar{y})=P \cdot t_{j}^{\prime}+v$. Here, the first equivalence follows from (the unique) representability of $x$ as $x=P u+v$ with $0 \leq v<P$, and the second equivalence from (1.32).

If $\operatorname{deg}(P)=0$ then the last formula is equivalent to $\bigvee_{j<n^{\prime}, m<P} \psi\left(t_{j, m}, \bar{y}\right)$, where $t_{j, m}(\bar{y})=t_{j}^{\prime \prime}(\underline{m}, \bar{y})$, which is what we wanted to prove.

Let $\operatorname{deg}(P)>0$. By Lemma 1.5:24, there is $\chi(v, \bar{y})$ with a linear period $p$ in $v$ such that $\operatorname{deg}(p)<\operatorname{deg}(P)$ and

$$
\bigvee_{j<n^{\prime}} \psi\left(t_{j}^{\prime \prime}, \bar{y}\right) \leftrightarrow \chi(v, \bar{y})
$$

for $0 \leq v<P$. Denote $\chi^{\prime}$ the formula $\chi \&(0 \leq v<P)$. By the induction assumption, there are terms $s_{i}(\bar{y}), i<k$, such that

$$
(\exists x) \psi(x, \bar{y}) \leftrightarrow(\exists v) \chi^{\prime}(v, \bar{y}) \leftrightarrow \bigvee_{i<k} \chi^{\prime}\left(s_{i}, \bar{y}\right)
$$

Denote $t_{i, j}(\bar{y})=t_{j}^{\prime \prime}\left(s_{i}(\bar{y}), \bar{y}\right)$. Then

$$
(\exists x) \psi(x, \bar{y}) \leftrightarrow \bigvee_{i<k, j<n^{\prime}} \psi\left(t_{i, j}, \bar{y}\right)
$$

### 1.6 Two-sorted solvability

The proof of Theorems $1.3: 4,1.3: 6$ and $1.3: 8$ from section 1.5 proves, in fact, more than we stated so far: For example, in the statement of Proposition S $1.5: 2$ (and consequently also in Theorem 1.3:4), all scalars which occur in terms $t_{i}$ can be constructed from scalars occuring in $\psi$ by ring operations and integer division.

This observation may be formulated as solvability (or quantifier elimination) for so called doded-modules - structures in a two-sorted language of the type "ordered ring-ordered module" (i.e. in a language with a special sort for scalar variables), which are just two-sorted variants of models of linear theories. This is stated in Theorem 1.6:1 and its Corollary 1.6:2,

These results are an ordered analogues of the quite well-known results by Lou van den Dries and Jan Holly in vdDH92 for two-sorted unordered modules and strengthen the result by Volker Weispfenning in Wei97, Theorem 4.1] for twosorted discretely ordered modules over the ring $\mathbb{Z}$ of integers (more precisely for the models of a two-sorted variant of Presburger arithmetic).

The results from vdDH92 are generalized by adding an ordering to the language but for a price of restricting ourselves only to modules over rings which are dodeds (see 1.3.1.2) (and adding another order-related conditions). Let us note that in vdDH92] the problem of generalizing the results to ordered modules
(even for the simplest case of the module $\mathbb{Z}$ of integers) is considered as "very interesting" but as one that "seems to be very hard".

The Weispfenning's result is strengthened in two directions: first, we admit the universe of the ring sort to be not only the ring $\mathbb{Z}$ of integers but an arbitrary doded (see Example 1.3:1 for examples of dodeds); second, we give a strictly smaller elimination set of formulas than Weispfenning's "scalar bounded formulas" (see Remark 1.6:3).

Let $\mathbb{L}$ denote the two-sorted language of the type "ordered ring-ordered module", i.e. $\mathbb{L}$ consists of

- the ordered ring sort $R$ with a language $\langle 0,1,+,-, \cdot, \leq\rangle$,
- the ordered group sort $M$ with a language $\langle 0,1,+,-, \leq\rangle$ (the symbol 1 is intended for the least positive element),
- a binary function symbol $\cdot: R \times M \rightarrow M$ (scalar multiplication).

Further on, unless stated otherwise, symbols $x, y, z$ denote $M$-variables while symbols $p, q, r$ stand for $R$-variables. We also refer to $R$-variables as scalar variables and to quantification of those variables as scalar quantification.

A doded-module is a (two-sorted) $\mathbb{L}$-structure $\mathcal{A}=\langle\mathcal{R}, \mathcal{M}, \cdot\rangle$ such that

1) $\mathcal{R}$ is a doded (see 1.3.1.2),
2) $\langle\mathcal{M}, r \cdot\rangle_{r \in R}$ is a discretely ordered (with 1 being the least positive element), integrally-divisible (i.e. $(\forall x)(\exists y)(\exists 0 \leq z<r \cdot 1)(x=r \cdot y+z)$ holds) $\mathcal{R}$-module.

We denote $\mathbb{L}^{\prime}$ the extension of $\mathbb{L}$ by

- a binary function symbol ${ }^{-1}: R^{2} \rightarrow R$ (scalar integer-division),
- a binary function symbol ${ }^{-1}: R \times M \rightarrow M$ (integer-division).

Usually, we write $r^{-1} q$ or $r^{-1} x$ instead of ${ }^{-1}(r, q)$ or ${ }^{-1}(r, x)$.
For a doded-module $\mathcal{A}$, we write $\mathcal{A}^{\prime}$ for its $\mathbb{L}^{\prime}$-expansion by definitions:

- $0 \leq q-r \cdot\left(r^{-1} q\right)<r$,
- $0 \leq x-r \cdot\left(r^{-1} x\right)<r \cdot 1$.

Now, we are ready to formulate the main result of this section. The following theorem may be understood as stating that every doded-module is " $M$-almost uniformly $M$-solvable".

Theorem 1.6:1. Let $\mathcal{A}=\langle\mathcal{R}, \mathcal{M}, \cdot\rangle$ be a doded-module, $\varphi(\bar{r}, \bar{y}, x)$ be an $\mathbb{L}^{\prime}$ formula without scalar quantifiers, and $\bar{\rho} \in R^{l(\bar{r})}$ be scalars. Then there are finitely many $\mathbb{L}^{\prime}$-terms $t_{i}(\bar{r}, \bar{y})$, for $i<n$, with $n \in \mathbb{N}$, such that

$$
\mathcal{A}^{\prime} \models(\exists x) \varphi(\bar{\rho}, \bar{y}, x) \leftrightarrow \bigvee_{i<n} \varphi\left(\bar{\rho}, \bar{y}, t_{i}(\bar{\rho}, \bar{y})\right)
$$

Proof. First observe that, for fixed scalars $\bar{\rho} \in R^{l(\bar{r})}$, the formula $\varphi(\bar{\rho}, \bar{y}, x)$ can be written naturally as a formula in the language of the (one-sorted) lineal which corresponds to the (two-sorted) doded-module $\mathcal{A}$.

Then it is enough to check that the proof of Proposition S 1.5:2 (which occupies most of section (1.5) constructs terms $t_{i}$ which contain only scalars expressible from the scalars occuring in $\psi$ only by ring operations and the operation ${ }^{-1}$ of scalar integer division (i.e. $\mathbb{L}^{\prime}$-scalar-operations).

This is obvious with only two exceptions:
a) In Lemma 1.5:24, we used an arbitrary balanced basis $B$. By doing that we can only assure scalars in formula $\chi$ to be expressible by $\mathbb{L}^{\prime}$-scalar-operations from the scalars occuring in $\psi$ and scalars from the basis $B$.
b) In the final part of the proof of Proposition S 1.5:2, we defined $P$ to be the linear period of $\psi$ of the least degree. Again, it is not clear that such $P$ is expressible by $\mathbb{L}^{\prime}$-scalar-operations from the scalars occuring in $\psi$.

Both these problems can be easily resolved:
a) Instead of taking an arbitrary balanced basis and applying Proposition B 1.5:4, we may use the backwards induction idea from the proof of Proposition B directly with a simple modification: At the induction step do not use Lemma $1.5: 23 \mathrm{~b}$ ) to get all divisors be divisible by $b_{d}$ but use the same lemma to get all divisors be divisible by the maximal divisor $q$ such that all divisors $q^{\prime} \leq q$ are up to multiplication by some $m \in \mathbb{N}$ linearly ordered by divisibility.

Then we get a balanced formula $\chi$ with all scalars expressible by $\mathbb{L}^{\prime}$-scalaroperations from the scalars occuring in $\psi$.
b) It is enough to take $P$ to be the least common multiple of all $x$-divisors in $\psi$.

Corollary 1.6:2. Let $\mathcal{A}$ be a doded-module. Every scalar-quantifier-free $\mathbb{L}^{\prime}$ formula is in $\mathcal{A}^{\prime}$ equivalent to a quantifier-free $\mathbb{L}^{\prime}$-formula.
Remark 1.6:3. To compare Corollary 1.6:2 with Wei97, Theorem 4.1], let us note that any quantifier-free $\mathbb{L}^{\prime}$-formula can be easily equivalently rewritten as a scalar bounded (i.e. with all quantifiers of the form $(\exists x,|x| \leq r \cdot 1)$ with $r$ a scalar term) formula in the language $\mathbb{L} \cup\left\langle^{-1}, \equiv\right\rangle$, where ${ }^{-1}$ is the scalar integer-division, and $\equiv$ is a ternary $R \times M^{2}$ congruence relation, defined as $x \equiv_{r} y \leftrightarrow(\exists z)(r \cdot z=x-y)$.

Indeed, this is easy since $r^{-1} x=y \leftrightarrow(\exists z,|z| \leq(r-1) \cdot 1)\left(z \equiv_{r} x \& y=\frac{x-z}{r}\right)$.

## Chapter 2

## Structure of Peano Products

In this chapter, we deal with a problem of understanding relations between local and global properties of an operation $o$ in a first-order structure of the form $\langle\mathcal{B}, o\rangle$, with a particular interest in the case where $\mathcal{B}$ is a model of Presburger arithmetic $\operatorname{Pr}$ and $o$ is a "Peano product" on $\mathcal{B}$, i.e. $\langle\mathcal{B}, o\rangle$ is a model of Peano arithmetic P .

This problem may be specified as follows: Given a "background model" $\mathcal{B}$ and a set $O$ of all $n$-ary operations on $B$ satisfying certain global property (e.g. being a Peano product), we want to describe the dependency closure

$$
\operatorname{icl}^{O}(E)=\left\{\bar{d} \in B^{n} ;\left(\forall o, o^{\prime} \in O\right)\left(o \upharpoonright E=o^{\prime} \upharpoonright E \Rightarrow o(\bar{d})=o^{\prime}(\bar{d})\right)\right\}
$$

for $E \subseteq B^{n}$. We call this task the $(\mathcal{B}, O, E)$-dependency problem.
Illustratively speaking, a point $\bar{d}$ lies in the dependency closure icl ${ }^{O}(E)$ of the set $E$ if the value $o(\bar{d})$ of any operation $o \in O$ is uniquely determined by its values on $E$.

Our particular interest is in the Peano dependency problem - a $(\mathcal{B}, O, E)$ dependency problem where $\mathcal{B}$ is a model of $\operatorname{Pr}$ and $O$ is the set of all (saturated) Peano products on $\mathcal{B}$. Both, the general dependency problem and the Peano dependency problem, may be further modified and specified by various modifications of the dependency closure (see the more general definition [2.1.1).

A $(\mathcal{B}, O, E)$-dependency problem with saturated $\mathcal{B}$ may be solved by studying a definability problem (regarding almost-uniform definability; see 2.1.3) in certain expansion of $\mathcal{B}$, called a fixator (2.1.4). This is formulated in the DD-theorem 2.1:2.

In Proposition 2.2:1, we completely solve two important cases of the Peano dependency problem - for $E=\emptyset$ (which is easy) and for $E=E_{a}=\{a\} \times B$, with $a$ nonstandard (an " $a$-slice"). We prove that, in these cases, $\operatorname{icl}(E)$ is as small as possible, i.e. it contains only the trivially dependent points (for $E=E_{a}$ that are points $\bar{d}=\left(d_{0}, d_{1}\right)$ where at least one of $d_{i}$ equals $p(a)$, for some polynomial $p \in \mathbb{Q}[x])$.

By the DD-theorem, the key for the proof is understanding definability in the respective fixators. The fixators are models of Presburger arithemetic $\operatorname{Pr}$ (1.1.3.1) and linear arithmetic LA (1.1.4.1), respectively. We use here the descriptions of elimination sets of formulas for the fixators which we provided in Corollaries 1.4:2 and 1.4:7, respectively, in chapter 1.

An important special case of Proposition 2.2:1 proves the existence of pairs of Peano products $(\cdot, \circ)$ which coincide on an " $a$-slice" $E_{a}=\{a\} \times B$, with $a \in B-\mathbb{N}$, but differ in some $\bar{d}<a$ and in some $\overline{d^{\prime}}>a$. We call such a couple a meeting pair of Peano products. By "piecing together" a meeting pair, it is possible to obtain a new "Robinson product" on $\mathcal{B}$, which satisfies certain portion of induction. In section [2.3, we put these ideas into the context of possible further research on constructions of models of Peano arithmetic.

Finally, section 2.4 contains a summary of our partial results regarding the problem of interpolating a given set of points in $B^{3}$ by the graph of some Peano product.

### 2.1 Dependency and definability

Given a saturated structure $\mathcal{B}$ (background-model) and a set $O$ of $n$-ary operations on $B$, we want to know whether, for an operation $o \in O$, its value in a point $\bar{d} \in B^{n}$ is determined by its values on a set $E \subseteq B^{n}$; this question is precised in a concept of dependency in 2.1.1. The main result of this section, the DD-theorem 2.1:2, provides an equivalent for dependency in terms of definability in an expansion $\mathcal{A}$ of $\mathcal{B}$, called fixator (see 2.1.4 for definition).

Throughout this section, $\mathcal{A}$ denotes a saturated expansion of a backgroundmodel $\mathcal{B}, n>0$ is an integer, $\bar{d} \in B^{n}$ and $E \subseteq B^{n}$. Further, all considered operations on $A=B$ are $n$-ary.

### 2.1.1 Dependency and marriages

Let $\sim$ be an equivalence relation on a set $O$ of $n$-ary operations on $B, o \in O$. We say that a point $\bar{d} \in B^{n} \sim$-depends on $E \subseteq B^{n}$ [for o] if, for $o^{\prime}, o^{\prime \prime} \in O$ such that $o^{\prime} \sim o^{\prime \prime}[=o]$ and $o^{\prime} \upharpoonright E=o^{\prime \prime} \upharpoonright E$, it is $o^{\prime}(\bar{d})=o^{\prime \prime}(\bar{d})$. The set

$$
\operatorname{icl}_{[\rho]}^{\sim}(E)=\left\{\bar{d} \in B^{n} ; \bar{d} \sim-\text { depends on } E[\text { for } o]\right\}
$$

is called the $\sim$-dependency closure of $E$ [for $o$ ]. It is easy to see that it is

$$
\operatorname{icl}^{\sim}(E)=\bigcap_{o \in O} \operatorname{icl}_{o}^{\sim}(E)
$$

A pair $\left(o, o^{\prime}\right)$ of operations on $B$ is called a $(\bar{d}, E)$-marriage if $o \upharpoonright E=o^{\prime} \upharpoonright E$, and $o(\bar{d}) \neq o^{\prime}(\bar{d})$. The purpose of marriages is to witness that a point $\bar{d}$ does not belong to the dependency closure of a set $E$.

### 2.1.2 Conjugation

We are going to construct marriages as pairs $\left(o, o^{g}\right)$, where $o^{g}=g^{-1} o g$, for an appropriate automorphism $g$ of $\mathcal{B}$. Clearly, it is $\left\langle\mathcal{B}, o^{g}\right\rangle \cong\langle\mathcal{B}, o\rangle$, via $g$, and

$$
\begin{equation*}
o^{g}(\bar{x})=o(\bar{x}) \Leftrightarrow o g(\bar{x})=g o(\bar{x}) \tag{2.1}
\end{equation*}
$$

for all $\bar{x} \in B^{n}$.
Instead of $\left\langle\mathcal{B}, o^{\prime}\right\rangle \cong\langle\mathcal{B}, o\rangle[$ via $g]$, we write shortly $o^{\prime} \cong o[$ via $g]$.

### 2.1.3 Almost uniform definability

An useful criterion for dependency can be formulated using the concept of almost uniform definability, which we state at this place.

Two sequences $\varepsilon, \varepsilon^{\prime}$ of elements of $A$ are said to be indistinguishable in $\mathcal{A}$ if they have the same complete type over $\emptyset$ in $\mathcal{A}$.

We say that a pair $\left(c, c^{\prime}\right) \in A^{2}$ is equidefinable from parameters $\left(\left\langle b_{i}\right\rangle_{i \in I},\left\langle b_{i}^{\prime}\right\rangle_{i \in I}\right)$ in $\mathcal{A}$ if there is an $L(\mathcal{A})$-formula $\varphi(\bar{x}, y)$ which equidefines $\left(c, c^{\prime}\right)$ in $\mathcal{A}$ from $\left(\left\langle b_{i}\right\rangle_{i \in I},\left\langle b_{i}^{\prime}\right\rangle_{i \in I}\right)$; i.e. there is $\left\{i_{j} ; j<l(\bar{x})\right\} \subseteq I$ such that $\varphi$ defines $c$ from $\left\langle b_{i_{j}}\right\rangle_{j<l(\bar{x})}$ and $c^{\prime}$ from $\left\langle b_{i_{j}}^{\prime}\right\rangle_{j<l(\bar{x})}$.

An operation $o$ is in $\mathcal{A}$ almost-uniformly definable (a.u.-definable) at a point $\bar{d} \in A^{n}$ over a set $E \subseteq A^{n}$ if, for all $\left\langle\bar{e}^{\prime}, \bar{d}^{\prime}\right\rangle_{\bar{e} \in E}$ such that $\varepsilon=\sqcup\langle\bar{e}, o(\bar{e}), \bar{d}\rangle_{\bar{e} \in E}$ and $\varepsilon^{\prime}=\sqcup\left\langle\bar{e}^{\prime}, o\left(\bar{e}^{\prime}\right), \bar{d}^{\prime}\right\rangle_{\bar{e} \in E}$ are indistinguishable in $\mathcal{A}$, the pair $\left(o(\bar{d}), o\left(\bar{d}^{\prime}\right)\right)$ is equidefinable from $\left(\varepsilon, \varepsilon^{\prime}\right)$ in $\mathcal{A}$.

Lemma 2.1:1. Let $o$ be an operation on $A$ and $\varepsilon=\left\langle a_{i}, \bar{d}\right\rangle_{i \in I}, \varepsilon^{\prime}=\left\langle a_{i}^{\prime}, \bar{d}^{\prime}\right\rangle_{i \in I}$ be two indistinguishable (in $\mathcal{A}$ ) systems of elements from $A$, with $|I|<|A|$, $l(\bar{d})=l\left(\bar{d}^{\prime}\right)=n$. Then the following statements are equivalent:

1) $\left(o(\bar{d}), o\left(\bar{d}^{\prime}\right)\right)$ is equidefinable in $\mathcal{A}$ from parameters $\left(\varepsilon, \varepsilon^{\prime}\right)$.
2) For every $g \in \operatorname{Aut}(\mathcal{A})$ such that $g\left(a_{i}\right)=a_{i}^{\prime}$ and $g(\bar{d})=\bar{d}^{\prime}$, it is og $(\bar{d})=g o(\bar{d})$.

Moreover, the implication "1) $\Rightarrow$ 2)" is true even for a non-saturated $\mathcal{A}$.
Proof. 1) $\Rightarrow$ 2): Easy.
$2) \Rightarrow 1)$ : Suppose $\left(o(\bar{d}), o\left(\bar{d}^{\prime}\right)\right)$ is not equidefinable from $\left(\varepsilon, \varepsilon^{\prime}\right)$. Define $g\left(a_{i}\right)=a_{i}^{\prime}$, $g(\bar{d})=\bar{d}^{\prime}$ and $g(o(\bar{d}))=e$, where $e \neq o\left(\bar{d}^{\prime}\right)=o g(\bar{d})$ is such that $\varepsilon_{\iota}\langle o(\bar{d})\rangle$ and $\varepsilon^{\prime}{ }_{\iota}\langle e\rangle$ are indistinguishable (we construct such $e$ later); then $g$ can be extended to an automorphism of $\mathcal{A}$ contradicting 2).

Existence of $e$ : Let $p(x)=\left\{\varphi\left(x, \varepsilon^{\prime}\right) ; \mathcal{A} \models \varphi(o(\bar{d}), \varepsilon)\right\} \cup\left\{x \neq o\left(\bar{d}^{\prime}\right)\right\}$. Since $\left(o(\bar{d}), o\left(\bar{d}^{\prime}\right)\right)$ is not equidefinable from $\left(\varepsilon, \varepsilon^{\prime}\right), p(x)$ is a type. Any $e$ realizing $p(x)$ has the demanded properties.

### 2.1.4 Fixators and DD-theorem

Let $G \subseteq \operatorname{Aut}(\mathcal{B})$ be a subgroup, o an $n$-ary operation on $B$, and $E \subseteq B^{n}$. We say that a saturated expansion $\mathcal{A}$ of $\mathcal{B}$ is a $(G, E)$-fixator for $o$ if

$$
g \in \operatorname{Aut}(\mathcal{A}) \Leftrightarrow g \in G, \text { and } o g(\bar{x})=g o(\bar{x}) \text { for all } \bar{x} \in E .
$$

For $o, o^{\prime} \in O$, we write $o \sim^{G} o^{\prime}$ if $o \cong o^{\prime}$ via some $g \in G$. Clearly, $\sim^{G}$ is an equivalence on $O$.

Proposition 2.1:2 (DD-theorem). Let o be an n-ary operation on $B, \bar{d} \in B^{n}$, and $E=E_{s} \cup E_{f} \subseteq B^{n}$, where $\left|E_{s}\right|<|B|$, and suppose that there is a saturated $\left(G, E_{f}\right)$-fixator $\mathcal{A}$ for o, with $G \subseteq \operatorname{Aut}(\mathcal{B})$. Then the following statements are equivalent:

1) $\bar{d} \in \operatorname{icl}_{o}^{\sim G}(E)$,
2) o is a.u.-definable at $\bar{d}$ over $E_{s}$ in $\mathcal{A}$.

Proof. We have the following: $\bar{d} \in \operatorname{icl}_{o}^{\mathcal{N}^{G}}(E) \Leftrightarrow$ For all $o^{\prime} \cong o$, via some $g \in G$ such that $o^{\prime} \upharpoonright E=o \upharpoonright E$, it is $o^{\prime}(\bar{d})=o(\bar{d})$.
$\Leftrightarrow \quad$ For all $g \in G$ and $\overline{d^{\prime}}, \overline{e^{\prime}}$ with $\bar{e} \in E$, if $g(\bar{d})=\overline{d^{\prime}}$, and $g(\bar{e})=\overline{e^{\prime}}, g(o(\bar{e}))=o\left(\overline{e^{\prime}}\right)$, for $\bar{e} \in E$, then $g(o(\bar{d}))=o\left(\overline{d^{\prime}}\right)$.
$\Leftrightarrow \quad o$ is a.u.-definable at $\bar{d}$ over $E_{s}$ in $\mathcal{A}$.
Above, the first $\Leftrightarrow$ is the definition of $\operatorname{icl}_{o}^{\sim^{G}}(E)$, the second $\Leftrightarrow$ is by setting $o^{\prime}=o^{g}$ and by (2.1), and the third $\Leftrightarrow$ follows by Lemma 2.1:1 and the definition of a $\left(G, E_{f}\right)$-fixator.

### 2.2 Dependency of Peano products

In this section, $\mathcal{B}$ will be a fixed saturated model of Presburger arithmetid ${ }^{1}$ and $O$ the set $\operatorname{sPP}(\mathcal{B})$ of all saturated Peano products on $\mathcal{B}$. Note that, up to an isomorphism, there are all saturated models of Peano arithmetic of size $|B|$ among the structures $\langle\mathcal{B}, \cdot\rangle$, where $\cdot \in \operatorname{sPP}(\mathcal{B})$.

We are going to prove the following proposition, which describes icl ${ }^{\operatorname{sPP}(\mathcal{B})}(E)$ for two particular cases: $E=\emptyset$ and $E=E_{a}=\{a\} \times B$, with $a \in B-\mathbb{N}$.

[^2]For $a \in B-\mathbb{N}$, let us denote

$$
D_{a}=\left\{\frac{p}{n} ; p \in{ }^{+} \mathbb{Z}[a], 0<n \in \mathbb{N} \text { and } \mathcal{B} \models n \mid p\right\}=\mathbb{Q}[a] \cap B
$$

(compare to $\mathrm{D}_{\mathcal{A}}$ from Example $\left.1.3: 2 \mathrm{~b}\right)$ ). We also write $\circ \cong_{[a]}$. if there is an isomorphism $f$ of $\langle\mathcal{B}, \circ\rangle$ and $\langle\mathcal{B}, \cdot\rangle$ such that $f(a)=a]$.

Proposition 2.2:1. Let $a \in B-\mathbb{N}$. The following holds:

1) $\mathrm{icl}^{\cong}(\emptyset)=(\mathbb{N} \times B) \cup(B \times \mathbb{N})$, for every $\cdot \in \operatorname{sPP}(\mathcal{B})$,
$\mathrm{icl}^{\cong}(\emptyset)=(\mathbb{N} \times B) \cup(B \times \mathbb{N})$,
2) $\operatorname{icl}^{\cong_{a}}\left(E_{a}\right)=\left(D_{a} \times B\right) \cup\left(B \times D_{a}\right)$, for every $\cdot \in \operatorname{sPP}(\mathcal{B})$, $\operatorname{icl}^{\cong_{a}}\left(E_{a}\right)=\left(D_{a} \times B\right) \cup\left(B \times D_{a}\right)$.

Let $\mathrm{P}(\mathcal{B})$ denote the set of all commutative, associative and distributive Robinson products on $\mathcal{B}$ and $\operatorname{PP}(\mathcal{B})$ the set of all Peano products on $\mathcal{B}$. We get the following easy corollary:

Corollary 2.2:2. Let $a \in B-\mathbb{N}$. The following holds:

1) $\operatorname{icl}^{\mathrm{P}(\mathcal{B})}(\emptyset)=\mathrm{icl}^{\mathrm{PP}(\mathcal{B})}(\emptyset)=\mathrm{icl}^{\mathrm{SPP}(\mathcal{B})}(\emptyset)=(\mathbb{N} \times B) \cup(B \times \mathbb{N})$,
2) $\operatorname{icl}^{\mathrm{P}(\mathcal{B})}\left(E_{a}\right)=\operatorname{icl}^{\mathrm{PP}(\mathcal{B})}\left(E_{a}\right)=\operatorname{icl}^{\operatorname{sPP}(\mathcal{B})}\left(E_{a}\right)=\left(D_{a} \times B\right) \cup\left(B \times D_{a}\right)$.

Proof. The inclusions " $\subseteq$ " follow from Proposition [2.2:1.
The opposite inclusions in 1) are trivial. In 2) they follow easily from commutativity, associativity and distributivity of the products.

Remark 2.2:3. Let us note that for the case $|E|<|B|$ the dependency problem is not difficult. Indeed, by the DD-theorem 2.1:2, $\bar{d} \in \operatorname{icl}{ }_{o}^{\cong}(E) \Leftrightarrow o$ is a.u.-definable at $\bar{d}$ over $E$ in $\mathcal{B}$ (because, clearly, $\mathcal{B}$ is an $(\operatorname{Aut}(\mathcal{B}), \emptyset)$-fixator for any operation $o$ on $B$ ). But $\mathcal{B} \models \operatorname{Pr}$, hence the definability problem can be easily solved.

Nevertheless, if $|E|=|B|$, the relevant fixator may be more complex structure than $\mathcal{B}$. For example, in the next section, we will see that for $E=E_{a}$ the respective fixator is a model of LA.

### 2.2.1 Fixators for Peano products

We will prove Proposition [2.2:1 using the DD-theorem [2.1:2. That is why we need to know the respective fixators:

## Observation 2.2:4.

a) $\mathcal{B}$ is an $(\operatorname{Aut}(\mathcal{B}), \emptyset)$-fixator for any operation o on $B$.
b) $\mathcal{B}_{a, \cdot}=\left\langle\mathcal{B}, a \cdot{ }_{-}\right\rangle$is an $\left(S_{a}, E_{a}\right)$-fixator for $\cdot \in \operatorname{sPP}(\mathcal{B})$, where $S_{a}$ is the stabilizer of a under the action of $\operatorname{Aut}(\mathcal{B})$.

Let us note that $\mathcal{B} \models$ Aa and $\mathcal{B}_{a} \models$ La. We are going to prove both cases of Theorem 2.2:1 at once. Further, we work in one of the following settings, which we fix:

- $\mathcal{A}=\mathcal{B}, G=\operatorname{Aut}(\mathcal{B}), E=\emptyset$,
- $\mathcal{A}=\mathcal{B}_{a, \cdot} G=S_{a}, E=E_{a}$, where $a \in B-\mathbb{N}$ and $\cdot \in \operatorname{sPP}(\mathcal{B})$.

In both cases, we have the following properties:
(*a) $\mathcal{A}$ is saturated,
${ }^{*} \mathrm{~b}$ ) every formula is in $\mathcal{A}$ equivalent to $\bigvee_{i<n}(\exists \bar{z}) \psi_{i}$, where $\psi_{i}$, with $i<n$, are systems of linear inequalities,
$\left({ }^{*} \mathrm{c}\right)$ the substructure $\mathcal{A}_{(\emptyset)}$ of all elements definable without parameters in $\mathcal{A}$ is an elementary substructure of $\mathcal{A}$.

This follows from Corollary 1.4:2 or Corollary 1.4:7, respectively.
The following Lemma is an adaptation of an idea by Jan Šaroch.
Lemma 2.2:5 (J. Šaroch). Let $p(\bar{x})$ be a complete type over $\emptyset$ in $\mathcal{A}$. Then $U=\{\bar{u} ; \mathcal{A} \models p(\bar{u})\}$ is closed under the operation $\bar{u}, \bar{v} \mapsto \frac{\bar{u}+\bar{v}}{2}$.

Proof. Let $\bar{u}, \bar{v} \in U$ and $\varphi(\bar{x}) \in p$. We prove $\varphi\left(\frac{\bar{u}+\bar{v}}{2}\right)$.
By $\left.{ }^{*} \mathrm{~b}\right)$, we may suppose that $\varphi$ is of the form $\bigvee_{i<n}(\exists \bar{z}) \psi_{i}$, where $\psi_{i}$, with $i<n$, are systems of linear inequalities. Since $\bar{u}$ and $\bar{v}$ have the same complete type, there is $i<n$ and $\bar{\pi} \in{ }^{l(z)} 2$ such that $\left(\exists \bar{z} \equiv_{2} \bar{\pi}\right) \psi_{i}$ holds for both $\bar{u}$ and $\bar{v}$. Then $(\exists \bar{z}) \psi_{i}$ holds for $\frac{\bar{u}+\bar{v}}{2}$, as well.

Lemma 2.2:6. Let $\bar{u} \in A^{2}$ and $U=\left\{\bar{u}^{\prime} \in A^{2} ; \operatorname{tp}(\bar{u})=\operatorname{tp}\left(\bar{u}^{\prime}\right)\right\}$. Then the following are equivalent:
a) None of $u_{0}, u_{1}$ is $\emptyset$-definable in $\mathcal{A}$.
b) $U$ contains $\bar{u}^{\prime}$ and $\bar{u}^{\prime \prime}$ such that $u_{i}^{\prime} \neq u_{i}^{\prime \prime}$, for $i=0,1$.

Proof. " b$) \Rightarrow \mathrm{a}$ )" is trivial.
"a) $\Rightarrow \mathrm{b})$ ": Set

$$
\begin{aligned}
U_{0} & =\left\{v_{1} ;\left(u_{0}, v_{1}\right) \in U\right\} \\
U_{1} & =\left\{v_{0} ;\left(v_{0}, u_{1}\right) \in U\right\}
\end{aligned}
$$

We show that each of $U_{0}, U_{1}$ has at least two elements. Then there are $v_{0} \neq u_{0}$ and $v_{1} \neq u_{1}$ such that $\left(u_{0}, v_{1}\right),\left(v_{0}, u_{1}\right) \in U$, and, by Lemma 2.2:5, the point $\left(u_{0}^{\prime}, u_{1}^{\prime}\right)=\left(\frac{u_{0}+v_{0}}{2}, \frac{u_{1}+v_{1}}{2}\right) \in U$ is different from $\left(u_{0}, u_{1}\right)$ in both coordinates.

Suppose that $U_{0}=\left\{u_{1}\right\}$. Then, by $(* a), u_{1}$ is definable from $u_{0}$, and thus it is $U=\left\{\left(u^{\prime}, f\left(u^{\prime}\right)\right) ; u^{\prime} \in \operatorname{dom}(U)\right\}$, for some definable function $f$. By our
assumption, $f\left(u^{\prime}\right)=u_{1}$, for all $u^{\prime} \in \operatorname{dom}(U)$. Therefore, again by $(*$ a), there is $\varphi \in \operatorname{tp}(\bar{u})$ such that $\mathcal{A} \models \varphi(\bar{x}) \rightarrow f\left(x_{0}\right)=u_{1}$. By $\left({ }^{*} \mathrm{c}\right)$, there is a $\emptyset$-definable $\bar{w} \in A^{2}$ such that $\varphi(\bar{w})$, and hence $u_{1}=f\left(w_{0}\right)$ is $\emptyset$-definable.

The case $U_{1}=\left\{u_{0}\right\}$ is symmetric.
Now, we are ready to prove Proposition 2.2:1,
Proposition 2.2:1, Let $a \in B-\mathbb{N}$. The following holds:

1) $\mathrm{icl}^{\cong}(\emptyset)=(\mathbb{N} \times B) \cup(B \times \mathbb{N})$, for every $\cdot \in \operatorname{sPP}(\mathcal{B})$, $\mathrm{icl}^{\cong}(\emptyset)=(\mathbb{N} \times B) \cup(B \times \mathbb{N})$,
2) $\mathrm{icl}_{\stackrel{\cong_{a}}{ }}\left(E_{a}\right)=\left(D_{a} \times B\right) \cup\left(B \times D_{a}\right)$, for every $\cdot \in \operatorname{sPP}(\mathcal{B})$, $\mathrm{icl}^{\cong_{a}}\left(E_{a}\right)=\left(D_{a} \times B\right) \cup\left(B \times D_{a}\right)$.

Proof. We need to prove that icl. $\sim^{\sim^{G}}(E)=\left(A_{(\emptyset)} \times A\right) \cup\left(A \times A_{(\emptyset)}\right)$. The inclusion " $\supseteq$ " is trivial. The opposite one is, by the DD-theorem 2.1:2 and Lemma 2.2:6, equivalent to the statement

- is a.u.-definable at $\bar{d}$ over $\emptyset$ in $\mathcal{A} \Rightarrow U=\{\bar{u} ; \operatorname{tp}(\bar{u})=\operatorname{tp}(\bar{d})\}$ does not contain $\bar{u}^{\prime}$ and $\bar{u}^{\prime \prime}$ such that $u_{i}^{\prime} \neq u_{i}^{\prime \prime}$, for $i=0,1$.

Suppose that • is a.u.-definable at $\bar{d}$ over $\emptyset$ in $\mathcal{A}$ by a formula $\varphi$ and that it is $\bar{u}^{\prime}, \bar{u}^{\prime \prime} \in U$. Then $U^{\prime}=\left\{\left(\bar{u}, u_{0} \cdot u_{1}\right) ; \bar{u} \in U\right\}$ is the set of all realizations of the type $\operatorname{tp}(\bar{d}) \cup\{\varphi(\bar{x}, y)\}$. Therefore, by Lemma 2.2:5, $\left(\frac{\bar{u}^{\prime}+\bar{u}^{\prime \prime}}{2}, \frac{u_{0}^{\prime} \cdot u_{1}^{\prime}+u_{0}^{\prime \prime} \cdot u_{1}^{\prime \prime}}{2}\right) \in U^{\prime}$, and hence $\frac{u_{0}^{\prime} \cdot u_{1}^{\prime}+u_{0}^{\prime \prime} \cdot u_{1}^{\prime \prime}}{2}=\frac{u_{0}^{\prime}+u_{0}^{\prime \prime}}{2} \cdot \frac{u_{1}^{\prime}+u_{1}^{\prime \prime}}{2}$. This implies $u_{0}^{\prime}=u_{0}^{\prime \prime}$ or $u_{1}^{\prime}=u_{1}^{\prime \prime}$.

### 2.3 Meeting pairs of Peano products

Let $\mathcal{B}$ be a fixed saturated model of Presburger arithmetic, as in section 2.2. For $a \in B-\mathbb{N}$, we denote $E_{a}=\{a\} \times B$ the "slice" of $B$ at $a$.

### 2.3.1 Meeting pair

Let $a \in B-\mathbb{N}$. A pair $(\cdot, \circ)$ of Peano products on $\mathcal{B}$ is called an a-meeting pair if it is $\cdot \upharpoonright E_{a}=\circ \upharpoonright E_{a}$, and $d_{0} \cdot d_{1} \neq d_{0} \circ d_{1}, d_{0}^{\prime} \cdot d_{1}^{\prime} \neq d_{0}^{\prime} \circ d_{1}^{\prime}$, for some $d_{0}, d_{1}<a<d_{0}^{\prime}, d_{1}^{\prime}$. The following is an easy consequence of Proposition 2.2:1;

Corollary 2.3:1. Let $a \in B-\mathbb{N}$, and $\cdot \in \operatorname{sPP}(\mathcal{B})$ be a saturated Peano product on $\mathcal{B}$. Then there is $\circ \in \operatorname{sPP}(\mathcal{B})$ such that $(\cdot, \circ)$ is an a-meeting pair of Peano products on $\mathcal{B}$. Moreover, ○ can be chosen in such a way that.$\cong_{a} \circ$.

Proof. By Proposition 2.2:1, there are points $\bar{d}, \overline{d^{\prime}} \notin \operatorname{icl}^{\cong_{a}}\left(E_{a}\right)$ with $\bar{d}<a<\overline{d^{\prime}}$. Let $\bullet, \bullet^{\prime}$ be witnesses for $\bar{d}, \overline{d^{\prime}}$ respectively, i.e. $\bullet \cong{ }_{a} \cong_{a} \bullet^{\prime}$ coincide with $\cdot$ on $E_{a}$, but $d_{0} \cdot d_{1} \neq d_{0} \bullet d_{1}, d_{0}^{\prime} \cdot d_{1}^{\prime} \neq d_{0}^{\prime} \bullet^{\prime} d_{1}^{\prime}$.

Suppose that neither $(\cdot, \bullet)$ nor $\left(\cdot, \bullet^{\prime}\right)$ is an $a$-meeting pair, then $\left(\bullet, \bullet^{\prime}\right)$ is one. Since $\bullet \cong_{a} \cdot$, via some $g$, we get $\bullet^{\prime} \cong_{a} \circ$, via $g$ (where $\circ=\bullet^{\prime g}$ is the " $g$-conjugate" of $\bullet^{\prime}$; see section 2.1.2), and $(\cdot, \circ)$ is an $a$-meeting pair.

Having a meeting pair $(\cdot, \circ)$, we construct a product $\times: B^{2} \rightarrow B$, different from both • and $\circ$, such that $\langle\mathcal{B}, \times\rangle \models T$, where $T$ is an extension of Robinson arithmetic Q by a set of induction axioms. This is stated as Proposition 2.3:2.

### 2.3.2 $\mathrm{LB}_{x}$ and $\mathrm{LcB}_{x}$ formulas

We denote $\mathrm{LB}_{x}\left[\mathrm{LcB}_{x}\right]$ the set of formulas $\varphi(x, \bar{y})$ in the language of arithmetic $L^{a r}=\langle 0, S,+, \cdot, \leq\rangle$ such that every occurrence of multiplication in $\varphi$ has the form $x \cdot z$ or $z \cdot x$, where $z$ is a variable which is bound by a quantifier $Q z \leq x$ $[Q z \geq x]$.

The $L^{a r}$-theory ILB [ILcB] is the extension of Robinson arithmetic Q by the scheme of induction $I\left(\mathrm{LB}_{x}\right)$ [ $\left.I\left(\mathrm{LcB}_{x}\right)\right]$ for all formulas $\varphi$ from $\mathrm{LB}_{x}\left[\mathrm{LcB}_{x}\right]$ (here, $x$ is the "induction variable").

Proposition 2.3:2. Let $a \in B-\mathbb{N}$, and $(\cdot, \circ)$ be an a-meeting pair of Peano products on $\mathcal{B}$.

1) For $\times=\cdot \upharpoonright[0, a]^{2} \cup \circ \upharpoonright\left(B^{2}-[0, a]^{2}\right)$, it is $\langle\mathcal{B}, \times\rangle \models \mathrm{ILB}$.
2) For $x^{\prime}=\upharpoonright[a, \infty)^{2} \cup \circ \upharpoonright\left(B^{2}-[a, \infty)^{2}\right)$, it is $\left\langle\mathcal{B}, x^{\prime}\right\rangle \models \mathrm{ILcB}$.

Proof. 1): Clearly, $\langle\mathcal{B}, \times\rangle \vDash \mathrm{Q}$. Let $\varphi(x, \bar{y}) \in \mathrm{LB}_{x}$. Then the following holds:

$$
\begin{align*}
& \langle\mathcal{B}, \times\rangle \models \varphi[b, \bar{c}] \quad \Leftrightarrow\langle\mathcal{B}, \cdot\rangle \models \varphi[b, \bar{c}], \text { for } b \leq a, \bar{c} \in B,  \tag{2.2}\\
& \langle\mathcal{B}, \times\rangle \models \varphi[b, \bar{c}] \quad \Leftrightarrow\langle\mathcal{B}, \circ\rangle \models \varphi[b, \bar{c}] \text {, for } b \geq a, \bar{c} \in B . \tag{2.3}
\end{align*}
$$

We prove that the axiom of induction for $\varphi$ holds in $\langle\mathcal{B}, \times\rangle$. Suppose that it is $\langle\mathcal{B}, \times\rangle \models \varphi[0, \bar{c}]$. Then, by (2.2) and by induction in $\langle\mathcal{B}, \cdot\rangle \vDash \mathrm{P}$, we get $\langle\mathcal{B}, \times\rangle \vDash \varphi[b, \bar{c}]$, for all $b \leq a$. Then, similarly, by (2.3) and by induction in $\langle\mathcal{B}, \circ\rangle \models \mathrm{P}$, we prove $\langle\mathcal{B}, \times\rangle \models \varphi[b, \bar{c}]$, for any $b \geq a$.
2) can be proven similarly.

We ask the following, a bit vague, open question:
Open question 2. Is it possible, by using similar methods, to construct Robinson products $\times$ which satisfy $I(\Gamma)$ for other sets $\Gamma \subseteq F m_{L^{a r}}$ ? In particular, is it possible to construct Peano products this way?

### 2.4 Peano interpolations

Let us remind that $\mathcal{B}$ stands for a fixed saturated model of Pr , and $O$ denotes a fixed set of $n$-ary operations on $B$. In this section, we deal with a more subtle problem connected with dependency: Given points $\left(\bar{b}_{i}, d_{i}\right) \in B^{n+1}$, for $i \in I$, is there an operation $o \in O$ such that $o\left(\bar{b}_{i}\right)=d_{i}$, for all $i \in I$ ?

We are going to prove the following partial answer:
Proposition 2.4:1. Let o: $B^{n} \rightarrow B$ satisfies

$$
\begin{equation*}
o(\bar{b})>\mathbb{N} \cdot \bar{b}, \text { for all } \bar{b}>\mathbb{N} \tag{2.4}
\end{equation*}
$$

and $\mathbb{N}<\bar{b}, d \in B$ be such that
i) $d>\mathbb{N} \cdot \bar{b}$,
ii) $d \equiv o(\bar{b}) \bmod n$, for all $0<n \in \mathbb{N}$.

Then there is $o^{\prime} \cong o$ such that $o^{\prime}(\bar{b})=d$.
Moreover, if $\bar{b} \equiv \overline{b^{\prime}} \bmod n \Rightarrow o(\bar{b}) \equiv o\left(\overline{b^{\prime}}\right) \bmod n$, for all $b \in B$ and $0<n \in \mathbb{N}$, then the other implication holds as well.
Proof. Since $\mathcal{B}$ is saturated, it is enough to show that, for a formula $\varphi(\bar{x}, y)$, with $l(x)=n$, it is

$$
\begin{equation*}
\mathcal{B} \models \varphi[\bar{b}, d] \Rightarrow \mathcal{B} \models \varphi\left[\overline{b^{\prime}}, o\left(\overline{b^{\prime}}\right)\right], \text { for some } \overline{b^{\prime}} \in B \text {. } \tag{2.5}
\end{equation*}
$$

Indeed, in that case we can find an automorphism $g$ of $\mathcal{B}$ (as in the proof of Lemma 2.1:1) such that $g\left(\overline{b^{\prime}}\right)=\bar{b}$ and $g\left(o\left(\overline{b^{\prime}}\right)\right)=d$, for some $\overline{b^{\prime}} \in B$. Then $o^{g}(\bar{b})=d$.

We prove (2.5). By Corollary 1.4:211), we may suppose that $\varphi$ is a conjunction of formulas $t=0, t>0$ and $n \mid t$, where $t$ is of the form $\sum_{i} k_{i} x_{i}+l y+m$, with $k_{i}, l, m \in \mathbb{Z}$, and $0<n \in \mathbb{N}$. Further, we work in $\mathcal{B}$.

Let $\varphi$ be $t=0$, and suppose $\varphi(\bar{b}, d)$. Then, by i), it is $l=0$, and hence also $\varphi(\bar{b}, o(\bar{b}))$. Let $\varphi$ be $t>0$. If $\varphi(\bar{b}, d)$ holds then either $l>0$, or $l=0$. In both cases, we get $\varphi(\bar{b}, o(\bar{b}))$; in the first case, we use (2.4). Finally, let $\varphi$ be $n \mid t$, and, again, suppose $\varphi(\bar{b}, d)$. Then $\varphi(\bar{b}, o(\bar{b}))$ holds, by ii).

The "moreover statement" is easy.
The following is an immediate consequence of Proposition 2.4:1.
Corollary 2.4:2. Let $\mathbb{N}<b_{0}, b_{1}, d \in B$. There is a Peano product $\circ$ on $\mathcal{B}$ such that $b_{0} \circ b_{1}=d$ if and only if
i) $d>\mathbb{N} \cdot b_{i}$, for $i<2$,
ii) " $d \equiv b_{0} \cdot b_{1} \bmod n "$, for all $0<n \in \mathbb{N}$.
(In ii), "..." means the obvious additive equivalent of ....)
Moreover, o may be chosen to be isomorphic to any given Peano product •.

## Chapter 3

## Quasi-Euclidean Subrings of $\mathbb{Q}[x]^{1}$

We present an algebraic connection of the material introduced in the previous chapters.

We show that the rings $\mathrm{D}_{\tau}$ from Example $1.3: 1 \mathrm{~b}$ ) are quasi-Euclidean subrings of $\mathbb{Q}[x]$ which are not $k$-stage Euclidean for any norm and positive integer $k$. These subrings can be either PID or non-UFD, depending on the choice of $\tau$. In both cases, there are $2^{\omega}$ such domains up to ring isomorphism. This solves the question of G. E. Cooke from Coo76, where he asked whether there is an example of quasi-Euclidean domain, which is not 2-stage Euclidean.

The quasi-Euclidean property of the rings $\mathrm{D}_{\tau}$ is proved in Theorem 3.4:2. The fact, that $\mathrm{D}_{\tau}$ 's are not $k$-stage Euclidean for any $0<k \in \mathbb{N}$, is showed in Theorem 3.4:9.

This chapter stands aside the chapters 1 and 2 as an independent part, and the connections to the previous chapters are rather loose. In order to keep the material of this chapter completely self-contained, we do not presume anything from chapters 1 and 2. This includes also a change in notation: Domains $\mathrm{D}_{\tau}$ from Example $1.3: 1$ (b) are denoted $R_{\tau}$ in this chapter, as this fits better the conventions of algebraic texts.

### 3.1 Introduction

Although Euclidean and principal ideal domains have been intensively studied for almost a century, examples of non-Euclidean PIDs are still rather scattered throughout the literature, and thought of as more or less singular, non-frequent objects. The oldest of these examples are arguably the rings of integers of $\mathbb{Q}(\sqrt{d})$ for $d=-19,-43,-67,-163$. However, these are the only cases for negative $d$ 's,

[^3]and the results from Wei73 and Har04 indicate that it is almost surely the case of positive values of $d$, too.

Another type of examples was given by Samuel in his famous paper [Sam71]. Leutbecher (in Leu78]) capitalized on his approach several years later, and proved that there are non-Euclidean PIDs which are even quasi-Euclidean (this was not the case of the four rings of integers mentioned above, as Cohn observed in [Coh66]).

Throughout this paper, by a quasi-Euclidean domain, we mean a commutative domain $R$ for which there is a function $\phi: R^{2} \rightarrow \omega$ such that, for all $(a, b) \in R^{2}$ with $b \neq 0$, there exists $q \in R$ with $\phi(b, a-b q)<\phi(a, b)$. The definition is similar to the one of classical Euclidean norm, with the important difference that by the norm function here, we do not measure elements of the ring but pairs of those. Also, $\omega$ can be equivalently replaced by some/any infinite ordinal in the definition; see Preliminaries section (in particular Proposition 3.2:1) for this and further equivalent definitions of quasi-Euclidean domain, and related concepts.

There are a few more published results on non-Euclidean PIDs. Unfortunately, they do not usually present a coherent class of these domains, or some sort of characterization of rings which are non-Euclidean PIDs in some distinguished class of domains. Nice attempts in this direction can be found in [And88] and [EH73].

In this text, we present a parametric construction which is in some sense a generalization of the approach used in [EH73]. We show that there are many discretely ordered non-Euclidean (even non- $k$-stage Euclidean in the sense of Cooke [Coo76]) subrings of $\mathbb{Q}[x]$ which are quasi-Euclidean. In fact, for each $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$, where $\mathbb{J}_{p}$ denotes the ring of $p$-adic integers, we define one such subring. Moreover, we observe that the set $\prod_{p \in \mathbb{P}} \mathbb{J}_{p}$ splits into two parts of full cardinalities, depending on whether the resulting ring is PID or non-UFD. Since each quasi-Euclidean ring is Bézout (Proposition 3.2:1), there are no inbetween cases, i.e. non-PID and UFD at the same time.

### 3.2 Preliminaries

Throughout this chapter, all rings are (commutative integral) domains. Further, we denote by $\mathbb{P}$ the set of all primes in $\mathbb{N}$. For each $p \in \mathbb{P}, \mathbb{J}_{p}$ stands for the ring of $p$-adic integers, while $\mathbb{Z}_{p}$ denotes the field $\mathbb{Z} / p \mathbb{Z}$. Since $\mathbb{J}_{p} \cong \lim _{\mathbb{Z}_{p^{k}}}$, we shall view $\mathbb{J}_{p}$ as a subring of $\prod_{k=1}^{\infty} \mathbb{Z}_{p^{k}}$, and denote, for a positive integer $k$, by $\pi_{k}$ the canonical projection from $\mathbb{J}_{p}$ to $\mathbb{Z}_{p^{k}}$. It will not cause any confusion that the notation $\pi_{k}$ does not reflect the prime $p$. Moreover, for technical reasons, we put $\pi_{0}: \mathbb{J}_{p} \rightarrow\{0\} ;$ again, regardless of the prime $p$.

If we deal with elements from the ring $\mathbb{Q}[x]$, we define $\operatorname{deg} 0=-1$, and we denote by $\operatorname{lc}(q)$ the leading coefficient of a polynomial $q$.

### 3.2.1 Quasi-Euclidean and $k$-stage Euclidean domains

Various generalizations of the concept of a Euclidean domain were proposed and studied in the past. The one we find very natural, is the concept of quasiEuclidean (used in Leu78] and [Bou80]) or the equivalent notion of $\omega$-stage Euclidean domain (used by Cooke in [Coo76]).

Given a ring $R$ and a partial order $\leq$ on $R^{2}$, we say that $\leq$ is quasi-Euclidean if it has the descending chain condition (dcc), and for any pair $(a, b) \in R^{2}$ with $b \neq 0$, there exists an element $q$ in $R$ such that $(b, a-b q)<(a, b)$. We call $R$ quasi-Euclidean provided there exists a quasi-Euclidean partial order on $R^{2}$.

Let $(a, b) \in R^{2}$ and $k$ be a non-negative integer. A $k$-stage division chain starting from the pair $(a, b)$ is a sequence of equations in $R$

$$
\begin{gathered}
a=q_{1} b+r_{1} \\
b=q_{2} r_{1}+r_{2} \\
r_{1}=q_{3} r_{2}+r_{3} \\
\vdots \\
r_{k-2}=q_{k} r_{k-1}+r_{k} .
\end{gathered}
$$

Such a division chain is called terminating if the last remainder $r_{k}$ is $0\left(r_{k-1}\right.$ is then easily seen to be the GCD of $a$ and $b$ ). Notice that a $k$-stage division chain is determined by its starting pair and the sequence of quotients $q_{1}, \ldots, q_{k}$. For the sake of compactness, in what follows, we shall denote this chain also by

$$
\left(\begin{array}{c|ccc}
a & q_{1} & \ldots & q_{k} \\
b & r_{1} & \ldots & r_{k}
\end{array}\right) .
$$

Given such a division chain, we define its 0 -th remainder $r_{0}$ as $b$.
In the following proposition, On denotes the class of all ordinal numbers.
Proposition 3.2:1. ( Bou80], Coo76], Leu78]) For a commutative domain $R$, the following conditions are equivalent:

1. There exists a function $\phi: R^{2} \rightarrow$ On (with $\operatorname{Rng}(\phi) \subseteq \omega$ ) such that, for all $(a, b) \in R^{2}$ with $b \neq 0$, there exists $q \in R$ such that $\phi(b, a-b q)<\phi(a, b)$.
2. $R$ is quasi-Euclidean.
3. $R$ is a Bézout domain, and the group $\mathrm{GL}_{2}(R)$ of regular $2 \times 2$ matrices over $R$ is generated by matrices of elementary transformations.
4. Every pair $(a, b) \in R^{2}$ with $b \neq 0$ has a terminating $k$-stage division chain for some positive integer $k$.

Proof. (1) $\Longrightarrow(2)$ is trivial, we just put $(a, b)<\left(a^{\prime}, b^{\prime}\right)$ if $\phi(a, b)<\phi\left(a^{\prime}, b^{\prime}\right)$.
$(2) \Longrightarrow(4)$ follows directly by the dcc.
The equivalence of (3) and (4) was proved already in [O'M64, 14.3].
$(4) \Longrightarrow(1):$ We put $\phi(a, 0)=0$ for all $a \in R$. If $b \neq 0$, we define $\phi(a, b)$ as the minimal $k \in \omega$ for which the pair $(a, b)$ has a terminating $k$-stage division chain. (So we even manage to find $\phi$ with the range in $\omega$.)

Notice that no notion of a norm is involved in the definition of a quasiEuclidean domain. However, given a norm $N$ on $R$ (i.e. a function $N: R \rightarrow \mathbb{N}$ with $N(a)=0$ iff $a=0$ ), we can measure how far $N$ is from being Euclidean: as in Coo76, for $0<k \leq \omega$, we say that $R$ is a $k$-stage Euclidean domain with respect to $N$ provided that, for every $(a, b) \in R^{2}$ with $b \neq 0$, there exists a positive integer $l \leq k$ such that for some $l$-stage division chain starting from $(a, b)$ it is $N\left(r_{l}\right)<N(b)$. As usual, we say that $R$ is $k$-stage Euclidean if there exists such a norm $N$ on $R$. So, in our notation, 1 -stage Euclidean means Euclidean (in the classic sense). On the other hand, by Proposition 3.2:1, $R$ is $\omega$-stage Euclidean (with respect to some/any norm) if and only if it is quasi-Euclidean.

Finally, observe that a quasi-Euclidean domain, being Bézout, is UFD if and only if it is PID. An example of non-UFD 2-stage Euclidean domain was given already by Cooke in Coo76, at the end of $\S 1$. It is at this place, where he admits that he does not know of any example of quasi-Euclidean domain which is not 2-stage Euclidean. Interestingly, all examples, we are going to construct, have got this property.

### 3.2.2 Peano arithmetic and weak saturation

Although our construction will be purely algebraic, we are going to give also a description derived from a nonstandard model of Peano arithmetic (PA). There are several reasons to do this: the description is very natural, only basic logical tools are needed, and it sheds more light at the entire situation.

Our models of PA are thought of as models in the language of arithmetic $L=(0,1,+, \cdot, \leq)$. The fact that it is an extension of the language of rings will make it more convenient for us to work with. In particular, we can immediately say that any model of PA is a (discretely ordered) commutative semiring with 0 and 1.

We will say that $\mathcal{M} \vDash \mathrm{PA}$ is weakly saturated if every 1-type in $\mathcal{M}$ without parameters is realized in $\mathcal{M}$, i.e. given any set $Y=\left\{\varphi_{i}(x) \mid i \in I\right\}$ of $L$-formulas with one free variable $x$, there is $m \in M$ such that $\mathcal{M} \models \varphi_{i}[m]$ for all $i \in I$, provided that, for each finite subset $S$ of $I$, one has $\mathcal{M} \models(\exists x) \bigwedge_{i \in S} \varphi_{i}(x)$. Indeed, weakly saturated models of PA exist, we can even take an appropriate elementary extension of $\mathbb{N}$, however, as we shall see, for such a model $\mathcal{M}$, it is necessarily $|M| \geq 2^{\omega}$.

### 3.3 Examples

### 3.3.1 Logical description

Let us fix a weakly saturated model $\mathcal{M}$. Then, as mentioned above, $\mathcal{M}$ forms a commutative semiring. Formally adding negative elements, we turn $\mathcal{M}$ into a commutative domain containing $\mathbb{Z}$ as a subring. We will denote this domain $\mathcal{M}^{ \pm}$. Notice that $\mathcal{M}^{ \pm}$shares several basic properties with $\mathbb{Z}$, namely it is a discretely ordered GCD domain; also for every $q, r$ with $r \neq 0$, there exists $0 \leq t<|r|$ such that $r$ divides $q+t$ (where $\left.\right|_{-} \mid$is the usual absolute value). However, unlike $\mathbb{Z}$, $\mathcal{M}^{ \pm}$is not Noetherian.

Let $a$ be a nonstandard element of $\mathcal{M}$, i.e. $a \in M \backslash \mathbb{N}$. We define a subring $R_{a}$ of $\mathcal{M}^{ \pm}$in the following way:

$$
R_{a}=\left\{m \in M^{ \pm} \mid(\exists n \in \mathbb{N})(\exists h \in \mathbb{Z}[x]) n \neq 0 \& n \cdot m=h(a)\right\}
$$

It is easily seen that $R_{a}$ is a ring. It can be naturally approached if we, in the first step, take a subring of $\mathcal{M}^{ \pm}$generated by $a$ (which is nothing else than $\mathbb{Z}[a] \cong \mathbb{Z}[x]$ ), and then allow division by nonzero integers in case it is possible in $\mathcal{M}^{ \pm}$. We immediately observe that $R_{a}$ is isomorphic to

$$
R_{a}^{\prime}=\left\{\left.\frac{h}{n} \in \mathbb{Q}[x] \right\rvert\, n \in \mathbb{N} \backslash\{0\}, h \in \mathbb{Z}[x], \text { and } n \mid h(a) \text { in } \mathcal{M}^{ \pm}\right\}
$$

Remark 3.3:1.

1. Regardless of $a$, we have $R_{a}^{\prime} \cap \mathbb{Q}=\mathbb{Z}$.
2. Notice that $R_{a}=R_{a+1}$ (for any nonstandard $a \in M$ ) but $R_{a}^{\prime} \neq R_{a+1}^{\prime}$ since precisely one of these two rings contains $x / 2$. On the other hand, as we shall see later, it is possible that we have nonstandard $a, b \in M$ such that $R_{a} \neq R_{b}$ but $R_{a}^{\prime}=R_{b}^{\prime}$.
3. For our considerations, we do not need the full strength of PA. In fact, instead of binary multiplication, it is enough to have an endomorphism $a$. of the monoid $(M,+, 0)$ such that $a \cdot 1 \notin \mathbb{N}$, and the induction for all formulas in the language $(0,1,+, a \cdot, \leq)$; so the resulting theory can be viewed as an extension of Presburger arithmetic rather than weakening of PA (see the theory LA from 1.1.4.1).

### 3.3.2 Algebraic description

As we have seen above, the definitions of $R_{a}$ and $R_{a}^{\prime}$ rely on the fixed model $\mathcal{M}$ of PA. However, there is only a little amount of information about $a \in M$ that we actually need. This makes it possible - as we are going to demonstrate - to
manage without refering to any Peano model. For $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$, we define a subring $R_{\tau}$ of $\mathbb{Q}[x]$.

$$
R_{\tau}=\left\{\left.\frac{h}{n} \in \mathbb{Q}[x] \right\rvert\, n \in \mathbb{N} \backslash\{0\}, h \in \mathbb{Z}[x], \text { and }(\forall p \in \mathbb{P}) \pi_{\mathrm{v}_{p}(n)}\left(h\left(\tau_{p}\right)\right)=0\right\}
$$

Here, $\mathrm{v}_{p}$ denotes the usual $p$-valuation. Further, $\tau_{p}$ is the $p$ th projection of $\tau$, and the substitution $h\left(\tau_{p}\right)$ is done inside $\mathbb{J}_{p}$ where $\mathbb{Z}$ is canonically embedded via $z \mapsto\left(z \bmod p, z \bmod p^{2}, z \bmod p^{3}, \ldots\right)$. We will use this substitution several times in the next section.

It follows easily from the definition that $\sigma \neq \tau$ implies $R_{\sigma} \neq R_{\tau}$. The correspondence between the rings $R_{a}^{\prime}$ and $R_{\tau}$ is made precise by Proposition 3.3:2,

Proposition 3.3:2. Let $\mathcal{M}$ be a weakly saturated model of PA. Then:

1. For each nonstandard $a \in M$ there exists precisely one $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$ such that $R_{a}^{\prime}=R_{\tau}$.
2. For each $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$ there is at least one nonstandard $a \in M$ such that $R_{a}^{\prime}=R_{\tau}$.

Proof. (1) There is even a ring homomorphism $\psi: \mathcal{M}^{ \pm} \rightarrow \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$ which assigns to $m \in M^{ \pm}$an element $\tau$ such that $\tau_{p}=\left(m \bmod p, m \bmod p^{2}, m \bmod p^{3}, \ldots\right)$ for each $p \in \mathbb{P}$. It is a matter of straightforward verification that $R_{a}^{\prime}=R_{\psi(a)}$ for any nonstandard $a \in M$.
(2) Let us consider the set $Y$ consisting of all congruences $x \equiv_{p^{k}} \tau_{p}(k)$ and inequalities $x>k$, where $k \in \mathbb{N} \backslash\{0\}$ and $p \in \mathbb{P}$. Then $Y$ is a 1-type in $\mathcal{M}$ (without parameters-positive integers are just constant terms in the language $L)$ since any finite subset of $Y$ has a solution in $\mathbb{N} \subset M$ by Chinese Remainder Theorem. So there is a global solution, $a \in M$, of all congruences and inequalities from $Y$, using the weak saturation of $\mathcal{M}$. (Now, it is clear that $|M| \geq 2^{\omega}$.) The inequalities assure that $a$ is nonstandard, and checking the definitions, we immediately see that $R_{a}^{\prime}=R_{\tau}$.

In the following section, we will freely use the fact (implicitly proved above) that, for every $\tau$, the ring $R_{\tau}$ inherits the discrete ordering from $\mathcal{M}^{ \pm}$via isomorphism with $R_{a}$ for some/any $a$.

### 3.4 Properties of the examples

### 3.4.1 Terminating division chains

We are going to show that, for every $\tau$, the $\operatorname{ring} R_{\tau}$ is quasi-Euclidean. So let $\tau$ be fixed for a while, put $R=R_{\tau}$, and let us denote by $R^{+}$the subsemiring of $R$
consisting of polynomials with nonnegative leading coefficients. First, we prove the following auxiliary result.

Lemma 3.4:1. Let $q, r \in R^{+}$with $r \neq 0$, then there are (unique) $p, s \in R^{+}$such that $q=p r+s$ and $s<r$.

Moreover: Let $\tilde{p}, \tilde{s} \in \mathbb{Q}[x]$ be such that $q=\tilde{p} r+\tilde{s}$ and $\operatorname{deg} \tilde{s}<\operatorname{deg} r$. Further let $\tilde{p}=p^{\prime} / m$ where $p^{\prime} \in \mathbb{Z}[x], m \in \mathbb{N} \backslash\{0\}$ and $0 \leq k<m$ such that $\left(p^{\prime}-k\right) / m \in R^{+}$. Then the pair $(p, s)$ satisfies

$$
(p, s)= \begin{cases}(\tilde{p}-1, \tilde{s}+r) & \text { for } k=0 \& \operatorname{lc}(\tilde{s})<0 \\ \left(\frac{p^{\prime}-k}{m}, \tilde{s}+\frac{k}{m} r\right) & \text { otherwise }\end{cases}
$$

Proof. Straightforward verification.
If we look at $R_{a}$ (for $a$ with $R_{a}^{\prime}=R$ ), there is only one pair $(p, s)$ in the model $\mathcal{M}$ satisfying the properties from Lemma 3.4:1, namely the pair $(q \operatorname{div} r, q \bmod r)$. Here, div stands for the binary operation of integer division. Thus in particular, we have that $R^{+}$as a subsemiring of $\mathcal{M}$ is closed under binary operations div and mod.

Consequently, we say that a division chain $\left(\begin{array}{c|ccc}r_{-1} & q_{1} & \ldots & q_{n} \\ r_{0} & r_{1} & \ldots & r_{n}\end{array}\right)$ in $R^{+}$with $r_{-1}, r_{0}>0$ is quasi-Euclidean if $q_{i+1}=r_{i-1} \operatorname{div} r_{i}$ and $r_{i+1}=r_{i-1} \bmod r_{i}$, for $i \geq 0$. A consequence of the proof of the following theorem is that, for any nonzero $a, b \in R^{+}$, there exists a positive integer $n$ such that the quasi-Euclidean chain of length $n$ starting from the pair $(a, b)$ is terminating.

Theorem 3.4:2. $R$ is a quasi-Euclidean domain. In particular, it is Bézout.
Proof. We will show that the condition (1) from Proposition 3.2:1 is satisfied. For this sake, we define $\phi: R^{2} \rightarrow\left(2 \times \mathbb{N}^{4}\right.$, lex $)$ by the formula $\phi(q, r)=(0,0,0,0,0)$ for $r=0$, and

$$
\phi(q, r)=\left(\delta_{q, r}, \operatorname{deg} q+1, \operatorname{deg} r, n_{q, r}, n_{q, r} \cdot|\operatorname{lc}(q)|\right)
$$

otherwise. Here, $\delta_{q, r}$ is 1 if $|q| \leq|r|$, and 0 otherwise; $n_{q, r} \in \mathbb{N}$ denotes the least common denominator of $q, r$. In the rest of the proof, we assume that $q>r>0$. The other cases follow easily. (Notice that $\phi(q, r)=\phi(|q|,|r|)$.)

Since $\mathbb{Q}[x]$ is a Euclidean ring with the norm $\operatorname{deg}(-)+1$, there are $\tilde{p}, \tilde{s} \in \mathbb{Q}[x]$ such that $q=\tilde{p} r+\tilde{s}$ and $\operatorname{deg} \tilde{s}<\operatorname{deg} r$. By Lemma 3.4:1, we get $p, s \in R^{+}$ satisfying $s<r$ and $q=p r+s$.

Suppose $s \neq 0$. We need to show that $\phi(r, s)<\phi(q, r)$ in the lexicographic order of $2 \times \mathbb{N}^{4}$. Since $0<s<r$, we have $\delta_{r, s}=0=\delta_{q, r}$. We may assume $\operatorname{deg} q=\operatorname{deg} r=\operatorname{deg} s$ (otherwise, we are done immediately). Then $p \in \mathbb{N}$. Further, we have $q, r \in \frac{\mathbb{Z}[x]}{n_{q, r}}$, and hence $s=q-p r \in \frac{\mathbb{Z}[x]}{n_{q, r}}$. Therefore $n_{r, s} \leq n_{q, r}$. Moreover, from $r<q$, we have $\operatorname{lc}(r) \leq \operatorname{lc}(q)$.

Assume $n_{r, s}=n_{q, r}$ and $\operatorname{lc}(r)=\operatorname{lc}(q)$. Then, from the definition of $\tilde{p}$, we have $\tilde{p}=1$, and thus $p^{\prime}=1=m, k=0$ in Lemma 3.4:1. The first case in the definition of $(p, s)$ leads to a contradiction, since we get $p=0$ (and so $q=s<r)$. So it must be that $p=\tilde{p}=1$ and $s=\tilde{s}$. In particular, we see that $\operatorname{deg} s=\operatorname{deg} \tilde{s}<\operatorname{deg} r$ which also contradicts one of our assumptions.

Finally, $R$ is Bézout by Proposition 3.2:1,

### 3.4.2 Separating the PID cases

In the following few paragraphs, we distinguish the choices of $\tau$ which imply that $R_{\tau}$ is a PID. We also show that there are $2^{\omega}$ pairwise nonisomorphic domains among the rings $R_{\tau}$ which are PID, and the same cardinality of those which are not PID. The next lemma will be useful.
Lemma 3.4:3. Let $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$. Then $R_{\tau}$ is a PID if and only if, for each nonzero $h \in \mathbb{Z}[x]$, the set $S_{h}=\left\{(p, k) \in \mathbb{P} \times(\mathbb{N} \backslash\{0\}) \mid \pi_{k}\left(h\left(\tau_{p}\right)\right)=0\right\}$ is finite.
Proof. Assume that $S_{h}$ is infinite for some nonzero $h \in \mathbb{Z}[x]$. Then either the set $\left\{p \in \mathbb{P} \mid h / p \in R_{\tau}\right\}$ is infinite, or there exists a prime $p$ such that $h / p^{k} \in R_{\tau}$ for any $k \in \mathbb{N}$. In the first case, we fix an enumeration $\left\{p_{1}, p_{2}, p_{3}, \ldots\right\}$ of that set, and-using the definition of $R_{\tau}$-we see that $\left(h / p_{1}, h /\left(p_{1} p_{2}\right), h /\left(p_{1} p_{2} p_{3}\right), \ldots\right)$ is an infinite descending (with respect to divisibility) sequence of elements in $R_{\tau}$; thus $R_{\tau}$ is not a UFD. In the second case, we use the same argument for the sequence $\left(h / p, h / p^{2}, h / p^{3}, \ldots\right)$.

If $R_{\tau}$ is not a PID, then (since it is Bézout by Theorem 3.4:2) there has to be an infinite sequence of elements in $R_{\tau}$ descending in divisibility $\left(h_{1} / n_{1}, h_{2} / n_{2}, \ldots\right)$; here $h_{i} \in \mathbb{Z}[x]$ and $n_{i}$ are positive integers coprime with $h_{i}$ in $\mathbb{Z}[x]$, for all $i>0$. The polynomials $h_{i}$ will eventually have the same degree $(\mathbb{Q}[x]$ is Euclidean) and absolute value of the leading coefficient ( $\mathbb{Z}$ is Noetherian), and so we may w.l.o.g. assume that all the polynomials $h_{i}$ are equal to a single nonzero $h \in \mathbb{Z}[x]$. It directly follows that, for this $h$, the set $S_{h}$ is infinite.

Let us take a representative subset $J$ of $\prod_{p \in \mathbb{P}} \mathbb{J}_{p}$ in the sense that, for each $\rho$, there is a $\tau \in J$ such that $R_{\tau} \cong R_{\rho}$, and for all $\tau, \sigma \in J, \tau \neq \sigma$, we have $R_{\tau} \neq R_{\sigma}$. Then $J$ is a disjoint union of the sets $A$ and $B$, where $A=\left\{\tau \in J \mid R_{\tau}\right.$ is a PID $\}$ and $B=\left\{\tau \in J \mid R_{\tau}\right.$ is not a UFD $\}$.
Proposition 3.4:4. $|A|=|B|=2^{\omega}$.
Proof. Let us assume that $|A|<2^{\omega}$. For each $p \in \mathbb{P}$, we define $\tau_{p} \in \mathbb{J}_{p}$ in such a way that:

1. $\pi_{1}\left(\tau_{p}\right)=\lfloor\log p\rfloor$,
2. $n \cdot \tau_{p} \notin\left\{h\left(\sigma_{p}\right) \mid \sigma \in A \& h \in \mathbb{Z}[x]\right\}$, for every positive integer $n$,
3. $\tau_{p}$ is not a root in $\mathbb{J}_{p}$ of a nonzero polynomial from $\mathbb{Z}[x]$.

This is clearly possible since the first two conditions are satisfied by $2^{\omega}$ different elements of $\mathbb{J}_{p}$. Let $\tau=\prod_{p \in \mathbb{P}} \tau_{p}$. We claim that $R_{\tau}$ is a PID which leads immediately to a contradiction (by (2), there cannot be $\sigma \in A$ with $R_{\tau} \cong R_{\sigma}$ ).

To prove this, we use Lemma 3.4:3, Let us fix a nonzero $h \in \mathbb{Z}[x]$. Then, using the limit comparison of $h$ and log, we deduce that, for all sufficiently large primes $p$, we have $0<|h(\lfloor\log p\rfloor)|<p$ which further implies $\pi_{1}\left(h\left(\tau_{p}\right)\right) \neq 0$. Together with the condition (3), we get that $S_{h}$ is finite. This finishes the proof that $|A|=2^{\omega}$.

To see that $|B|=2^{\omega}$, it is enough to fix a $\sigma \in A$, and for each nonzero subset $P$ of $\mathbb{P}$ define $\tau^{P} \in B$ by setting $\tau_{p}^{P}=(0,0,0, \ldots)$ for $p \in P$, and $\tau_{p}^{P}=\sigma_{p}$ otherwise.

### 3.4.3 Keeping distance from Euclidean domains

Here, we prove that no $R_{\tau}$ is a $k$-stage Euclidean domain, whatever positive integer $k$ we take. From now on, we work in a fixed ring $R_{\tau}$. We start with two slightly technical lemmas ${ }^{2}$.

Lemma 3.4:5. Let $Q=\left(\begin{array}{c|ccc}a & q_{1} & \ldots & q_{k} \\ b & r_{1} & \ldots & r_{k}\end{array}\right)$ be a division chain starting from $(a, b)$ with $a, b, k>0$. There is a division chain $Q^{\prime}=\left(\begin{array}{c|ccc}a & q_{1}^{\prime} & \ldots & q_{l}^{\prime} \\ b & r_{1}^{\prime} & \ldots & r_{l}^{\prime}\end{array}\right)$ with $q_{i}^{\prime}>0$ for $i>1$ such that $\left|r_{k}\right|=\left|r_{l}^{\prime}\right|$ and $l \leq 2 k-1$.

Proof. Denote $T_{1}, T_{2}$ the following two transformations on the set of all division chains starting from $(a, b)$ :

$$
\left.\begin{array}{c}
T_{1}:\left(\begin{array}{c|ccc}
a & q_{1} & \ldots & q_{k} \\
b & r_{1} & \ldots & r_{k}
\end{array}\right) \mapsto \\
\left(\begin{array}{c|ccccccc}
a & q_{1} & \ldots & q_{i-1} & q_{i}-1 & 1 & -\left(q_{i+1}+1\right) \\
b & r_{1} & \ldots & r_{i-1} & r_{i}+r_{i-1} & -r_{i} & (-1)^{2} r_{i+1} & (-1)^{3} r_{i+2}
\end{array} \ldots_{i} \pm r_{k}\right.
\end{array}\right) .
$$

where $i$ is the first index such that $q_{i+1}<0\left(T_{1}\right.$ is identity if there is no such $i$ ) and $\pm$ stands for $(-1)^{k-i+1}$;

$$
T_{2}:\left(\begin{array}{c|ccc}
a & q_{1} & \ldots & q_{k} \\
b & r_{1} & \ldots & r_{k}
\end{array}\right) \mapsto\left(\begin{array}{c|ccccccc}
a & q_{1} & \ldots & q_{i-1} & q_{i}+q_{i+2} & q_{i+3} & \ldots & q_{k} \\
b & r_{1} & \ldots & r_{i-1} & r_{i+2} & r_{i+3} & \ldots & r_{k}
\end{array}\right)
$$

where $i$ is the first index such that $q_{i+1}=0\left(T_{2}\right.$ is identity if there is no such $\left.i\right)$.
We will show a little bit more than stated; instead of $l \leq 2 k-1$, we prove even that $l \leq k+n$ where $n=\max \left\{k-i+1 ; i>1 \& q_{i}<0\right\}(n=0$ if there is no

[^4]such $i$ ). Put $Q=\left(\begin{array}{c|ccc}a & q_{1} & \ldots & q_{k} \\ b & r_{1} & \ldots & r_{k}\end{array}\right)$ and denote the corresponding pair $(n, k)$ as $p_{Q}=\left(n_{Q}, k_{Q}\right)$. We prove the statement by induction on the pairs $\left(n_{Q}, k_{Q}\right)$ with lexicographic ordering. The case $p_{Q}=(0,1)$ is trivial.

If there is $i$ such that $q_{i+1}=0$, we get $p_{T_{2}(Q)} \leq_{l e x}\left(n_{Q}, k_{Q}-2\right)$, and the induction assumption gives some $Q^{\prime}$. It is easy to verify that this $Q^{\prime}$ meets all the requirements. (Note that in the case $i+1=k$ we get $T_{2}(Q)=\left(\begin{array}{c|ccc}a & q_{1} & \ldots & q_{i-1} \\ b & r_{1} & \ldots & r_{i-1}\end{array}\right)$ and $r_{i-1}=r_{i+1}$.)

Otherwise, we have $q_{i} \neq 0$ whenever $i>1$, and using $T_{1}$ we get

$$
p_{T_{1}(Q)} \leq_{l e x}\left(n_{Q}-1, k_{Q}+1\right)
$$

Again, the $Q^{\prime}$ given by the induction assumption is what we wanted.
Lemma 3.4:6. Let $\left(\begin{array}{c|ccc}a & q_{1} & \ldots & q_{k} \\ b & r_{1} & \ldots & r_{k}\end{array}\right)$ be a division chain starting from ( $a, b$ ) such that $a, b, q_{i}>0$ for $i>1$, and let $\left(\begin{array}{c|ccc}a & e_{1} & \ldots & e_{m} \\ b & f_{1} & \ldots & 0\end{array}\right)$ be the quasi-Euclidean division chain in $R_{\tau}$ starting from $(a, b)$. Assume $m \geq k$.

Then $\left|r_{k}\right| \geq f_{k+1}$, and in particular $\operatorname{deg}\left(r_{k}\right) \geq \operatorname{deg}\left(f_{k+1}\right)$ (we put $f_{k+1}=0$ if $m=k)$.

Proof. Take the least $l$ such that $q_{l} \neq e_{l}$ (if there is no such, we are done since $\left(f_{i}\right)$ is decreasing). By an inductive argument, it is easy to observe that the following holds (recall that we put $f_{0}=r_{0}=b$ ):
If $q_{l}<e_{l}$ then $\left\{\begin{array}{l}r_{l+2 i} \geq r_{l-1} \text { for } i \geq 0, \\ r_{l+2 i+1} \leq-r_{l-1} \text { for } i \geq 1, \\ r_{l+1} \leq-r_{l-1} \text { or } r_{l+1}=-f_{l} ;\end{array}\right.$
and if $q_{l}>e_{l}$ then $\left\{\begin{array}{l}r_{l+2 i}<-r_{l-1} \text { for } i \geq 1, \\ r_{l+2 i+1}>r_{l-1} \text { for } i \geq 0, \\ r_{l} \leq-r_{l-1} \text { or }\left(m>k \& r_{l} \leq-f_{l+1}\right) \text {. }\end{array}\right.$
The statement follows since $r_{l-1}=f_{l-1}$ and $\left(f_{i}\right)$ is decreasing.
Combining both lemmas together, we obtain the following corollary which gives us a bound on the speed of decrease of remainders in a division chain, compared to the quasi-Euclidean one. By letting $a, b$ be any two consecutive Fibonacci numbers, one can see that the bound is optimal.
Corollary 3.4:7. Given $a, b>0$, let $\left(\begin{array}{c|ccc}a & e_{1} & \ldots & e_{n} \\ b & f_{1} & \ldots & 0\end{array}\right)$ be the quasi-Euclidean division chain starting from $(a, b)$, and $\left(\begin{array}{c|ccc}a & q_{1} & \ldots & q_{k} \\ b & r_{1} & \ldots & r_{k}\end{array}\right)$ be an arbitrary division chain. Then, for $l \leq \min (k, n / 2)$, we have $\left|r_{l}\right| \geq f_{2 l}$.

Now, we have all the tools for proving that no $R_{\tau}$ is $k$-stage Euclidean domain, independently of the choice of $k>0$. For the sake of better readability, we state the key step of the proof as a separate lemma.

Lemma 3.4:8. Let $k$ be a positive integer and $0<b \in R_{\tau}$ such that $\operatorname{deg}(b) \geq 1$. Then there is $0<a \in R_{\tau}$ such that every division chain $\left(\begin{array}{c|ccc}a & q_{1} & \ldots & q_{l} \\ b & r_{1} & \ldots & r_{l}\end{array}\right)$ of length $l \leq k$ starting from $(a, b)$ satisfies $\operatorname{deg}\left(r_{l}\right) \geq \operatorname{deg}(b)$.

Proof. By Corollary 3.4:7, it is enough to prove the statement for the quasiEuclidean division chain instead of an arbitrary one.

Set $a=\frac{c}{d}(b-\beta)$ where $c, d \in \mathbb{N}$ are such that no division chain in $\mathbb{Z}$ of length $l \leq k$ starting from $(c, d)$ is terminating (such $c, d$ exist since Corollary 3.4:7 holds also in $\mathbb{Z})$ and $0 \leq \beta<d$ is such that $d \mid(b-\beta)$ in $R_{\tau}$.

For a contradiction, let the quasi-Euclidean division chain

$$
\left(\begin{array}{c|ccc}
a & e_{1} & \ldots & e_{l} \\
b & f_{1} & \ldots & f_{l}
\end{array}\right)
$$

starting from $(a, b)$ satisfy $\operatorname{deg}\left(f_{l}\right)<\operatorname{deg}(b)$. We may w.l.o.g. assume $\operatorname{deg}\left(f_{l-1}\right)=$ $\operatorname{deg}(b)$; then we have $e_{i} \in \mathbb{Z}$ for all $i=1,2, \ldots, l$.

Define the operation ${ }^{\wedge}: R_{\tau} \rightarrow \mathbb{Q}$ as $\hat{r}=\operatorname{lc}(d r) / \operatorname{lc}(b)$. Easily $\hat{a}, \hat{b} \in \mathbb{Z}$, and therefore also $\hat{f}_{i} \in \mathbb{Z}$, for all $i \neq l$. Hence,

$$
\left(\begin{array}{c|cccc}
\hat{a} & e_{1} & \ldots & e_{l-1} & e_{l} \\
\hat{b} & \hat{f}_{1} & \ldots & \hat{f}_{l-1} & 0
\end{array}\right)
$$

is a division chain in $\mathbb{Z}$ starting from $(\hat{a}, \hat{b})=(c, d)$, a contradiction.

Theorem 3.4:9. Let $\tau \in \prod_{p \in \mathbb{P}} \mathbb{J}_{p}$ be arbitrary. Then the ring $R_{\tau}$ is not $k$-stage Euclidean for any positive integer $k$.

Proof. Assume the contrary and let $N$ be a norm such that $R_{\tau}$ is $k$-stage Euclidean with respect to $N$. To get a contradiction, we construct an infinite sequence $\left(b_{0}, b_{1}, \ldots\right)$ of elements from $R_{\tau}$ with $N\left(b_{i}\right)>N\left(b_{i+1}\right)$ and such that $\operatorname{deg} b_{i+1} \geq \operatorname{deg} b_{i} \geq 1$, for all $i \in \mathbb{N}$.

As the first step, put $b_{0}=x \in R_{\tau}$. Now assume we have defined $b_{i}$ for all $i \leq j \in \mathbb{N}$. Suppose $b_{j}>0$. For $b_{j}$ we find some $a_{j}$ using Lemma 3.4:8, By the $k$-stage Euclidean property, there is an $l$-stage division chain $\left(\begin{array}{c|ccc}a_{j} & q_{1} & \ldots & q_{l} \\ b_{j} & r_{1} & \ldots & r_{l}\end{array}\right)$ with $l \leq k$ starting from the pair $\left(a_{j}, b_{j}\right)$ such that $N\left(r_{l}\right)<N\left(b_{j}\right)$. So we can set $b_{j+1}=r_{l}$. By Lemma 3.4:8, we know that $\operatorname{deg} b_{j+1} \geq \operatorname{deg} b_{j} \geq 1$.

The case $b_{j}<0$ is similar. For $-b_{j}$ find $-a_{j}$ by Lemma 3.4:8, take a division chain $\left(\begin{array}{c|ccc}a_{j} & q_{1} & \ldots & q_{l} \\ b_{j} & r_{1} & \ldots & r_{l}\end{array}\right)$ with $N\left(r_{l}\right)<N\left(b_{j}\right)$ and set $b_{j+1}=r_{l}$. If it were
$\operatorname{deg} r_{l}<\operatorname{deg} b_{j}$, we would have the division chain $\left(\begin{array}{c|ccc}-a_{j} & q_{1} & \ldots & q_{l} \\ -b_{j} & -r_{1} & \ldots & -r_{l}\end{array}\right)$ with $\operatorname{deg}-r_{l}<\operatorname{deg}-b_{j}$, contradicting the choice of $-a_{j}$.

We conclude this chapter by the following
Open question 3. Is there an example of a $k$-stage Euclidean domain which is not $(k-1)$-stage Euclidean, for $k>2$ ?

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[^0]:    ${ }^{1}$ The name "linear arithmetic" is used somewhat vaguely and/or inconsistently through the literature. It denotes more different concepts where an important role is played by inequalities of "linear" combinations of "unknowns". In this thesis, linear arithmetic denotes the first order theory LA from section 1.1.4.1 (or an equivalent theory).

[^1]:    ${ }^{1}$ It seems that no attention has been yet paid to the extensions of Pr by linear functions. The reason is, probably, that they are trivial when eximined in the standard model $\langle\mathbb{N}, 0,1,+\rangle$. Even the quite general result of Semënov in [Sem84, Theorem 2, p.617] dismisses linear functions as trivial definable cases.

[^2]:    ${ }^{1}$ Let us note that existence of saturated models is, in general, not provable in ZFC. However, a saturated model of Presburger aritmhetic of size $\kappa$ exists provided that $\omega<\kappa=\kappa^{<\kappa}$. In particular, under the assumption of continuum hypothesis, there is such a model of size $2^{\omega}$. Here and further on, we therefore assume continuum hypothesis (which even implies that every countable structure in a countable language has a saturated elementary extension of size $2^{\omega}$ ).

[^3]:    ${ }^{1}$ This chapter is joint work with Jan Šaroch, and is essentially identical to the paper GŠ13. The authors would like to thank Josef Mlček and Jan Trlifaj for reading parts of this text and giving several valuable comments.

[^4]:    ${ }^{2}$ Lemma 3.4:5 is a modified version of a classical result on continued fractions by Perron (see [Per13, §37]).

