# Extensions and Restrictions of Abstract Categorial Grammars （抽象的範疇文法の拡張と制限） 

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# Extensions and Restrictions of Abstract Categorial Grammars 

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Thesis submitted for the degree of
Doctor of Philosophy (Information Studies)
to the Graduate School of Interdisciplinary Information Studies, the University of Tokyo

September 2006

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## Acknowledgment

I am very much grateful to Dr. Makoto Kanazawa, who led and encouraged me to get motivated to work on Abstract Categorial Grammars, the subject of this thesis, and gave me crucial advice and comments. He has supported me throughout my dissertation research, not only while he was my official supervisor, but also after he moved from the University of Tokyo to the National Institute of Informatics.

Dr. Kanazawa organized the Workshop on Lambda Calculus and Formal Grammar twice in Tokyo. I got explicit and implicit suggestions for my research from the talks and discussions by speakers at the workshop, Dr. Philippe de Groote, who is the advocator of Abstract Categorial Grammars, Dr. Reinhard Muskens, Dr. Sylvain Pogodalla, Dr. Kazushige Terui, Dr. Sylvain Salvati and Dr. Kanazawa. It is my pleasure to express my gratitude to them.

Dr. de Groote and Dr. Pogodalla offered me kind hospitality and valuable academic experience when I visited their laboratory. I deeply appreciate their support.

Dr. Salvati provided me many occasions to discuss Abstract Categorial Grammars and the Lambda Calculus with him and gave me many interesting suggestions and useful advice. I thank him very much for his help and friendship.

I would like to thank Dr. Makoto Tatsuta and the members of his seminar, Dr. Kanazawa, Dr. Terui, Mr. Daisuke Kimura and Mr. Takayuki Koai. I talked about my work in progress there and got many helpful comments from them.

Last but not least, I am sincerely indebted to the members of the thesis committee, Dr. Jun'ichi Tsujii, my official supervisor, Dr. Hiroshi Nakagawa, my official vice-supervisor, Dr. Ryu Hasegawa, Dr. Kumiko Tanaka-Ishii and Dr. Kanazawa. They gave me not only valuable criticism but also warm encouragement.

## Abstract

Phillippe de Groote has introduced a new grammar formalism called Abstract Categorial Grammars (ACGs) based on simply typed linear lambda calculus. Not only can ACGs explain various problematic linguistic phenomena in elegant ways, but also they encode mildly context-sensitive formalisms in straightforward ways. This thesis is concerned with the mathematical properties of some simple extensions and restrictions of ACGs. In particular, the aspect of ACGs as a generalization of well-established grammar formalisms including those mildly context-sensitive formalisms is investigated. Extensions and restrictions of ACGs that we discuss are non-linear ACGs, lexicalized forms of ACGs, and two-dimensional ACGs.

While the linearity constraint on ACGs is considered reasonable for several reasons, non-linear $\lambda$-terms enable ACGs to represent some linguistic phenomena in a more natural fashion. We show that relaxing the linearity constraint by allowing vacuous $\lambda$-abstraction does not increase the expressive power of ACGs. Our conversion entails that allowing or disallowing deleting rules in some existing grammar formalisms does not change their generative power.

Although the origin of ACGs is located in a history of categorial grammars, ACGs are not necessarily lexicalized grammars by definition, unlike usual categorial grammar formalisms. Not only from the lexicalist's point of view but also from a computational point of view, the restriction to lexicalized ACGs is an important issue. We show that some subclasses of ACGs admit lexicalization in the sense that each grammar in one of these subclasses can be converted into an equivalent lexicalized grammar belonging to the same class.

Two-dimensional extensions of ACGs define relations among languages generated by ACGs, and are able to model the relations between sentences and their meanings of natural language. We investigate the generative capacity of two-dimensional ACGs through encoding some well-known tree transducers.

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## Chapter 1

## Introduction

One important attractive feature of so-called categorial grammars as a tool for modeling natural language is that a sentence and its meaning are derived from one common proof in the underlying logic. However, several linguistic phenomena are found that cannot be explained by the Lambek Grammar [29], which is the most representative categorial grammar formalism. Among several modifications (e.g., Morrill [37], Moortgat [35,36], Oehrle [40,41]) of the Lambek Grammar, de Groote [15] has introduced a new grammar formalism called Abstract Categorial Grammars (ACGs), which is based on the implicational fragment of the linear logic. De Groote's approach is essentially the same as Oehrle's, but the ACG formalism radically simplifies Oehrle's formalism. While in Oehrle's formalism, the underlying logic contains the conjunction and word forms are represented by a special kind of typed $\lambda$ terms, ACGs represent word forms and meanings as well as the derivation structures for them by simply typed linear $\lambda$-terms. While Muskens [38] has independently proposed a formalism almost equivalent to ACGs, we deal with ACGs because of their formal simplicity and tractability for mathematical discussion.

ACGs have a number of theoretically and linguistically attractive features as a grammar formalism for natural language. Although ACGs are not a variety of Lambek grammars, they inherit the virtues of categorial grammars. Moreover, ACGs give elegant explanations for several linguistic phenomena that cannot be treated well by Lambek grammars, such as the ambiguity of the scopes of quantifiers in a sentence like "every man loves a woman". While usual categorial grammars treat word forms and meanings asymmetrically, ACGs treat word forms as $\lambda$-terms as well as meanings. This simplification enables ACGs generate diverse types of languages, because typed $\lambda$-terms, which constitute the languages of ACGs, can represent diverse types of data including not only strings and meaning representations,
but also trees, lists, and so on. Besides, the simply typed linear lambda calculus has been studied very well so far. Hence we have plenty of useful tools for investigating the mathematical properties of ACGs.

The mathematical properties of ACGs have close relationship with socalled mildly context-sensitive grammars rather than usual categorial grammars. Since some linguistic phenomena, such as cross-serial dependencies in Dutch, were found that cannot be captured by context-free grammars (CFGs), several cautious extensions of CFGs, called mildly context-sensitive grammars, have been proposed and studied. Some formalisms are known to be equivalent; classes of languages defined by combinatory categorial grammars, head grammars, linear indexed grammars, and tree adjoining grammars (TAGs) are all equivalent $[22,55,56]$. Moreover, linear context-free rewriting systems (LCFRSs), multiple context-free grammars, minimalist grammars, and multi-component TAGs (MCTAGs) define the same class of languages $[33,34,49,56]$. Recent studies $[16,18,19,58]$ on ACGs have shown that ACGs are powerful enough to encode those types of mildly contextsensitive grammars. They have presented straightforward encodings of CFGs, TAGs, non-deleting non-duplicating context-free tree grammars (CFTGs), LCFRSs, MCTAGs, where derivation structures in the original grammar are preserved. For instance, the ACG encoding a TAG generates derived trees through $\lambda$-terms encoding derivation trees of the TAG. Moreover, de Groote and Pogodalla [19] implicitly introduce a hierarchy of ACGs according to the complexity of $\lambda$-terms and types appearing in each ACG, and then associate the variety of context-free formalisms, namely, CFGs, non-deleting nonduplicating CFTGs, and LCFRSs, with the hierarchy of ACGs. In this sense, the ACG can be thought of as a generalization of mildly context-sensitive formalisms. To study ACGs would be to study those well-established grammar formalisms. In fact, de Groote and Pogodalla [19] state "Abstract Categorial Grammars . . . should rather be seen as the kernel of a grammatical framework ... in which other existing grammatical models may be encoded".

This thesis is devoted to studying the mathematical properties of some extensions and restrictions of ACGs. In particular, the aspect of ACGs that generalizes well-established grammar formalisms is investigated. Through discussion on the three main issues of this thesis, elimination of deleting operations, lexicalization, and two-dimensional formalisms, it is demonstrated that ACGs can be thought of as a generalization of existing grammars in various senses.

## Elimination of deleting operations.

Preceding research on several grammar formalisms have shown that deleting operations in a grammar can be eliminated preserving the language. We
show that a similar result holds for ACGs, which generalizes the results by preceding research on elimination of deleting operations in some grammar formalisms.

## Lexicalization.

A grammar is called lexicalized if each of its lexical entries that contribute to deriving an element of the language contains an item that explicitly appears on the surface of the derived structure. We show that a certain subclass of ACGs admits lexicalization. As Schabes [47] has shown that every finitely ambiguous ACG can be lexicalized as a TAG, our lexicalization method converts the ACG encoding a CFG into the one encoding a lexicalized TAG as a string generator.

## Two-dimensional formalisms.

Not a few grammar formalisms, which define languages, have been extended into two-dimensional formalisms, in which grammars define relations between two languages. We present some encoding methods for existing two-dimensional formalisms by two-dimensional ACGs.

## Overview of This Thesis

In the next chapter, we give a formal definition of ACGs and see basic mathematical properties of ACGs as well as the results on ACGs obtained by preceding research. In particular, it is shown that if the universal membership problem or the non-emptiness problem of ACGs is decidable, then so is the multiplicative exponential fragment of the linear logic, which is in open, and that non-semilinear languages can be generated by ACGs. Those results might imply that ACGs have too much powerful generative capacity and intractable computational complexity for modeling natural language, so finding an appropriate restriction on ACGs is desired.

Chapter 3 is concerned with non-linear extensions of ACGs. While ACGs are based on the simply typed linear lambda calculus, admitting non-linear $\lambda$-terms as lexical entries may allow ACGs to describe linguistic phenomena in a more natural and concise fashion. On the other hand, preceding research concerning mildly context-sensitive grammars has shown that deleting operations in grammars can be eliminated preserving their languages. Namely, Seki et al. [49] have shown that every multiple context-free grammar (MCFG) has an equivalent LCFRS, which can be thought of as an MCFG that has no deleting operations, and Fujiyoshi [10] has established the equivalence between non-duplicating monadic CFTGs and non-deleting non-duplicating monadic CFTGs. Along this line, we show that the generative capacity of
the usual ACGs is equivalent to that of the more liberal type of ACGs called affine $A C G s$, where vacuous $\lambda$-abstraction is allowed. Our conversion from an affine ACG into an equivalent linear ACG derives the results by Seki et al. and by Fujiyoshi as corollaries. This demonstrates that not only individual mildly context-sensitive grammars are encodable by ACGs, but also some transformations of grammars can be encoded in the ACG formalism.

In Chapter 4, we focus on lexicalized ACGs. A grammar is called lexicalized if each of its lexical entries that contribute to deriving an element of the language contains an item that explicitly appears on the surface of the derived structure. From the point of lexicalists' view, which thinks that linguistic phenomena should be accounted for by the inherent information in the lexical entries, to be lexicalized is a natural requirement. Schabes [47] has shown that every finitely ambiguous TAG admits lexicalization. Besides, because each production of a CFG is encoded as a lexical entry in the ACG encoding the CFG, a CFG in Greibach normal form can be considered lexicalized. While usual categorial grammars are lexicalized by definition, ACGs are not necessarily lexicalized. We first show that the universal membership problem for lexicalized ACG is NP-complete. This result contrasts with the one for general ACGs and suggests that the restriction to lexicalized ACGs may be preferable from the point of view of computational complexity as well as of lexicalism. Meanwhile, Salvati has introduced the notion of semilexicalized ACGs by relaxing the notion of lexicalized ACGs and shown that the languages generated by semilexicalized ACGs are in NP. His result suggests that the generative capacity of semilexicalized ACGs is not very much richer than that of lexicalized ACGs. The main result of Chapter 4 is lexicalization of semilexicalized ACGs. We show that each grammar in one level of the ACG hierarchy can be converted into an equivalent lexicalized grammar belonging to the same level if the level is higher than a certain level. Our lexicalization, however, generalizes neither lexicalization of TAGs nor Greibach normalization of CFGs, since the ACGs encoding TAGs or CFGs are located at a lower level than that level. Nevertheless, the fact that our lexicalization method converts ACGs encoding CFGs into lexicalized ACGs of the form encoding TAGs as string generators has a certain correspondence with Schabes's lexicalization method that lexicalizes finitely ambiguous CFGs as TAGs.

Chapter 5 is devoted to investigating the generative capacity of twodimensional ACGs. Lambek grammars are two-dimensional grammars in the sense that they generate pairs of a sentence and its meaning. Although the ACG formalism is designed to be easily extended into a two-dimensional formalism, there is little research on the mathematical properties of twodimensional ACGs. In order to evaluate the generative capacity of two-
dimensional ACGs, we present encoding methods that convert several existing two-dimensional formalisms, which define relations between two languages, into two-dimensional ACGs. In particular, a precise characterization of the class of non-deleting non-duplicating macro tree transducers $[2,6]$ is given in the ACG hierarchy. Besides, it is shown that synchronous TAGs $[50,53]$ are encodable by two-dimensional ACGs. These results show that the ACG formalism generalizes existing two-dimensional formalisms as well as one-dimensional formalisms. On the other hand, we show that several properties of usual one-dimensional ACGs hold also for two-dimensional ones. In particular, as Salvati has shown that a part of the hierarchy of one-dimensional ACGs collapses, we show that the corresponding part of the hierarchy of two-dimensional ACGs also collapses through encoding deterministic tree-walking transducers [1] by two-dimensional ACGs.

The subject of Chapter 6 is a restricted kind of higher-order matching. The higher-order matching problem is an important topic in computer science. Parsing with ACGs has a close connection with a special form of the higher-order matching problem called linear interpolation. We show that linear interpolation is NP-complete. While each variable occurs exactly once in a linear $\lambda$-term, the number of occurrences of constants are not restricted. It is natural to ask whether the NP-hardness of linear interpolation still holds when we exclude multiple occurrences of constants from linear $\lambda$-terms. We answer the question in the affirmative.

## Chapter 2

## Abstract Categorial Grammars

### 2.1 Definitions and Notations

$\varnothing$ is the empty set, $\varepsilon$ is the empty string, $\mathbb{Z}$ is the set of integers, $\mathbb{N}$ is the set of non-negative integers. $|X|$ denotes the cardinality of $X$ if $X$ is a set, and the length of $X$ if $X$ is a sequence. The sequence consisting of elements $X_{1}, \ldots, X_{n}$ in this order is denoted by $\left\langle X_{1}, \ldots, X_{n}\right\rangle$. If no confusion occurs, in particular if each $X_{i}$ is not a sequence, we sometimes write $X_{1} \ldots X_{n}$ for $\left\langle X_{1}, \ldots, X_{n}\right\rangle$. If $\vec{Y}$ denotes $\left\langle Y_{1}, \ldots, Y_{m}\right\rangle, X_{1} \ldots X_{m} \vec{Y} Z_{1} \ldots Z_{l}$ should be read as $\left\langle X_{1}, \ldots, X_{n}, Y_{1}, \ldots, Y_{m}, Z_{1}, \ldots, Z_{l}\right\rangle$ and so on. $X_{1} \ldots X_{m}$ is said to be a subsequence of a sequence $\vec{Y}$ if there are some possibly empty sequences $\vec{Y}_{0}, \ldots, \vec{Y}_{m}$ such that $\vec{Y}=\vec{Y}_{0} X_{1} \vec{Y}_{1} X_{2} \ldots \vec{Y}_{m-1} X_{m} \vec{Y}_{m}$. For a fixed index set $I$ and a set represented as $S=\left\{X_{i} \mid i \in I\right\}$, we write $\left\langle X_{i} \in S \mid \phi(i)\right\rangle$ (or $\left.\left\langle X_{i}\right\rangle_{\phi(i)}\right)$ for an appropriate sequence of $X_{i}$ such that $\phi(i)$ holds, if the order of the elements is determined uniquely from the context, or can be arbitrarily determined. Standard language-theoretic notions like concatenation, Kleene star (*), homomorphism etc. are defined as usual.

### 2.1.1 Lambda-Terms

Types Let $\mathscr{A}$ be a finite non-empty set of atomic types. The set $\mathcal{T}(\mathscr{A})$ of types built on $\mathscr{A}$ is defined as the smallest superset of $\mathscr{A}$ such that

- if $\alpha, \beta \in \mathcal{T}(\mathscr{A})$, then $(\alpha \rightarrow \beta) \in \mathcal{T}(\mathscr{A})$.

For $\alpha \in \mathcal{T}(\mathscr{A}),|\alpha|$ represents the number of occurrences of atomic types in $\alpha$. The set $\mathcal{S T}^{+}(\alpha)$ of positive subtypes of a type $\alpha$, the set $\mathcal{S T}^{-}(\alpha)$ of negative subtypes of a type $\alpha$, and the set $\mathcal{S T}(\alpha)$ of subtypes of a type $\alpha$ are defined as follows:

- $\mathcal{S T}^{+}(p)=\{p\}, \mathcal{S T}^{-}(p)=\varnothing$ for $p \in \mathscr{A}$.
- $\mathcal{S T}^{+}((\alpha \rightarrow \beta))=\{(\alpha \rightarrow \beta)\} \cup \mathcal{S T}^{+}(\beta) \cup \mathcal{S T}^{-}(\alpha)$, and $\mathcal{S T}^{-}((\alpha \rightarrow$ $\beta))=\mathcal{S T}^{-}(\beta) \cup \mathcal{S T}^{+}(\alpha)$.
- $\mathcal{S T}(\alpha)=\mathcal{S T}^{+}(\alpha) \cup \mathcal{S T}^{-}(\alpha)$.

Let ord : $\mathcal{T}(\mathscr{A}) \rightarrow \mathbb{N}$ be defined as

- $\operatorname{ord}(p)=1$ for all $p \in \mathscr{A}$,
- $\operatorname{ord}((\alpha \rightarrow \beta))=\max \{\operatorname{ord}(\alpha)+1, \operatorname{ord}(\beta)\}$.

We say that a type $\alpha$ is $n$-th order if ord $(\alpha) \leq n$. We use $o, p, q, r, s$ for atomic types and early lower case Greek letters $\alpha, \beta, \gamma, \ldots$ for types unless otherwise noted. We write $\alpha_{1} \rightarrow \cdots \rightarrow \alpha_{n} \rightarrow \beta$ for $\left(\alpha_{1} \rightarrow\left(\cdots \rightarrow\left(\alpha_{n} \rightarrow \beta\right) \ldots\right)\right)$ and if $\vec{\alpha}=\alpha_{1} \ldots \alpha_{n}, \vec{\alpha} \rightarrow \beta$ means $\alpha_{1} \rightarrow \cdots \rightarrow \alpha_{n} \rightarrow \beta . \alpha^{n} \rightarrow \beta$ stands for $\underbrace{\alpha \rightarrow \cdots \rightarrow \alpha}_{n \text {-times }} \rightarrow \beta$. A second-order type $\alpha \in \mathcal{T}(\mathscr{A})$ is said to be $n$-ary if $\alpha=p_{1} \rightarrow \cdots \rightarrow p_{n} \rightarrow q$ for some $p_{i}, q \in \mathscr{A}$. Thus an atomic type is called a nullary second-order type.

Higher-Order Signatures A higher-order signature $\Sigma$ is a triple $\langle\mathscr{A}, \mathscr{C}, \tau\rangle$ where $\mathscr{A}$ is a finite non-empty set of atomic types, $\mathscr{C}$ is a finite set of constants, and $\tau$ is a function from $\mathscr{C}$ to $\mathcal{T}(\mathscr{A})$ called a type assignment. We say that a higher-order signature $\Sigma$ is $n$-th order if $\operatorname{ord}(\Sigma) \leq n$ where $\operatorname{ord}(\Sigma)=\max \{\operatorname{ord}(\tau(\mathrm{a})) \mid \mathrm{a} \in \mathscr{C}\}$. When we write $\Sigma, \Sigma^{\prime}, \Sigma_{i}$ etc. to denote higher-order signatures, we assume $\Sigma=\langle\mathscr{A}, \mathscr{C}, \tau\rangle, \Sigma^{\prime}=\left\langle\mathscr{A}^{\prime}, \mathscr{C}^{\prime}, \tau^{\prime}\right\rangle$, $\Sigma_{i}=\left\langle\mathscr{A}_{i}, \mathscr{C}_{i}, \tau_{i}\right\rangle$ and so forth.

Lambda Terms Let $\mathscr{X}$ be a countably infinite set of variables. The set $\Lambda^{\mathrm{K}}(\Sigma)$ of $\lambda$-terms (terms for short) built upon $\Sigma$ and the type $\hat{\tau}(M)$ of a term $M \in \Lambda^{\mathrm{K}}(\Sigma)$ are defined inductively as follows:

- For every $\mathrm{a} \in \mathscr{C}, \mathrm{a} \in \Lambda^{\mathrm{K}}(\Sigma)$ and $\hat{\tau}(\mathrm{a})=\tau(\mathrm{a})$.
- For every $x \in \mathscr{X}$ and $\alpha \in \mathcal{T}(\mathscr{A}), x^{\alpha} \in \Lambda^{\mathrm{K}}(\Sigma)$ and $\hat{\tau}\left(x^{\alpha}\right)=\alpha$.
- For $M, N \in \Lambda^{\mathrm{K}}(\Sigma)$, if $\hat{\tau}(M)=(\alpha \rightarrow \beta), \hat{\tau}(N)=\alpha$, then $(M N) \in$ $\Lambda^{\mathrm{K}}(\Sigma)$ and $\hat{\tau}((M N))=\beta$.
- For $x \in \mathscr{X}, \alpha \in \mathcal{T}(\mathscr{A})$ and $M \in \Lambda^{\mathrm{K}}(\Sigma),\left(\lambda x^{\alpha} \cdot M\right) \in \Lambda^{\mathrm{K}}(\Sigma)$ and $\hat{\tau}\left(\left(\lambda x^{\alpha} \cdot M\right)\right)=(\alpha \rightarrow \hat{\tau}(M))$.

We assume that every occurrence of a variable $x$ in one term has the same type; e.g., $\lambda x^{p} . x^{p \rightarrow q} x^{p}$ is disallowed. Constants and variables are called atomic terms, terms of the form ( $M N$ ) are called application terms, terms of the form $\left(\lambda x^{\alpha} . M\right)$ are called abstraction terms. For convenience, we simply write $\tau$ instead of $\hat{\tau}$ and often omit the superscript on a variable if its type is clear from the context. We write $M_{1} \ldots M_{n}$ for $\left(\left(\ldots\left(M_{1} M_{2}\right) \ldots\right) M_{n}\right)$, $\lambda x_{1} \ldots x_{n} . M$ for $\left(\lambda x_{1} \cdot\left(\lambda x_{2} \ldots .\left(\lambda x_{n} \cdot M\right) \ldots\right)\right)$, and so on. If $\vec{x}=x_{1} \ldots x_{n}$ and $\vec{M}=M_{1} \ldots M_{m}$, then $\lambda \vec{x} \cdot M_{0} \vec{M}$ represents the term $\lambda x_{1} \ldots x_{n} \cdot M_{0} M_{1} \ldots M_{m}$. The notions of free variables, closed terms, $\beta$-normal form, $\beta \eta$-normal form, $\eta$-long form, are defined as usual (see a standard text book, e.g. [20]). A term $M$ is a combinator iff $M$ is closed and $M$ contains no constants. The head of $M$ is the atomic term $\times$ if $M=\lambda \vec{x} \times \vec{M}$. In a term $\times M_{1} \ldots M_{m}$, each occurrence of $M_{i}$ is said an argument of the head occurrence of x , and conversely the functor of the occurrence of $M_{i}$ is the head occurrence of x . We often say that a term $M$ is $k$-th order ( $k$-ary) if $\tau(M)$ is $k$-th order ( $k$-ary).

A capture-avoiding substitution of a term $M$ for a free variable $x$ is denoted by $[M / x]$. When $\vec{x}=x_{1} \ldots x_{n}$ and $\vec{M}=M_{1} \ldots M_{n},[\vec{M} / \vec{x}]$ means the simultaneous substitution $\left[M_{1} / x_{1}, \ldots, M_{n} / x_{n}\right] .\left[M_{i} / x_{i}\right]_{\phi(i)}$ means the simultaneous substitution $\left[\left\langle M_{i}\right\rangle_{\phi(i)} /\left\langle x_{i}\right\rangle_{\phi(i)}\right]$. For a substitution $\sigma, N \sigma$ denotes the term obtained from $N$ by applying the substitution and $\lambda x y z . L M N \sigma$ means $\lambda x y z .((L M N) \sigma)$. A substitution that may capture free variables is denoted by $[x:=M]$. Thus, $\lambda x \cdot y[x / y]=\lambda x \cdot x,(\lambda x \cdot y)[x / y]=\lambda z \cdot x$, $(\lambda x . y)[y:=x]=\lambda x . x$, and so on.

As usual, let $\rightarrow_{\beta},={ }_{\beta},={ }_{\beta \eta}$, $\equiv$ denote $\beta$-reduction, $\beta$-equality, $\beta \eta$-equality, and $\alpha$-equality respectively. A long normal term means a $\eta$-long $\beta$-normal term. The $\beta \eta$-equality is often simply written by $=.|M|_{\beta},|M|_{\beta \eta}$ and $\operatorname{Fv}(M)$ respectively represent the $\beta$-normal form, $\beta \eta$-normal form, and the set of free variables of $M$. We use upper case italic letters $M, N, P, \ldots$ for terms, late lower case italic letters $x, y, z, \ldots$ for variables, and sanserif $\mathrm{a}, \mathrm{A}, \ldots$ for constants in principle.

The size of a term is defined by

$$
\begin{aligned}
\operatorname{size}(\mathrm{a}) & =\operatorname{size}(x)=1 \\
\operatorname{size}(M N) & =\operatorname{size}(M)+\operatorname{size}(N), \\
\operatorname{size}(\lambda x \cdot M) & =\operatorname{size}(M)+1
\end{aligned}
$$

For a term $M \in \Lambda(\Sigma)$ and $\mathrm{a} \subseteq \mathscr{C}, \#_{\mathrm{a}}(M)$ denotes the number of occurrences of the constant a in $M$. Moreover, for $\mathscr{C}^{\prime} \subseteq \mathscr{C}, \# \mathscr{C}^{\prime}(M)$ denotes the total number of occurrences of the constants in $\mathscr{C}^{\prime}$. This notation is defined in the same way for strings.

A term is said to be linear if every $\lambda$-abstraction binds exactly one occurrence of a variable and no free variable of it occurs more than once.

Convention 2.1. Since this thesis mainly concerns linear $\lambda$-terms, we denote the set of linear $\lambda$-terms on $\Sigma$ by $\Lambda(\Sigma)$ and moreover we omit the modifier "linear", so if we say simply a term, it means a linear $\lambda$-term hereafter.

Strings, Trees and Lambda-Terms A second-order signature $\Sigma$ is called string signature if $\mathscr{A}=\{o\}$ and $\tau(\mathrm{a})=o \rightarrow o$ for every a $\in \mathscr{C}$. We write str for $o \rightarrow o$. For an (unranked) alphabet $V, \Sigma_{V}$ denotes the corresponding string signature such that $\mathscr{C}_{V}=V$. For a string $\mathrm{a}_{1} \ldots \mathrm{a}_{n} \in V^{*}$, we define the corresponding closed linear term as

$$
/ \mathrm{a}_{1} \ldots \mathrm{a}_{n} / \equiv \lambda z^{o} \mathrm{a}_{1}\left(\mathrm{a}_{2}\left(\ldots\left(\mathrm{a}_{n} z\right) \ldots\right)\right) .
$$

The concatenation is represented by the combinator $\mathbf{B} \equiv \lambda x^{s t r} y^{s t r} z^{o} . x(y z)$. For $\mathrm{w}_{1}, \mathrm{w}_{2} \in V^{*}$, we have

$$
\mathbf{B} / \mathrm{w}_{1} / / \mathrm{w}_{2} /=/ \mathrm{w}_{1} \mathrm{w}_{2} / .
$$

For notational convenience, we write $M+N$ instead of $\mathbf{B} M N$ for two terms of type str. Moreover, by the associativity of + , i.e., $(L+M)+N=L+(M+N)$, we can omit parentheses as $L+M+N$ if $\beta \eta$-equivalent terms can be identified.

A ranked alphabet is a pair $\langle F, \rho\rangle$ where $F$ is an (unranked) alphabet and $\rho$ is a function from $F$ to the non-negative integers. $F^{(n)}$ denotes the set $\{\mathrm{f} \in F \mid \rho(\mathrm{f})=n\}$. The set $\mathbb{T}(\langle F, \rho\rangle)$ of trees over $\langle F, \rho\rangle$ is the smallest set such that

- if $\mathrm{f} \in F^{(n)}$ and $M_{1}, \ldots, M_{n} \in \mathbb{T}(\langle F, \rho\rangle)$, then $\left(\mathrm{f} M_{1} \ldots M_{n}\right) \in \mathbb{T}(\langle F, \rho\rangle)$.

The outer most pair of parentheses of a tree is usually omitted. For a possibly infinite set $\mathscr{X}$ of variables, by $\mathbb{T}(\langle F, \rho\rangle \cup \mathscr{X})$ we denote the set of trees that may contain variables as symbols of rank 0 . A second-order signature $\Sigma$ is called a tree signature if $\mathscr{A}=\{o\}$. For a ranked alphabet $\langle F, \rho\rangle$, we define the corresponding tree signature as $\Sigma_{\langle F, \rho\rangle}=\left\langle\{o\}, F,\left\{\mathrm{f} \mapsto o^{\rho(\mathrm{f})} \rightarrow o \mid \mathrm{f} \in\right.\right.$ $F\}\rangle$. We often identify a ranked alphabet $\langle F, \rho\rangle$ with the corresponding tree signature $\Sigma_{\langle F, \rho\rangle}$. Then $\mathbb{T}(\Sigma)$ is the subset of $\Lambda(\Sigma)$ whose elements are variable-free and of atomic types. The yield string yield $(M)$ of a tree $M$ is defined as

$$
\begin{aligned}
\operatorname{yield}(\mathrm{f}) & =\mathrm{f} & & \text { if } \mathrm{f} \in F^{(0)}, \\
\operatorname{yield}\left(\mathrm{f} M_{1} \ldots M_{n}\right) & =\operatorname{yield}\left(M_{1}\right) \ldots \operatorname{yield}\left(M_{n}\right) & & \text { if } \mathrm{f} \in F^{(n)} \text { with } n \geq 1 .
\end{aligned}
$$

### 2.1.2 Abstract Categorial Grammars

Definition 2.2 (Lexicon). For two sets of atomic types $\mathscr{A}_{0}$ and $\mathscr{A}_{1}$, a type substitution $\sigma$ is a mapping from $\mathscr{A}_{0}$ to $\mathcal{T}\left(\mathscr{A}_{1}\right)$, which can be extended homomorphically as

$$
\sigma(\alpha \rightarrow \beta)=\sigma(\alpha) \rightarrow \sigma(\beta) .
$$

For two higher-order signatures $\Sigma_{0}$ and $\Sigma_{1}$, a term substitution $\theta$ is a mapping from $\mathscr{C}_{0}$ to $\Lambda\left(\Sigma_{1}\right)$ such that $\theta(\mathrm{a})$ is closed for all a $\in \mathscr{C}_{0}$. For two higher-order signatures $\Sigma_{0}$ and $\Sigma_{1}$, we say that a type substitution $\sigma: \mathscr{A}_{0} \rightarrow \mathcal{T}\left(\mathscr{A}_{1}\right)$ and a term substitution $\theta: \mathscr{C}_{0} \rightarrow \Lambda\left(\Sigma_{1}\right)$ are compatible iff $\sigma\left(\tau_{0}(\mathrm{a})\right)=\tau_{1}(\theta(\mathrm{a}))$ holds for all a $\in \mathscr{C}_{0}$. A lexicon from $\Sigma_{0}$ to $\Sigma_{1}$ is a compatible pair of a type substitution and a term substitution. On a lexicon $\mathscr{L}=\langle\sigma, \theta\rangle$, the function $\widehat{\mathscr{L}}$ from $\Lambda\left(\Sigma_{0}\right)$ to $\Lambda\left(\Sigma_{1}\right)$ is uniquely determined as follows:

$$
\begin{aligned}
\widehat{\mathscr{L}}\left(x^{\alpha}\right) & =x^{\sigma(\alpha)} \quad \text { for } \alpha \in \mathcal{T}\left(\mathscr{A}_{0}\right), \\
\widehat{\mathscr{L}}(\mathrm{a}) & =\theta(\mathrm{a}) \quad \text { for } \mathrm{a} \in \mathscr{C}_{0}, \\
\widehat{\mathscr{L}}\left(\lambda x^{\alpha} \cdot M\right) & =\lambda x^{\sigma(\alpha)} \cdot \widehat{\mathscr{L}}(M) \quad \text { for } \lambda x^{\alpha} \cdot M \in \Lambda\left(\Sigma_{0}\right), \\
\widehat{\mathscr{L}}(M N) & =\widehat{\mathscr{L}}(M) \widehat{\mathscr{L}}(N) \quad \text { for } M N \in \Lambda\left(\Sigma_{0}\right) .
\end{aligned}
$$

Indeed, $\widehat{\mathscr{L}}(M)$ always a well-typed $\lambda$-term if so is $M$; if $M$ has type $\alpha$, then $\widehat{\mathscr{L}}(M)$ has type $\sigma(\alpha)$. Hereafter we simply write $\mathscr{L}$ for $\sigma$ or $\widehat{\mathscr{L}}$ depending on the context. A lexicon $\mathscr{L}=\langle\sigma, \theta\rangle$ is $n$-th $\operatorname{order}$ if $\operatorname{ord}(\mathscr{L})=$ $\max \left\{\operatorname{ord}(\sigma(p)) \mid p \in \mathscr{A}_{0}\right\} \leq n$. For two lexicons $\mathscr{L}_{1}=\left\langle\sigma_{1}, \theta_{1}\right\rangle: \Sigma_{0} \rightarrow \Sigma_{1}$ and $\mathscr{L}_{2}=\left\langle\sigma_{2}, \theta_{2}\right\rangle: \Sigma_{1} \rightarrow \Sigma_{2}$, it is easily seen that $\sigma_{2} \circ \sigma_{1}$ and $\theta_{2} \circ \theta_{1}$ are also compatible. Thus, the composite lexicon $\mathscr{L}_{2} \circ \mathscr{L}_{1}: \Sigma_{0} \rightarrow \Sigma_{2}$ can be defined as $\mathscr{L}_{2}=\left\langle\sigma_{2} \circ \sigma_{1}, \theta_{2} \circ \theta_{1}\right\rangle$, where it holds that $\operatorname{ord}\left(\mathscr{L}_{2} \circ \mathscr{L}_{1}\right) \leq$ $\operatorname{ord}\left(\mathscr{L}_{1}\right)+\operatorname{ord}\left(\mathscr{L}_{2}\right)-1$.

Definition 2.3 (de Groote [15]). An abstract categorial grammar (ACG) is a quadruple $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$, where

- $\Sigma_{0}$ is a higher-order signature, called the abstract vocabulary,
- $\Sigma_{1}$ is a higher-order signature, called the object vocabulary,
- $\mathscr{L}$ is a lexicon from $\Sigma_{0}$ to $\Sigma_{1}$,
- $s \in \mathscr{A}_{0}$ is called the distinguished type.

When we use the modifier abstract or object, it specifies the vocabulary that a given type or term belongs to. Thus, we speak of abstract atomic types,
object atomic types, abstract constants, object constants, etc. We sometimes call the triple $\left\langle\mathrm{a}, \tau_{0}(\mathrm{a}), \mathscr{L}(\mathrm{a})\right\rangle$ for $\mathrm{a} \in \mathscr{C}_{0}$ a lexical entry, and specify an ACG by giving the set of lexical entries and the distinguished type.

Definition 2.4. An ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ generates two languages, the abstract language $\mathcal{A}(\mathscr{G})$ and the object language $\mathcal{O}(\mathscr{G})$, defined as follows:

$$
\begin{aligned}
& \mathcal{A}(\mathscr{G})=\left\{M \in \Lambda\left(\Sigma_{0}\right) \mid M \text { is a closed } \beta \eta \text {-normal term of type } s\right\} \\
& \mathcal{O}(\mathscr{G})=\left\{|\mathscr{L}(M)|_{\beta \eta} \mid M \in \mathcal{A}(\mathscr{G})\right\}
\end{aligned}
$$

The abstract language can be thought as a set of abstract grammatical structures, and the object language is regarded as the set of concrete forms obtained from these abstract structures and the lexicon. Thus, we simply say the language generated by an ACG for its object language. The term abstract categorial languages (ACLs) means the object languages of ACGs. Two ACGs are equivalent iff their object languages coincide. For conciseness we write $M \in \mathcal{A}(\mathscr{G})(P \in \mathcal{O}(\mathscr{G}))$ instead of $|M|_{\beta \eta} \in \mathcal{A}(\mathscr{G})\left(|P|_{\beta \eta} \in \mathcal{O}(\mathscr{G})\right)$ even if $M(P)$ is not $\beta \eta$-normal.

A string $A C G$ is an ACG whose object vocabulary is a string signature and whose distinguished type is mapped to the type str. A tree $A C G$ is an ACG whose object vocabulary is a tree signature and whose distinguished type is mapped to the type $o$. By $\mathbf{G}_{\text {string }}$ and $\mathbf{G}_{\text {tree }}$, we denote the classes of string ACGs and tree ACGs, respectively.

Let us denote by $\mathbf{G}(m, n)$ the subclass of ACGs $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ such that $\operatorname{ord}\left(\Sigma_{0}\right) \leq m$ and $\operatorname{ord}(\mathscr{L}) \leq n$. An ACG in $\mathbf{G}(m, n)$ is also called an $m$-th order $A C G$. For a second-order ACG, its lexicon is said to be semirelabeling if it is first-order and it maps each abstract constant to an object constant or the identity $\lambda z^{o} . z$. A relabeling lexicon is a semi-relabeling lexicon which maps every abstract constant to an object constant. $\mathbf{G}(2,1(\mathrm{sr}))$ and $\mathbf{G}(2,1(r))$ denote the class of second-order ACGs whose lexicons are semirelabeling and relabeling, respectively. A second-order lexicon $\mathscr{L}$ is said to be monadic if for each abstract atomic type $p \in \mathscr{A}_{0}, \mathscr{L}(p)$ is either an atomic or a unary type. $\mathbf{G}(2,2(\mathrm{mon}))$ is the class of second-order ACGs whose lexicons are monadic.

We specify a subclass of ACGs by combining these notations; for instance, $\mathrm{G}_{\text {tree }}(2,1(\mathrm{r}))$ denotes the class of tree ACGs whose abstract vocabulary is second-order and lexicon is relabeling.

Example 2.5. Let $\mathscr{G}_{1}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}_{1}, s\right\rangle \in \mathbf{G}_{\text {string }}$ consist of the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}_{1}(\mathrm{x})$ |
| :---: | :---: | :---: |
| M | $n$ | $/ \mathrm{man} /$ |
| W | $n$ | $/$ woman $/$ |
| L | $n p^{2} \rightarrow s$ | $\lambda y x \cdot x+/$ loves $/+y$ |
| A | $n \rightarrow(n p \rightarrow s) \rightarrow s$ | $\lambda y w \cdot w(/ \mathrm{a} /+y)$ |
| E | $n \rightarrow(n p \rightarrow s) \rightarrow s$ | $\lambda x w \cdot w(/$ every $/+x)$ |

where $\mathscr{L}(s)=\mathscr{L}(n p)=\mathscr{L}(n)=s t r$.
Moreover, let $\mathscr{G}_{2}=\left\langle\Sigma_{0}, \Sigma_{2}, \mathscr{L}_{2}, s\right\rangle$ consist of the following lexical entries: ${ }^{1}$

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}_{2}(\mathrm{x})$ |
| :---: | :---: | :---: |
| M | $n$ | $\lambda z \cdot \operatorname{man} z$ |
| W | $n$ | $\lambda z \cdot \mathbf{\text { woman } z}$ |
| L | $n p^{2} \rightarrow s$ | $\lambda y \cdot$ love $y x$ |
| A | $n \rightarrow(n p \rightarrow s) \rightarrow s$ | $\lambda w_{1} w_{2} \cdot \exists y\left(\left(w_{1} y\right) \wedge\left(w_{2} y\right)\right)$ |
| E | $n \rightarrow(n p \rightarrow s) \rightarrow s$ | $\lambda w_{1} w_{2} \cdot \forall x\left(\left(w_{1} x\right) \rightarrow\left(w_{2} x\right)\right)$ |

where $\mathscr{L}(s)=t, \mathscr{L}(n p)=e, \mathscr{L}(n)=e \rightarrow t$. Let

$$
\begin{aligned}
M & =\operatorname{AW}\left(\lambda x^{n p} \cdot \operatorname{EM}\left(\lambda y^{n p} \cdot \mathrm{~L} x y\right)\right), \\
N & =\operatorname{EM}\left(\lambda y^{n p} \cdot \operatorname{AW}\left(\lambda x^{n p} \cdot \operatorname{Lxy}\right)\right) .
\end{aligned}
$$

We have

$$
\begin{aligned}
\mathscr{L}_{1}(M) & =\mathscr{L}_{1}(N)=/ \text { every man loves a woman } \\
\mathscr{L}_{2}(M) & =\exists y((\operatorname{woman} y) \wedge(\forall x((\operatorname{man} x) \rightarrow(\operatorname{love} x y)))) \\
\mathscr{L}_{2}(N) & =\forall x((\operatorname{man} x) \rightarrow(\exists y((\operatorname{woman} y) \wedge(\operatorname{love} x y)))) .
\end{aligned}
$$

This way the above two ACGs give an explanation of the ambiguity of the scopes of the quantifiers.

### 2.2 Basic Properties of Abstract Categorial Grammars

### 2.2.1 Hierarchy of Second-Order ACGs

De Groote and Pogodalla [19] have shown that a variety of context-free formalisms can be encoded by second-order ACGs in a straightforward way.

[^0]Namely, string languages generated by context-free grammars (CFGs), linear context-free grammars (LCFTGs), linear context-free rewriting systems (LCFRSs) are all generated by ACGs belonging to $\mathbf{G}_{\text {string }}(2,2), \mathbf{G}_{\text {string }}(2,3)$, $\mathbf{G}_{\text {string }}(2,4)$ respectively. Here, we let the term linear CFTGs mean nonduplicating non-deleting CFTGs following the terminology by de Groote and Pogodalla, although usually "linearity" means just non-duplication. Their encodings are straightforward in the sense that the derivation structures in the original grammar are preserved in the resultant ACG. For the converse, it is not difficult to see that ACGs in $\mathbf{G}_{\text {string }}(2,2)$ can generate only contextfree languages. The most remarkable result on the hierarchy of second-order string ACGs is given by Salvati [45]. The hierarchy collapses from fourthorder.

Theorem 2.6 (Salvati [45]). Every second-order string ACG in $\mathbf{G}_{\text {string }}(2, n)$ for $n \geq 1$ has an equivalent LCFRS, and thus an equivalent $A C G$ belonging to $\mathbf{G}_{\text {string }}(2,4)$.

It is conjectured that the languages generated by ACGs in $\mathbf{G}_{\text {string }}(2,3)$ are also generated by LCFTGs. If that is the case, these results give us a precise correspondence between $\mathbf{G}_{\text {string }}(2, n)$ and context-free formalisms for each $n$.

Let us turn our attention to the generative capacity of second-order tree ACGs. The equivalence between $\mathbf{G}_{\text {tree }}(2,1)\left(\mathbf{G}_{\text {tree }}(2,1(\mathrm{r}))\right)$ and regular tree grammars is easily seen. As discussed by de Groote and Pogodalla [19], linear context-free tree languages can be generated by ACGs in $\mathbf{G}_{\text {tree }}(2,2)$ (see also [27] for a complete proof). Moreover it is easy to see that the converse of the relation also holds. De Groote [18] has shown that treeadjoining grammars are also simulated by ACGs in $\mathbf{G}_{\text {tree }}(2,2(\mathrm{mon}))$. The converse relation also holds with respect to the string generating capacity.

Those correspondences are summarized in Table 2.1.
Definition 2.7 (Linear Context-Free Tree Grammar). A context-free tree grammar is a quadruple $G=\langle\langle V, \rho\rangle, T, R, S\rangle$ where $\langle V, \rho\rangle$ is a ranked alphabet, elements of $T \subsetneq V$ are called terminal symbols, elements of $V-T$ are called nonterminal symbols, $S \in V-T$ is the start symbol with $\rho(S)=0$, and $P$ is a finite set of productions of the form

$$
A\left[x_{1}, \ldots, x_{\rho(A)}\right] \rightarrow N
$$

where $A \in V-T$ and $N \in \mathbb{T}\left(\langle V, \rho\rangle \cup\left\{x_{1}, \ldots, x_{\rho(A)}\right\}\right)$. A CFTG is nonduplicating iff for every production of the above form, $x_{i}$ appears in $N$ at most once for each $x_{1}, \ldots, x_{\rho(A)}$. A CFTG is non-deleting iff for every production

Table 2.1: Hierarchy of second-order ACGs

| Finite Languages | $=$ | $\mathbf{G}(1, m)$ |  |
| :---: | :---: | :--- | :---: |
| String Languages |  |  |  |
| Context-Free Grammars | $=$ | $\mathbf{G}_{\text {string }}(2,2)$ |  |
| Linear Context Free Tree Grammars | $\subseteq$ | $\mathbf{G}_{\text {string }}(2,3)$ |  |
| Linear Context Free Rewriting Systems | $=$ | $\mathbf{G}_{\text {string }}(2, n)$ for $n \geq 4$ |  |
| Tree Languages |  |  |  |
| Regular Tree Grammars | $=$ | $\mathbf{G}_{\text {tree }}(2,1)\left(\mathbf{G}_{\text {tree }}(2,1(\mathrm{r}))\right)$ |  |
| Linear Context-Free Tree Grammars | $=$ | $\mathbf{G}_{\text {tree }}(2,2)$ |  |
| Tree Adjoining Grammars | $\subseteq$ | $\mathbf{G}_{\text {tree }}(2,2($ mon $))$ |  |

rule of the above form, $x_{i}$ appears in $N$ at least once for each $x_{1}, \ldots, x_{\rho(A)}$. A CFTG is called linear iff it is non-duplicating and non-deleting at the same time.

We write

$$
M \Rightarrow M^{\prime}
$$

iff there are $A \in V-T, M_{0} \in \mathbb{T}(\langle V, \rho\rangle \cup\{z\})$ ( $z$ occurs exactly one in $\left.M_{0}\right), M_{1}, \ldots, M_{\rho(A)} \in \mathbb{T}(\langle V, \rho\rangle)$, and $A\left[x_{1}, \ldots, x_{\rho(A)}\right] \rightarrow N \in R$ such that $M=M_{0}\left[A M_{1} \ldots M_{\rho(A)} / z\right]$ and $M^{\prime}=M_{0}\left[N\left[M_{i} / x_{i}\right]_{1 \leq i \leq \rho(A)} / z\right]$. The (tree) language generated by $G$ is defined by

$$
L(G)=\{M \in \mathbb{T}(\langle T, \rho\rangle) \mid S \stackrel{*}{\Rightarrow} M\}
$$

where $\stackrel{*}{\Rightarrow}$ is the reflexive transitive closure of $\Rightarrow$.
De Groote and Pogodalla [19] encode an LCFTG $G=\langle\langle V, \rho\rangle, T, R, S\rangle$ as an ACG $\mathscr{G}^{G}=\left\langle\Sigma_{0}, \Sigma_{\langle T, \rho\rangle}, \mathscr{L}, S\right\rangle$ as follows. The object vocabulary is the tree signature $\Sigma_{\langle T, \rho\rangle}$, and the set $\mathscr{A}_{0}$ of abstract atomic types is the set $V-T$ of nonterminals. An abstract atomic type $A \in \mathscr{A}_{0}=V-T$ is mapped to $o^{\rho(A)} \rightarrow o$ by the lexicon $\mathscr{L}$. For each production rule $r: A\left[x_{1}, \ldots, x_{\rho(A)}\right] \rightarrow$ $N$, we put the following lexical entry into $\mathscr{G}^{G}$ :

$$
\left\langle\mathrm{A}_{r}, B_{1} \rightarrow \cdots \rightarrow B_{k} \rightarrow A, \lambda y_{1} \ldots y_{k} x_{1} \ldots x_{\rho(A)} \cdot N^{\prime}\right\rangle
$$

where $B_{1}, \ldots, B_{k}$ is a sequence of nonterminals occurring in $N$ (if the same nonterminal $B$ appears in $N m$-times, $B$ occurs in the sequence $m$-times) and $N^{\prime}$ is obtained from $N$ by replacing each $B_{i}$ with $y_{i}$.

Example 2.8. Let an LCFTG $G$ consist of the following production rules:

$$
\begin{aligned}
S & \rightarrow A \mathrm{abc}, \\
A\left[x_{1}, x_{2}, x_{3}\right] & \rightarrow A\left(\mathrm{fa} x_{1}\right)\left(\mathrm{fb} x_{2}\right)\left(\mathrm{fc} x_{3}\right), \\
A\left[x_{1}, x_{2}, x_{3}\right] & \rightarrow \mathrm{g} x_{1} x_{2} x_{3},
\end{aligned}
$$

where the ranks of $S, A$, a, b, c, f, g, are $0,3,0,0,0,2,3$, respectively. $S$ and $A$ are nonterminals and $\mathrm{a}, \mathrm{b}, \mathrm{c}, \mathrm{f}, \mathrm{g}$ are terminals. The ACG $\mathscr{G}^{G}$ that encodes $G$ consists of the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}(\mathrm{x})$ |
| :---: | :---: | :---: |
| $\mathrm{A}_{1}$ | $A \rightarrow S$ | $\lambda y_{A}^{o^{3} \rightarrow o} \cdot y_{A} \mathrm{abc}$ |
| $\mathrm{A}_{2}$ | $A \rightarrow A$ | $\lambda y_{A}^{\sigma^{3} \rightarrow o} x_{1}^{o} x_{2}^{o} x_{3}^{o} \cdot y_{A}\left(\mathrm{fa} x_{1}\right)\left(\mathrm{fb} x_{2}\right)\left(\mathrm{fc} x_{3}\right)$ |
| $\mathrm{A}_{3}$ | $A$ | $\lambda x_{1}^{o} x_{2}^{o} x_{3}^{o} . \mathrm{g} x_{1} x_{2} x_{3}$ |

Definition 2.9 (Linear Context-Free Rewriting System). A contextfree rewriting system (CFRS) is a quadruple $G=\langle\langle V, \rho\rangle, T, R, S\rangle$ where $\langle V, \rho\rangle$ is a ranked alphabet called the set of nonterminals, $T$ is an (unranked) alphabet disjoint from $V$ called the set of terminal symbols, $S \in V$ is the start symbol with $\rho(S)=1$, and $R$ is a finite set of productions of the form

$$
A \rightarrow f\left(B_{1}, \ldots, B_{m}\right)
$$

where $A, B_{1}, \ldots, B_{m} \in V$ and $f$ is a function from $\left(T^{*}\right)^{\rho\left(B_{1}\right)} \times \cdots \times\left(T^{*}\right)^{\rho\left(B_{m}\right)}$ to $\left(T^{*}\right)^{\rho(A)}$ such that

$$
f\left(\left\langle x_{1,1}, \ldots, x_{1, \rho\left(B_{1}\right)}\right\rangle, \ldots,\left\langle x_{m, 1}, \ldots, x_{m, \rho\left(B_{m}\right)}\right\rangle\right)=\left\langle\mathrm{w}_{1}, \ldots, \mathrm{w}_{\rho(A)}\right\rangle
$$

with $\mathrm{w}_{i} \in\left(T \cup\left\{x_{1,1}, \ldots, x_{m, \rho\left(B_{m}\right)}\right\}\right)^{*}$. For $f$ as above, $f_{i}$ for $i \in\{1, \ldots, \rho(A)\}$ means the $i$-th project of $f$, i.e., $f_{i}\left(\vec{x}_{1}, \ldots, \vec{x}_{m}\right)=\mathrm{w}_{i}$ for $\vec{x}_{i}=\left\langle x_{i, 1}, \ldots, x_{i, \rho(i)}\right\rangle$. A CFRS is linear iff for every function appearing in a production rule as the above form, $x_{i, j}$ appears in $f\left(\vec{x}_{1}, \ldots, \vec{x}_{m}\right)$ exactly once for each $i \in\{1, \ldots, m\}$ and $j \in\left\{1, \ldots, \rho\left(B_{i}\right)\right\}$.

The set $L(G, A) \subseteq\left(T^{*}\right)^{\rho(A)}$ for $A \in V$ is defined as follows:

- If $A \rightarrow f() \in R\left(f\right.$ is a constant function and $\left.f() \in\left(T^{*}\right)^{\rho(A)}\right)$, then $f() \in L(G, A)$.
- if $A \rightarrow f\left(B_{1}, \ldots, B_{m}\right) \in R$, and $\overrightarrow{\mathrm{w}}_{i} \in L\left(G, B_{i}\right)$ for $i \in\{1, \ldots, m\}$, then $f\left(\overrightarrow{\mathrm{w}}_{1}, \ldots, \overrightarrow{\mathrm{w}}_{m}\right) \in L(G, A)$.

The language generated by $G$ is defined as

$$
L(G)=\left\{\mathrm{w} \in T^{*} \mid\langle\mathrm{w}\rangle \in L(G, S)\right\} .
$$

De Groote and Pogodalla [19] encode an LCFRS $G=\langle\langle V, \rho\rangle, T, R, S\rangle$ as an ACG $\mathscr{G}^{G}=\left\langle\Sigma_{0}, \Sigma_{T}, \mathscr{L}, s^{\prime}\right\rangle$ as follows. The object vocabulary is the string signature $\Sigma_{T}$, and the set $\mathscr{A}_{0}$ of abstract atomic types is $V \cup\left\{s^{\prime}\right\}$ with $s^{\prime} \notin V$. An abstract atomic type $A \in V$ is mapped to $\left(\operatorname{str}^{\rho(A)} \rightarrow s t r\right) \rightarrow s t r$ and the distinguished type $s^{\prime}$ is mapped to str by the lexicon $\mathscr{L}$. For each production rule $r: A \rightarrow f\left(B_{1}, \ldots, B_{k}\right)$, we put a lexical entry representing the rule:

$$
\left\langle\mathrm{c}_{r}, B_{1} \rightarrow \cdots \rightarrow B_{k} \rightarrow A, \lambda y_{1} \ldots y_{k} z \cdot y_{1}\left(\lambda \vec{x}_{1} \ldots y_{k}\left(\lambda \vec{x}_{k} \cdot z M_{1} \ldots M_{\rho(A)}\right) \ldots\right)\right\rangle
$$

where

$$
M_{i}=/ f_{i}\left(\vec{x}_{1}, \ldots, \vec{x}_{k}\right) /
$$

Moreover, we put the following special lexical entry:

$$
\left\langle\mathrm{S}, S \rightarrow s^{\prime}, \lambda y^{(s t r \rightarrow s t r) \rightarrow s t r} \cdot y\left(\lambda \vec{x}^{s t r} \cdot x\right)\right\rangle .
$$

Then, whenever $\left\langle\mathrm{w}_{1}, \ldots, \mathrm{w}_{\rho(A)}\right\rangle \in L(G, A)$, we can find an abstract term $M \in \Lambda\left(\Sigma_{0}\right)$ of type $A$ such that $\mathscr{L}(\mathrm{A})=\lambda z^{(\operatorname{str} \rho(A) \rightarrow \operatorname{str}) \rightarrow \operatorname{str} . z / \mathrm{w}_{1} / \ldots / \mathrm{w}_{\rho(A)} / .}$

Example 2.10. Let an LCFRS $G$ consist of the following production rules:

$$
\begin{array}{llrl}
S \rightarrow f(A) & \text { with } & f\left(\left\langle x_{1}, x_{2}, x_{3}\right\rangle\right) & =\left\langle x_{1} x_{2} x_{3}\right\rangle \\
A \rightarrow g(A) & \text { with } & g\left(\left\langle x_{1}, x_{2}, x_{3}\right\rangle\right) & =\left\langle\mathrm{a} x_{1}, \mathrm{~b} x_{2}, \mathrm{c} x_{3}\right\rangle \\
A \rightarrow h() & \text { with } & h() & =\langle\mathrm{a}, \mathrm{~b}, \mathrm{c}\rangle
\end{array}
$$

The ACG that encodes $G$ consists of the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}(\mathrm{x})$ |
| :---: | :---: | :---: |
| F | $A \rightarrow S$ | $\lambda y_{A} z \cdot y_{A}\left(\lambda x_{1} x_{2} x_{3} . z\left(x_{1}+x_{2}+x_{3}\right)\right)$ |
| G | $A \rightarrow A$ | $\lambda y_{A} z . y_{A}\left(\lambda x_{1} x_{2} x_{3} \cdot z\left(\mathrm{a}+x_{1}\right)\left(\mathrm{b}+x_{2}\right)\left(\mathrm{c}+x_{3}\right)\right)$ |
| H | $A$ | $\lambda z . z \mathrm{abc}$ |
| S | $S \rightarrow s^{\prime}$ | $\lambda y_{S} \cdot y_{S}(\lambda x . x)$ |

A more general result on the generative capacity of second-order ACG is given by Salvati.

Theorem 2.11 (Salvati [43]). Every second-order ACG generates a PTIME language.

### 2.2.2 General Properties of ACGs

There are some results on the mathematical properties of ACGs which are themselves, or are corollaries to, already known facts. These properties demonstrate the rich expressive power of ACGs. However, some fundamental mathematical properties of ACGs are not yet to be revealed. For instance, no recursively enumerable language has been found that cannot be generated by any ACG.

Lemma 2.12. Let a type substitution $\sigma:\{o\} \rightarrow \mathcal{T}(\{o\})$ be defined as $\sigma(o)=$ str. Then, for any $\alpha \in \mathcal{T}(\{o\})$, there is a linear combinator $Z^{\sigma(\alpha)}$ of type $\sigma(\alpha)$.

Proof. By induction on $\alpha$, we define a linear combinator $Z^{\sigma(\alpha)}$ of type $\sigma(\alpha)$. Let $\sigma(\alpha)=\beta_{1} \rightarrow \cdots \rightarrow \beta_{m} \rightarrow$ str and $\beta_{i}=\beta_{i, 1} \rightarrow \cdots \rightarrow \beta_{i, k_{i}} \rightarrow$ str. Define

$$
Z^{\sigma(\alpha)}=\lambda y_{1}^{\beta_{1}} \ldots y_{m}^{\beta_{m}} z^{o} \cdot R_{1}\left(R_{2}\left(\ldots\left(R_{m} z\right) \ldots\right)\right)
$$

$$
\text { where } R_{i}=y_{i}^{\beta_{i}} Z^{\beta_{i, 1}} \ldots Z^{\beta_{i, k_{i}}} .
$$

Hereafter by $Z^{\beta}$ we denote such linear combinator of type $\beta$ provided that $\beta$ has the above form.
Definition 2.13. The universal membership problem is the problem of determining whether $P \in \mathcal{O}(\mathscr{G})$. The non-emptiness problem for ACGs is the problem of determining whether $\mathcal{O}(\mathscr{G}) \neq \varnothing$.

Though these problems are fundamental, they are still open. We have only partial answers to them.

Proposition 2.14. The universal membership problem for ACGs is at least as hard as the non-emptiness problem for ACGs.

Proof. We present a reduction from the non-emptiness problem to the universal membership problem. Let an ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ be given as an instance of the non-emptiness problem. We define $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}, \Sigma_{1}^{\prime}, \mathscr{L}^{\prime}, s\right\rangle$ as

$$
\begin{aligned}
\Sigma_{1}^{\prime} & =\langle\{o\}, \varnothing, \varnothing\rangle \\
\mathscr{L}^{\prime}(p) & =s t r \text { for all } p \in \mathscr{A}_{0} \\
\mathscr{L}^{\prime}(\mathrm{c}) & =Z^{\mathscr{L}^{\prime}\left(\tau_{0}(\mathrm{c})\right)} \text { for all } \mathrm{c} \in \mathscr{C}_{0},
\end{aligned}
$$

where $Z^{\alpha}$ is as in Lemma 2.12. Then, $\mathscr{L}^{\prime}(M) \rightarrow_{\beta} \lambda z^{o} . z$ for every $M \in$ $\mathcal{A}\left(\mathscr{G}^{\prime}\right)=\mathcal{A}(\mathscr{G})$. Therefore, $\mathcal{O}(\mathscr{G}) \neq \varnothing$ iff $\lambda z^{o} . z \in \mathcal{O}\left(\mathscr{G}^{\prime}\right)$.

The following proposition is an easy corollary to a result obtained by de Groote et al. [17]

Proposition 2.15. The emptiness problem for ACGs is decidable iff the multiplicative exponential linear logic (MELL) is decidable.

Proof. Let $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ be an ACG, where $\Sigma_{0}=\left\langle\mathscr{A}_{0}, \mathscr{C}_{0}, \tau_{0}\right\rangle, \mathscr{C}_{0}=$ $\left\{\mathrm{c}_{1}, \ldots, \mathrm{c}_{n}\right\}$, and $\tau_{0}\left(\mathrm{c}_{i}\right)=A_{i}$ for $1 \leq i \leq n$. Then $\mathcal{O}(\mathscr{G}) \neq \varnothing$ iff $\mathcal{A}(\mathscr{G}) \neq \varnothing$ iff $!A_{1}, \ldots,!A_{n} \Rightarrow s$ is provable in MELL. This proves the "if" direction.

The "only if" direction can be proved as follows. De Groote et al. [17] show that the decidability of MELL is equivalent to the decidability of a fragment of it called $\mathbf{I M E L L}_{\mathbf{0}}{ }^{-}$. Formulas of $\mathbf{I M E L L}_{\mathbf{0}}{ }^{-\circ}$ are of the form ! $A$ or $A$, where $A$ is a pure implicative formula, and the right-hand side of a sequent of $\mathbf{I M E L L}_{\mathbf{0}}^{-0}$ must be a pure implicative formula. Given a sequent

$$
\mathcal{S}:!A_{1}, \ldots,!A_{n}, B_{1}, \ldots, B_{m} \Rightarrow C
$$

of $\mathbf{I M E L L}_{\mathbf{0}}{ }^{\circ}$ where each $B_{i}$ is pure implicative, let an ACG $\mathscr{G} \mathcal{S}=\left\langle\Sigma_{0}, \Sigma_{0}\right.$, $\left.\mathscr{L}_{\text {id }}, s\right\rangle$ be such that $\Sigma_{0}=\left\langle\mathscr{A}_{\mathcal{S}} \cup\{s\}, \mathscr{C}_{0}, \tau_{0}\right\rangle, \mathscr{A}_{\mathcal{S}}$ is the set of atomic formulas in the sequent $\mathcal{S}, s \notin \mathscr{A}_{\mathcal{S}}, \mathscr{C}_{0}=\left\{\mathrm{c}_{i} \mid 1 \leq i \leq n\right\} \cup\{\mathrm{b}\}, \tau_{0}\left(\mathrm{c}_{i}\right)=A_{i}$, $\tau_{0}(\mathrm{~b})=\left(B_{1} \rightarrow \cdots \rightarrow B_{m} \rightarrow C\right) \rightarrow s$, and $\mathscr{L}_{\text {id }}$ is the identity. Then $\mathcal{S}$ is provable in IMELL $_{0}^{-0}$ iff $\mathcal{S}$ is provable in MELL iff $\mathcal{A}\left(\mathscr{G}^{\mathcal{S}}\right) \neq \varnothing$ iff $\mathcal{O}\left(\mathscr{G}^{\mathcal{S}}\right) \neq \varnothing$.

Restricting to string ACGs does not change the situation. We can find a string ACG $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}^{\prime}, s\right\rangle$ such that $\mathcal{A}\left(\mathscr{G}^{\prime}\right)=\mathcal{A}\left(\mathscr{G}^{\mathcal{S}}\right)$. For instance, let $\Sigma_{1}=\langle\{o\}, \varnothing, \varnothing\rangle, \mathscr{L}^{\prime}(p)=\operatorname{str}$ for all $p \in \mathscr{A}_{\mathcal{S}} \cup\{s\}$ and let $\mathscr{L}^{\prime}(\mathrm{c})=Z^{\mathscr{L}\left(\tau_{0}(\mathrm{c})\right)}$ for all $\mathrm{c} \in \mathscr{C}_{0}$.

The decidability of MELL is still open but it is known to be at least as hard as Petri-net reachability, which is at least EXPSPACE-hard [31]. Indeed, we can represent Petri-net reachability sets by ACGs directly using a reduction from vector addition systems (VASs), which are equivalent to Petri-nets.

Definition 2.16. An $m$-dimensional vector addition system ( $m$-VAS) $\mathscr{V}$ is a pair $\langle\Delta, \vec{s}\rangle$, where $\Delta$ is a finite subset of $\mathbb{Z}^{m}$ and $\vec{s} \in \mathbb{N}^{m}$. We call $\mathbb{N}^{m}$ the set of configurations. $\Rightarrow_{\vec{d}}$ for $\vec{d} \in \mathbb{Z}^{m}$ is a binary relation on $\mathbb{N}^{m}$ such that for $\vec{a}, \vec{b} \in \mathbb{N}^{m}, \vec{a} \Rightarrow_{\vec{d}} \vec{b}$ iff $\vec{b}=\vec{a}+\vec{d}$. The union of $\Rightarrow_{\vec{d}}$ for $\vec{d} \in \Delta$ is denoted by $\Rightarrow_{\Delta}$. A configuration $\vec{b}$ is reachable iff $\vec{s} \Rightarrow_{\Delta}^{*} \vec{b}$ where $\Rightarrow_{\Delta}^{*}$ is the reflexive, transitive closure of $\Rightarrow_{\Delta}$. The reachability set $\mathcal{R}(\mathscr{V}) \subseteq \mathbb{N}^{m}$ of an m-VAS $\mathscr{V}$ is the set of reachable configurations.

Definition 2.17. For an alphabet $V$, suppose that its elements are ordered as $\mathrm{a}_{1}, \ldots, \mathrm{a}_{n}$. The Parikh vector of a string $\mathrm{w} \in V^{*}$ is defined as

$$
\#(\mathrm{w})=\left\langle \#_{\mathrm{a}_{1}}(\mathrm{w}), \ldots, \#_{\mathrm{a}_{n}}(\mathrm{w})\right\rangle \in \mathbb{N}^{n} .
$$

and the Parikh image of a language $L \subseteq V^{*}$ is defined as

$$
\#(L)=\{\#(\mathrm{w}) \mid \mathrm{w} \in L\} \subseteq \mathbb{N}^{n}
$$

A set of vectors of natural numbers $\Gamma \subseteq \mathbb{N}^{m}$ is linear iff there are $\vec{a}_{1}, \ldots, \vec{a}_{n}, \vec{b} \in \mathbb{N}^{m}$ such that $\Gamma=\left\{\vec{b}+\sum_{j=1}^{n} k_{j} \vec{a}_{j} \mid k_{j} \in \mathbb{N}\right\} . \Gamma \subseteq \mathbb{N}^{m}$ is semilinear iff there are linear sets $\Gamma_{1}, \ldots, \Gamma_{n} \subseteq \mathbb{N}^{m}$ such that $\Gamma=\bigcup_{i=1}^{n} \Gamma_{i}$. A language $L$ is linear (semilinear) iff the Parikh image of $L$ is linear (semilinear).

It is known that there is a VAS whose reachability set is not semilinear [21]. The notion of Parikh vector can be applied to $\lambda$-terms on a higherorder signature $\Sigma$; the Parikh vector of $M$ is defined as $\#(M)=\left\langle \# \mathrm{a}_{i}(M)\right|$ $\left.a_{i} \in \mathscr{C}\right\rangle \in \mathbb{N}^{|\mathscr{G}|}$. Then the notions of Parikh image and the semilinearity of a set of terms are defined in the obvious way.
Proposition 2.18. For every reachability set $\mathcal{R}(\mathscr{V})$, there is an $A C G \mathscr{G}^{\mathscr{V}} \in$ $\mathrm{G}_{\text {string }}(3,2)$ such that $\mathcal{R}(\mathscr{V})=\#\left(\mathcal{O}\left(\mathscr{G}^{V}\right)\right)$.

Proof. Given an $m$-VAS $\mathscr{V}=\langle\Delta, \vec{s}\rangle$, we define an ACG $\mathscr{G}^{\mathscr{V}}=\left\langle\Sigma_{0}, \Sigma_{V}, \mathscr{L}, q\right\rangle$ as follows. The set $\mathscr{A}_{0}$ of abstract atomic types is

$$
\mathscr{A}_{0}=\left\{p_{1}, \ldots, p_{m}, q\right\} .
$$

Let a function $\rho$ map $\vec{d}=\left\langle d_{1}, \ldots, d_{m}\right\rangle \in \mathbb{Z}^{m}$ to $\rho(\vec{d}) \in \mathcal{T}\left(\mathscr{A}_{0}\right)$ as

$$
\begin{array}{r}
\rho(\vec{d})=\left(p_{1}^{\nu\left(d_{1}\right)} \rightarrow \cdots \rightarrow p_{m}^{\nu\left(d_{m}\right)} \rightarrow q\right) \rightarrow p_{1}^{\pi\left(d_{1}\right)} \rightarrow \cdots \rightarrow p_{m}^{\pi\left(d_{m}\right)} \rightarrow q \\
\text { where } \nu\left(d_{i}\right)=\left\{\begin{array}{ll}
0 & \text { if } d_{i} \geq 0 \\
-d_{i} & \text { otherwise }
\end{array}, \quad \pi\left(d_{i}\right)= \begin{cases}0 & \text { if } d_{i} \leq 0 \\
d_{i} & \text { otherwise }\end{cases} \right.
\end{array}
$$

The set of abstract constants and the type assignment on them are defined as

$$
\begin{aligned}
\mathscr{C}_{0} & =\left\{\mathrm{a}_{i} \mid 1 \leq i \leq m\right\} \cup\left\{\mathrm{b}_{\vec{d}} \mid \vec{d} \in \Delta\right\} \cup\{\mathrm{c}\}, \\
\tau_{0}\left(\mathrm{a}_{i}\right) & =p_{i}, \quad \tau_{0}\left(\mathrm{~b}_{\vec{d}}\right)=\rho(\vec{d}), \quad \tau_{0}(\mathrm{c})=p_{1}^{s_{1}} \rightarrow \cdots \rightarrow p_{m}^{s_{m}} \rightarrow q
\end{aligned}
$$

where $\vec{s}=\left\langle s_{1}, \ldots, s_{m}\right\rangle$. The object vocabulary is the string signature $\Sigma_{V}$ for the alphabet

$$
V=\left\{\mathrm{e}_{1}, \ldots, \mathrm{e}_{m}\right\} .
$$

The lexicon $\mathscr{L}$ is defined as

$$
\begin{aligned}
& \mathscr{L}\left(p_{i}\right)=\mathscr{L}(q)=s t r, \\
& \mathscr{L}\left(\mathrm{a}_{i}\right)=\mathrm{e}_{i}, \quad \mathscr{L}\left(\mathrm{~b}_{\vec{d}}\right)=Z^{\mathscr{L}(\rho(\vec{d}))}, \quad \mathscr{L}(\mathrm{c})=Z^{\mathscr{L}\left(\tau_{0}(\mathrm{c})\right)} .
\end{aligned}
$$

We show that $\vec{a} \in \mathcal{R}(\mathscr{V})$ iff there is $M \in \mathcal{A}\left(\mathscr{G}^{\mathscr{V}}\right)$ such that $\#(\mathscr{L}(M))=\vec{a}$.
First we show the "only if" part by induction on $n$ where $n$ is such that $\vec{s} \Rightarrow_{\Delta}^{n} \vec{a}$. For $n=0$, i.e., $\vec{a}=\vec{s}$, let

$$
M \equiv \mathrm{ca}_{1}^{s_{1}} \ldots \mathrm{a}_{m}^{s_{m}} \in \mathcal{A}\left(\mathscr{G}^{\mathscr{V}}\right)
$$

Suppose that

$$
\vec{s} \Rightarrow_{\Delta}^{*} \vec{b}=\left\langle b_{1}, \ldots, b_{m}\right\rangle \Rightarrow_{\vec{d}} \vec{a}=\left\langle a_{1}, \ldots, a_{m}\right\rangle
$$

for some $\vec{d}=\left\langle d_{1}, \ldots, d_{m}\right\rangle \in \Delta$. By the induction hypothesis, we have $N \in$ $\mathcal{A}\left(\mathscr{G}^{\mathscr{V}}\right)$ such that $\#(\mathscr{L}(N))=\vec{b}$. Since $a_{i}=b_{i}+d_{i} \geq 0$, we have $b_{i} \geq \nu\left(d_{i}\right)$ for every $i$. In other words, $N$ contains at least $\nu\left(d_{i}\right)$ occurrences of the constant $\mathrm{a}_{i}$ for every $i$. Let $N^{\prime}$ be obtained from $N$ by replacing $\nu\left(d_{i}\right)$ occurrences of the constant $\mathrm{a}_{i}$ with fresh distinct variables $x_{i, 1}, \ldots, x_{i, \nu\left(d_{i}\right)}$ for each $i$, i.e.,

$$
N \equiv N^{\prime}\left[\mathrm{a}_{1} / x_{1,1}, \ldots, \mathrm{a}_{1} / x_{1, \nu\left(d_{1}\right)}, \ldots, \mathrm{a}_{m} / x_{m, 1}, \ldots, \mathrm{a}_{m} / x_{m, \nu\left(d_{m}\right)}\right]
$$

and $N^{\prime}$ contains $b_{i}-\nu\left(d_{i}\right)$ occurrences of $\mathbf{a}_{i}$ for each $i$. Let

$$
M \equiv \mathrm{~b}_{\vec{d}}\left(\lambda \vec{x} \cdot N^{\prime}\right) \mathrm{a}_{1}^{\pi\left(d_{1}\right)} \ldots \mathrm{a}_{m}^{\pi\left(d_{m}\right)} \in \mathcal{A}(\mathscr{G})
$$

where $\vec{x}=\left\langle x_{1,1}, \ldots, x_{1, \nu\left(d_{1}\right)}, \ldots, x_{m, 1}, \ldots, x_{m, \nu\left(d_{m}\right)}\right\rangle$. Since $M$ contains $a_{i}=$ $b_{i}-\nu\left(d_{i}\right)+\pi\left(d_{i}\right)=b_{i}+d_{i}$ occurrences of $\mathrm{a}_{i}$, we have $\#(\mathscr{L}(M))=\vec{a}$.

Let $\mathscr{B}=\left\{\mathrm{b}_{\vec{d}} \in \mathscr{C}_{0} \mid \vec{d} \in \Delta\right\}$. The "if" direction is shown by induction on $\#_{\mathscr{B}}(M)$ for $M \in \mathcal{A}\left(\mathscr{G}^{\mathscr{V}}\right) . M$ must be in the form of either

$$
\begin{align*}
& \mathrm{ca}_{1}^{s_{1}} \ldots \mathrm{a}_{m}^{s_{m}}, \quad \text { or }  \tag{2.1}\\
& \mathrm{b}_{\vec{d}} N \mathrm{a}_{1}^{\pi\left(d_{1}\right)} \ldots \mathrm{a}_{m}^{\pi\left(d_{m}\right)} \text { for some } \vec{d} \in \Delta \text { and } N \in \Lambda\left(\Sigma_{0}\right) . \tag{2.2}
\end{align*}
$$

$M$ has the form (2.1) iff $\# \mathscr{C}_{b}(M)=0$. In this case, clearly $\#(\mathscr{L}(M))=\vec{s}$ is reachable. If $\#_{\mathscr{B}}(M) \geq 1, M$ has the form (2.2). Let $\vec{d}=\left\langle d_{1}, \ldots, d_{m}\right\rangle$. By the type $\rho(\vec{d})$ of $\mathrm{b}_{\vec{d}}$,

$$
\tau_{0}(N)=p_{1}^{\nu\left(d_{1}\right)} \rightarrow \cdots \rightarrow p_{m}^{\nu\left(d_{m}\right)} \rightarrow q
$$

Let

$$
N^{\prime} \equiv N \mathrm{a}_{1}^{\nu\left(d_{1}\right)} \ldots \mathrm{a}_{m}^{\nu\left(d_{m}\right)}
$$

By applying the induction hypothesis to $N^{\prime} \in \mathcal{A}\left(\mathscr{G}^{\mathcal{V}}\right)$, we get that $\#\left(\mathscr{L}\left(N^{\prime}\right)\right)$ is reachable. Since $\# \mathrm{a}_{i}(M)=\#_{\mathrm{a}_{i}}\left(N^{\prime}\right)-\nu\left(d_{i}\right)+\pi\left(d_{i}\right)=\# \mathrm{a}_{i}\left(N^{\prime}\right)+d_{i}$, we have

$$
\vec{s} \Rightarrow_{\Delta}^{*} \#\left(\mathscr{L}\left(N^{\prime}\right)\right) \Rightarrow_{\vec{d}} \#(\mathscr{L}(M)) .
$$

$\#(\mathscr{L}(M))$ is a reachable configuration.

Corollary 2.19. There is an $A C G$ in $\mathbf{G}(3,1)$ whose language is not semilinear.

Proof. We redefine $\mathscr{G}^{\mathscr{V}}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, q\right\rangle$ in the proof of Proposition 2.18 as $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}, \Sigma_{0}, \mathscr{L}_{\text {id }}, q\right\rangle$, where $\mathscr{L}_{\text {id }}$ is the identity. Since $\#\left(\mathcal{O}\left(\mathscr{G}^{\mathscr{V}}\right)\right)$ can be obtained from $\#\left(\mathcal{A}\left(\mathscr{G}^{2}\right)\right)=\#\left(\mathcal{O}\left(\mathscr{G}^{\prime}\right)\right)$ by a projection, and semilinearity is preserved under projections, the non-semilinearity of the former implies the non-semilinearity of the latter.

In contrast, the following result is easily seen.
Proposition 2.20. Every second-order ACG generates a semilinear language.

Proof. First we show that if an ACG $\mathscr{G}$ is second-order, then $\#(\mathcal{A}(\mathscr{G}))$ is semilinear. For the second-order abstract vocabulary $\Sigma_{0}$ of $\mathscr{G}$, let us define a CFG $G^{\Sigma_{0}}$ so that the set of nonterminal symbol is $\mathscr{A}_{0}$, the set of terminal symbols is $\mathscr{C}_{0}$, the start symbol is the distinguished type $s$ of $\mathscr{G}$, and the set of productions is

$$
\left\{q \rightarrow \mathrm{~A} p_{1} \ldots p_{m} \mid \mathrm{A} \in \mathscr{C}_{0}, \tau_{0}(\mathrm{~A})=p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow q\right\}
$$

Clearly $\#(\mathcal{A}(\mathscr{G}))=\#\left(L\left(G^{\Sigma_{0}}\right)\right)$, where $L\left(G^{\Sigma_{0}}\right)$ denotes the language generated by $G^{\Sigma_{0}}$. Since every context-free language is semilinear, so is the abstract language of every second-order ACG.

Since the lexicon $\mathscr{L}$ induces a linear translation which maps $\#(M)$ to $\#(\mathscr{L}(M))$ for $M \in \Lambda\left(\Sigma_{0}\right)$ and semilinearity is preserved under linear translations, $\#(\mathcal{O}(\mathscr{G}))$ is also semilinear.

Additionally, recent study by Kanazawa [26] shows that string ACLs form a substitution-closed full abstract family of languages in the sense of Ginsburg and Greibach [11]. Let $\mathbf{L}$ be a class of string languages. We say that $\mathbf{L}$ is a full AFL (full abstract family of languages) if $\mathbf{L}$ is closed under union, concatenation, Kleene star, homomorphism, inverse homomorphism, and intersection with regular sets. $\mathbf{L}$ is closed under substitution if, for every $L \in \mathbf{L}$ over an alphabet $V_{1}$ and every substitution $f$ from $V_{1}$ to $\mathcal{P}\left(V_{2}^{*}\right)$ such that $f(\mathrm{a}) \in \mathbf{L}$ for all $\mathrm{a} \in V_{1}$, we have $\hat{f}(L) \in \mathbf{L}$, where $\mathcal{P}\left(V_{2}^{*}\right)$ denotes the power set of $V_{2}^{*}$ and $\hat{f}$ is the extension of $f$ such that $\hat{f}(\varepsilon)=\{\varepsilon\}, \hat{f}(\mathrm{wa})=\hat{f}(\mathrm{w}) \hat{f}(\mathrm{a})$ for $\mathrm{w} \in V_{1}^{*}$ and $\mathrm{a} \in V_{1}, \hat{f}(L)=\bigcup_{\mathrm{w} \in L} \hat{f}(\mathrm{w})$.

Theorem 2.21 (Kanazawa [26]). The class of string languages generated by ACGs in $\mathbf{G}_{( }(m, n)$ is a substitution-closed full AFL for all $m, n \geq 2$.

## Chapter 3

## Non-Linear Extensions of Abstract Categorial Grammars

The theorems presented in Section 3.4 has been published in [60].

### 3.1 Introduction

The ACG formalism is based on simply typed linear lambda calculus in two senses,
(i) lexical entries of the grammar are all linear $\lambda$-terms,
(ii) grammatical combinations of them, i.e., elements of the abstract language, are also represented by linear $\lambda$-terms.

This accordance is, however, not mandatory. While the linearity constraint on the abstract language is thought to be reasonable, admitting non-linear $\lambda$-terms as lexical entries may allow ACGs to describe linguistic phenomena in a more natural and concise fashion.

The ACG $\mathscr{G}_{2}$ in Example 2.5 can be improved by allowing non-linear terms. Let $\mathscr{G}_{2}^{\prime}$ have the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}_{2}^{\prime}(\mathrm{x})$ |
| :---: | :---: | :---: |
| M | $n$ | $\lambda x \cdot \operatorname{man} x$ |
| W | $n$ | $\lambda x \cdot \mathbf{\text { woman } x}$ |
| J | $n p$ | John |
| R | $n p \rightarrow s$ | $\lambda x \cdot \mathbf{r u n} x$ |
| L | $n p^{2} \rightarrow s$ | $\lambda x_{1} x_{2} \cdot \operatorname{love} x_{2} x_{1}$ |
| A | $n \rightarrow(n p \rightarrow s) \rightarrow s$ | $\lambda y_{1} y_{2} \cdot \exists\left(\lambda x \cdot \wedge\left(y_{1} x\right)\left(y_{2} x\right)\right)$ |
| E | $n \rightarrow(n p \rightarrow s) \rightarrow s$ | $\lambda y_{1} y_{2} \cdot \forall\left(\lambda x . \rightarrow\left(y_{1} x\right)\left(y_{2} x\right)\right)$ |

where $\mathscr{L}_{2}^{\prime}(s)=t, \mathscr{L}_{2}^{\prime}(n p)=e, \mathscr{L}_{2}^{\prime}(n)=e \rightarrow t$, and $\exists, \forall, \wedge, \rightarrow$ are all object constants with $\tau_{2}^{\prime}(\exists)=\tau_{2}^{\prime}(\forall)=(e \rightarrow t) \rightarrow t$ and $\tau_{2}^{\prime}(\wedge)=\tau_{2}^{\prime}(\rightarrow)=t^{2} \rightarrow t$. The meanings of quantifiers $A$ and $E$ are represented by non-linear $\lambda$-terms which have multiple occurrences of bound variables,

Moreover, even vacuous $\lambda$-abstraction is useful for more flexible representations. Let the pair of two $\lambda$-terms $M_{1}$ and $M_{2}$ be represented by Pair $M_{1} M_{2}=\lambda v . v M_{1} M_{2}$ (Pairequiv $\lambda u_{1} u_{2} v . v u_{1} u_{2}$ ). One can get the $i$-th projection of a given pair of $\lambda$-terms by the non-linear $\lambda$-term $\pi_{i} \equiv \lambda w \cdot w\left(\lambda u_{1} u_{2} \cdot u_{i}\right)$ as

$$
\pi_{i}\left(\lambda v . v M_{1} M_{2}\right) \rightarrow_{\beta}\left(\lambda u_{1} u_{2} \cdot u_{i}\right) M_{1} M_{2} \rightarrow_{\beta} M_{i} .
$$

For $\mathscr{G}_{1}$ consisting of the following lexical entries,

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}_{1}(\mathrm{x})$ |
| :---: | :---: | :---: |
| M | $n$ | $/ \mathrm{man} /$ |
| W | $n$ | $/$ woman $/$ |
| J | $n p$ | $/$ John $/$ |
| R | $n p \rightarrow s$ | $\lambda x \cdot x+/$ runs $/$ |
| L | $n p^{2} \rightarrow s$ | $\lambda x_{1} x_{2} \cdot x_{2}+/$ loves $/+x_{1}$ |
| A | $n \rightarrow(n p \rightarrow s) \rightarrow s$ | $\lambda z w \cdot w(/ \mathrm{a} /+z)$ |
| E | $n \rightarrow(n p \rightarrow s) \rightarrow s$ | $\lambda z w \cdot w(/$ every $/+z)$ |

suppose that the word "every" is replaced with "all". Though the meanings of these two words are similar,

## * all man runs

is not a correct sentence. The correct sentence is
all men run.
A modification $\mathscr{G}_{1}^{\prime}$ is given as follows:

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}_{1}^{\prime}(\mathrm{x})$ |
| :---: | :---: | :---: |
| M | $n$ | $P / \operatorname{man} / /$ men $/$ |
| W | $n$ | $P /$ woman $/ /$ women $/$ |
| J | $n p$ | $\lambda y \cdot y / \mathrm{John} / \pi_{1}$ |
| R | $n p \rightarrow s$ | $\lambda x \cdot x(\lambda u v \cdot u+v(P /$ runs $/ /$ run $/))$ |
| L | $n p^{2} \rightarrow s$ | $\lambda x_{1} x_{2} \cdot x_{2}(\lambda u v . u+v(P /$ loves $/ /$ love $/))+x_{1}(\lambda u v . u)$ |
| A | $n \rightarrow(n p \rightarrow s) \rightarrow s$ | $\lambda z w \cdot w\left(\lambda y \cdot y\left(/ \mathrm{a} /+\pi_{1} z\right) \pi_{1}\right)$ |
| E | $n \rightarrow(n p \rightarrow s) \rightarrow s$ | $\lambda z w \cdot w\left(\lambda y \cdot y\left(/ \mathrm{all} /+\pi_{2} z\right) \pi_{2}\right)$ |

where $P=\lambda u_{1}^{s t r} u_{2}^{s t r} v^{s t r^{2} \rightarrow s t r} \cdot v u_{1} u_{2}, \pi_{i} \equiv \lambda w^{\left(s t r^{2} \rightarrow s t r\right) \rightarrow s t r} \cdot w\left(\lambda u_{1}^{s t r} u_{2}^{s t r} \cdot u_{i}\right), \mathscr{L}_{1}^{\prime}(n)=$ $\left(s t r^{2} \rightarrow s t r\right) \rightarrow s t r, \mathscr{L}_{1}^{\prime}(n p)=\left(s t r \rightarrow\left(\left(s t r^{2} \rightarrow s t r\right) \rightarrow s t r\right) \rightarrow s t r\right) \rightarrow s t r$,
$\mathscr{L}_{1}^{\prime}(s)=$ str. Then $\mathcal{O}\left(\mathscr{G}_{1}^{\prime}\right)$ consists of terms representing some English sentences such as John runs, all men run, all women love a man, and so on. The agreement between words is handled without using feature structures.

Both modifications on $\mathscr{G}_{1}$ and $\mathscr{G}_{2}$ are independently made without changing the abstract vocabulary $\Sigma_{0}$. The new ACGs $\mathscr{G}_{1}^{\prime}=\left\langle\Sigma_{0}, \Sigma_{1}^{\prime}, \mathscr{L}_{1}^{\prime}, s\right\rangle$ and $\mathscr{G}_{2}^{\prime}=\left\langle\Sigma_{0}, \Sigma_{2}^{\prime}, \mathscr{L}_{2}^{\prime}, s\right\rangle$ can generate correct pairs of a grammatical sentence and its meaning through the common abstract language.

On the other hand, it is known that the expressive power of some grammar formalisms involving a linearity constraint (non-duplication and nondeletion) does not change when the constraint is relaxed to just non-duplication, allowing deleting operations. Seki et al. [49] have shown the equivalence between linear context-free rewriting systems (LCFRSs) and multiple contextfree grammars (MCFGs), which correspond to the relaxed version of LCFRSs that may have deleting operations. Fujiyoshi [10] has established the equivalence between linear (non-duplicating non-deleting) monadic CFTGs and non-duplicating monadic CFTGs. Fisher's result [8,9] is rather general. He has shown that the string IO-languages defined by general CFTGs coincide with the string IO-languages defined by non-deleting CFTGs.

This chapter extends the definition of ACGs and moreover introduces variants of the definition of the abstract languages. A brief discussion on the relation among the resultant various classes of ACLs defined by those extensions is given in Section 3.3. In Section 3.4 we then focus on the relation between usual linear ACGs and affine ACGs, whose lexical entries may contain vacuous $\lambda$-abstraction, along the line mentioned above. We present a procedure for constructing a linear ACG corresponding to a given affine ACG such that the language of the constructed ACG is exactly the set of the linear $\lambda$-terms generated by the original ACG. Therefore, we conclude that affine ACGs are not essentially more expressive than linear ACGs, since strings and trees are usually represented by linear $\lambda$-terms.

As de Groote and Pogodalla $[18,19]$ have constructed linear ACGs encoding linear CFTGs and LCFRSs, non-duplicating CFTGs and MCFGs are also encodable by affine ACGs in straightforward ways. For such affine ACGs, our linearization method constructs linear ACGs which have the form corresponding to linear CFTGs or LCFRSs. Thus, our result is a generalization of the results we have mentioned above with the exception of Fisher's, which covers CFTGs involving duplication.

### 3.2 Definitions

In this chapter, we deal with non-linear $\lambda$-terms, so the word " $\lambda$-term" does not mean "linear $\lambda$-term", but a general unrestricted $\lambda$-term belonging to $\Lambda^{\mathrm{K}}(\Sigma)$ defined in Chapter 2. To emphasize the absence of the restriction to linear $\lambda$-terms, we call them $\lambda \mathrm{K}$-terms. A $\lambda \mathrm{K}$-term is a $\lambda \mathrm{I}$-term, if for every subterm of $M$ of the form $\lambda x . N, x \in \operatorname{Fv}(N)$. An affine term is a $\lambda$ K-term such that no variable occurs free twice or more in any subterm of $M$. Thus, a linear term is a $\lambda$ K-term that is $\lambda I$ and affine at the same time. We write $\Lambda^{\mathrm{K}}(\Sigma), \Lambda^{\mathrm{I}}(\Sigma), \Lambda^{\text {aff }}(\Sigma), \Lambda^{\text {lin }}(\Sigma)$, respectively for the sets of simply typed $\lambda$ K-terms, $\lambda$ I-terms, affine terms, linear terms on $\Sigma$.

We extend the definition of a lexicon $\mathscr{L}$ from $\Sigma_{0}$ to $\Sigma_{1}$ so that $\mathscr{L}($ a) is not restricted to linear terms but any simply typed $\lambda$ K-term on $\Sigma_{1}$ provided that it is a compatible pair of a type substitution and a term substitution. We call the generalized lexicons $\lambda \mathrm{K}$-lexicons. Similarly we define a $\lambda \mathrm{I}$-lexicon, an affine lexicon and a linear lexicon. A $\lambda \mathrm{K}-A C G$, a $\lambda \mathrm{I}-A C G$, an affine $A C G$, a linear $A C G$ are defined in the same way as usual ACGs but the lexicons $\mathscr{L}$ are $\lambda \mathrm{K}, \lambda \mathrm{I}$, affine, linear, respectively. $\mathbf{G}^{\mathrm{K}}, \mathbf{G}^{\mathrm{I}}, \mathbf{G}^{\text {aff }}, \mathbf{G}^{\text {lin }}$ denotes the class of $\lambda \mathrm{K}, \lambda \mathrm{I}$, affine, linear ACGs respectively, so ACGs satisfying the original definition are called with the modifier "linear" in this chapter. For each $\mathbf{G}^{*} \in\left\{\mathbf{G}^{\mathrm{K}}, \mathbf{G}^{\mathrm{I}}, \mathbf{G}^{\text {aff }}, \mathbf{G}^{\text {lin }}\right\}$, the subclasses denoted by $\mathbf{G}^{*}(m, n) \subseteq \mathbf{G}^{*}$ are defined in the same way as in Chapter 2.

Moreover, as we have done for the definition of an ACG, we can give variations of the definitions of the abstract language and object language of an ACG. For a $\lambda \mathrm{K}$-ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$, let

$$
\begin{aligned}
& \mathcal{A}^{\mathrm{K}}(\mathscr{G})=\left\{M \in \Lambda^{\mathrm{K}}\left(\Sigma_{0}\right) \mid M \text { is a closed } \beta \eta \text {-normal term of type } s\right\} \\
& \mathcal{O}^{\mathrm{K}}(\mathscr{G})=\left\{|\mathscr{L}(M)|_{\beta \eta} \mid M \in \mathcal{A}^{\mathrm{K}}(\mathscr{G})\right\} \\
& \mathcal{A}^{\operatorname{lin}}(\mathscr{G})=\left\{M \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}\right) \mid M \text { is a closed } \beta \eta \text {-normal term of type } s\right\} \\
& \mathcal{O}^{\operatorname{lin}}(\mathscr{G})=\left\{|\mathscr{L}(M)|_{\beta \eta} \mid M \in \mathcal{A}^{\operatorname{lin}}(\mathscr{G})\right\} .
\end{aligned}
$$

Note that while $\mathcal{A}^{\operatorname{lin}}(\mathscr{G})$ is a subset of $\Lambda^{\operatorname{lin}}\left(\Sigma_{0}\right), \mathcal{O}^{\operatorname{lin}}(\mathscr{G})$ is not necessarily a subset of $\Lambda^{\text {lin }}\left(\Sigma_{1}\right)$ if $\mathscr{G} \notin \mathbf{G}^{\text {lin }}$. We define $\mathcal{A}^{\text {aff }}(\mathscr{G})$, $\mathcal{O}^{\text {aff }}(\mathscr{G}), \mathcal{A}^{\mathrm{I}}(\mathscr{G}), \mathcal{O}^{\mathrm{I}}(\mathscr{G})$ similarly. For a class $\mathbf{G}^{*}$ of ACGs, let

$$
\begin{aligned}
\mathbf{L}^{\mathrm{K}}\left(\mathbf{G}^{*}\right) & =\left\{\mathcal{O}^{\mathrm{K}}(\mathscr{G}) \mid \mathscr{G} \in \mathbf{G}^{*}\right\} \\
\mathbf{L}^{\operatorname{lin}}\left(\mathbf{G}^{*}\right) & =\left\{\mathcal{O}^{\left.\operatorname{lin}(\mathscr{G}) \mid \mathscr{G} \in \mathbf{G}^{*}\right\}}\right.
\end{aligned}
$$

and $\mathbf{L}^{\text {aff }}\left(\mathbf{G}^{*}\right), \mathbf{L}^{\mathrm{I}}\left(\mathbf{G}^{*}\right)$ be defined similarly. $\mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {lin }}\right)$ is the class of original abstract categorial languages.

### 3.3 Relations among Variants of ACLs

By extending the definition of ACGs and their languages, we have obtained sixteen classes of languages. In this section, we discuss the relation among those classes. Let us focus on the four classes of abstract categorial languages $\mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {lin }}\right), \mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {aff }}\right), \mathbf{L}^{\text {aff }}\left(\mathbf{G}^{\text {lin }}\right)$, and $\mathbf{L}^{\text {aff }}\left(\mathbf{G}^{\text {aff }}\right)$. We can compare other classes of languages defined in the previous section similarly. First, by $\mathbf{G}^{\text {lin }} \subseteq \mathbf{G}^{\text {aff }}$, trivially $\mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {lin }}\right) \subseteq \mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {aff }}\right)$ and $\mathbf{L}^{\text {aff }}\left(\mathbf{G}^{\text {lin }}\right) \subseteq \mathbf{L}^{\text {aff }}\left(\mathbf{G}^{\text {aff }}\right)$ hold. Note that the fact that $\mathcal{O}^{\operatorname{lin}}(\mathscr{G}) \subseteq \mathcal{O}^{\text {aff }}(\mathscr{G})$ for all $\mathscr{G} \in \mathbf{G}$ does not imply that $\mathbf{L}^{\operatorname{lin}}\left(\mathbf{G}^{*}\right) \subseteq$ $\mathbf{L}^{\text {aff }}\left(\mathbf{G}^{*}\right)$ for $\mathbf{G}^{*} \in\left\{\mathbf{G}^{\text {lin }}, \mathbf{G}^{\text {aff }}\right\}$. In fact, the following proposition is rather easy.

Proposition 3.1. $\mathbf{L}^{\text {aff }}\left(\mathbf{G}^{\text {aff }}(m, n)\right) \subseteq \mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {aff }}(m, n)\right)$.
Proof. Given $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle \in \mathbf{G}^{\text {aff }}$, let us define $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L}^{\prime}, s\right\rangle \in$ $\mathrm{G}^{\text {aff }}$ as follows:
$\mathscr{A}_{0}^{\prime}=\mathscr{A}_{0}$,
$\mathscr{C}_{0}^{\prime}=\mathscr{C}_{0} \cup\left\{\mathrm{~K}_{\alpha}^{\beta} \mid \beta\right.$ is a negative subtype of $\tau_{0}(\mathrm{~b})$ for some $\mathrm{b} \in \mathscr{C}_{0}$, and
$\alpha$ is a negative subtype of a negative subtype of $\tau_{0}(\mathrm{a})$ for some $\left.\mathrm{a} \in \mathscr{C}_{0}\right\}$,
$\tau_{0}^{\prime}=\tau_{0} \cup\left\{\mathbf{K}_{\alpha}^{\beta} \mapsto \beta \rightarrow \alpha \rightarrow \beta\right\}$,
$\mathscr{L}^{\prime}=\mathscr{L} \cup\left\{\mathrm{K}_{\alpha}^{\beta} \mapsto \lambda x^{\mathscr{L}(\beta)} y^{\mathscr{L}(\alpha)} . x\right\}$.
It is enough to show that $\mathcal{O}^{\text {aff }}(\mathscr{G})=\mathcal{O}^{\text {lin }}\left(\mathscr{G}^{\prime}\right)$.
We first show that $\mathcal{O}^{\text {aff }}(\mathscr{G}) \subseteq \mathcal{O}^{\text {lin }}\left(\mathscr{G}^{\prime}\right)$. For $M \in \Lambda^{\text {aff }}\left(\Sigma_{0}\right)$, let $(M)^{*}$ be defined as follows:

$$
\begin{aligned}
(M)^{*} & =M \quad \text { if } M \in \mathscr{C}_{0} \cup \mathscr{X}, \\
\left(M_{1} M_{2}\right)^{*} & =\left(M_{1}\right)^{*}\left(M_{2}\right)^{*}, \\
\left(\lambda x^{\alpha} \cdot M\right)^{*} & = \begin{cases}\lambda x^{\alpha} \cdot(M)^{*} & \text { if } x \in \operatorname{Fv}(M), \\
\mathrm{K}_{\alpha}^{\beta}(M)^{*} & \text { if } x \notin \operatorname{Fv}(M) \text { and } \tau_{0}(M)=\beta .\end{cases}
\end{aligned}
$$

It is easy to see that for every $M \in \mathcal{A}^{\text {aff }}(\mathscr{G}),(M)^{*} \in \mathcal{A}^{\text {lin }}\left(\mathscr{G}^{\prime}\right)$ and $\mathscr{L}^{\prime}\left((M)^{*}\right)=$ $\mathscr{L}(M)$.
 ing each occurrence of a constant $\mathrm{K}_{\alpha}^{\beta}$ with $\lambda x^{\beta} y^{\alpha}$.x. Clearly $M \in \mathcal{A}^{\operatorname{aff}}(\mathscr{G})$ and $\mathscr{L}^{\prime}(N)=\mathscr{L}(M)$.

Thus, $\mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {aff }}\right)$ is the largest class among the four classes. The above proof transfers deleting operations in the abstract level into the object level via constants K that are mapped to $\lambda x y . x$. Therefore, for $\mathbf{G}^{*} \in\left\{\mathbf{G}^{\text {aff }}, \mathbf{G}^{\mathrm{K}}\right\}$,


Figure 3.1: Relation among variants of ACLs
we get $\mathbf{L}^{\mathrm{K}}\left(\mathbf{G}^{*}(m, n)\right) \subseteq \mathbf{L}^{\mathrm{I}}\left(\mathbf{G}^{*}(m, n)\right)$ and $\mathbf{L}^{\text {aff }}\left(\mathbf{G}^{*}(m, n)\right) \subseteq \mathbf{L}^{\text {lin }}\left(\mathbf{G}^{*}(m, n)\right)$. Similarly, duplicating operations in the abstract level can be relegated to the object level via constants W that are mapped to $\lambda x y . x y y$. For $\mathbf{G}^{*} \in$ $\left\{\mathbf{G}^{\mathrm{I}}, \mathbf{G}^{\mathrm{K}}\right\}$, we get $\mathbf{L}^{\mathrm{K}}\left(\mathbf{G}^{*}(m, n)\right) \subseteq \mathbf{L}^{\text {aff }}\left(\mathbf{G}^{*}(m, n)\right)$ and $\mathbf{L}^{\mathrm{I}}\left(\mathbf{G}^{*}(m, n)\right) \subseteq$ $\mathbf{L}^{\text {lin }}\left(\mathbf{G}^{*}(m, n)\right)$. Figure 3.1 summarizes the relations among variants of ACLs.

Corollary 3.2. Let $*$ be among $\{\mathrm{K}, \mathrm{I}$, aff $\}$. The class $\mathbf{L}^{\operatorname{lin}}\left(\mathbf{G}^{*}(m, n)\right)$ is the largest among $\mathbf{L}^{\operatorname{lin}}\left(\mathbf{G}^{*}(m, n)\right), \mathbf{L}^{*}\left(\mathbf{G}^{*}(m, n)\right), \mathbf{L}^{*}\left(\mathbf{G}^{\operatorname{lin}}(m, n)\right), \mathbf{L}^{\operatorname{lin}}\left(\mathbf{G}^{\operatorname{lin}}(m, n)\right)$.

Therefore, relaxing the linearity constraint on the abstract language does not enlarge the class of ACLs as much as extending the definition of ACGs themselves. In the sequel, hence we focus only on ACLs constructed on linear abstract languages.

### 3.4 Linearization of Affine ACGs

In the remainder of this chapter, we investigate the relation between $\mathbf{L}^{\operatorname{lin}}\left(\mathbf{G}^{\operatorname{lin}}\right)$ and $\mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {aff }}\right)$ in more detail. While $\mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {lin }}\right)$ consists of languages whose elements are all linear, $\mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {aff }}\right)$ contain languages including non-linear terms. Therefore, $\mathbf{L}^{\text {lin }}\left(\mathbf{G}^{\text {aff }}\right)$ is properly larger than $\mathbf{L}^{\operatorname{lin}}\left(\mathbf{G}^{\text {lin }}\right)$. However, since $\lambda$ terms representing strings or trees are all linear, non-linear terms in the object languages are not very interesting. The main result of this chapter is that for every $\mathscr{G} \in \mathbf{G}^{\text {aff }}(m, n)$, we can construct $\mathscr{G}^{l} \in \mathbf{G}^{\operatorname{lin}}(m, \max \{2, n\})$ such that

$$
\begin{equation*}
\mathcal{O}^{\operatorname{lin}}\left(\mathscr{G}^{l}\right)=\left\{P \in \mathcal{O}^{\operatorname{lin}}(\mathscr{G}) \mid P \text { is linear }\right\} . \tag{3.1}
\end{equation*}
$$

Moreover, in case of $m=2$, we can find $\mathscr{G}^{l} \in \mathbf{G}^{\operatorname{lin}}(2, n)$ satisfying the equation (3.1). Therefore extending the definition of an ACG to allow lexical entries to be affine does not enrich the expressive power of ACGs in an essential way.

Before proceeding with our construction, we mention a partially stronger result on the special case of this problem on second-order string ACGs, obtained from Salvati's work [45]. He presents an algorithm that converts a linear second-order string ACG $\mathscr{G} \in \mathbf{G}_{\text {string }}^{\operatorname{lin}}(2, n)$ into an equivalent LCFRS (via a deterministic tree-walking transducer). Even if an input is an affine ACG $\mathscr{G} \in \mathrm{G}_{\text {string }}^{\text {aff }}(2, n)$, his algorithm still outputs an equivalent LCFRS. Since every LCFRS is encodable by a linear ACG belonging to $\mathbf{G}_{\text {string }}^{\operatorname{lin}}(2,4)[18,19]$, this entails the following corollary.
Corollary 3.3. For every affine $A C G \mathscr{G} \in \mathbf{G}_{\text {string }}^{\mathrm{aff}}(2, n)$, there is a linear $A C G \mathscr{G}^{\prime} \in \mathbf{G}_{\text {string }}^{\operatorname{lin}}(2,4)$ such that $\mathcal{O}^{\operatorname{lin}}\left(\mathscr{G}^{\prime}\right)=\mathcal{O}^{\operatorname{lin}}(\mathscr{G})$.

### 3.4.1 Basic Idea

We explain our basic idea for linearization method for affine ACGs through a small example. Let us consider the affine ACG $\mathscr{G}$ consisting of the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}(\mathrm{x})$ |
| :---: | :---: | :---: |
| A | $p \rightarrow s$ | $\lambda w^{o^{2} \rightarrow o} . \mathrm{a}^{\circ} \mathrm{b}^{o}$ |
| B | $p$ | $\lambda x^{o} y^{o} . x$ |

where $\mathscr{L}(s)=o$ and $\mathscr{L}(p)=o^{2} \rightarrow o$. Corresponding to $\mathrm{AB} \in \mathcal{A}^{\operatorname{lin}}(\mathscr{G})$, we have $\mathrm{a} \in \mathcal{O}^{\operatorname{lin}}(\mathscr{G})$ by

$$
\begin{equation*}
\mathscr{L}(\mathrm{AB}) \equiv\left(\lambda w^{o \rightarrow o \rightarrow o} . w \mathbf{a}^{o} \mathrm{~b}^{o}\right)\left(\lambda x^{o} y^{o} . x\right) \rightarrow_{\beta}\left(\lambda x^{o} y^{o} . x\right) \mathrm{a}^{o} \mathrm{~b}^{o} \rightarrow_{\beta} \mathrm{a}^{o} . \tag{3.2}
\end{equation*}
$$

The occurrences of vacuous $\lambda$-abstraction $\lambda y^{o}$ causes the deletion of b in (3.2). Such deleting operation is what we want to eliminate in order to linearize the affine ACG $\mathscr{G}$. Let us retype $\lambda y^{o}$ with $\lambda y^{\bar{\sigma}}$ and replace $\mathrm{b}^{o}$ with $\overline{\mathrm{b}}^{\bar{o}}$ to indicate that they should be eliminated. Then (3.2) is decorated by bars as

$$
\begin{equation*}
\left(\lambda w^{o \rightarrow \bar{o} \rightarrow o} \cdot w \mathrm{a}^{o} \overline{\mathrm{~b}}^{\bar{o}}\right)\left(\lambda x^{o} y^{\bar{o}} \cdot x\right) \rightarrow_{\beta}\left(\lambda x^{o} y^{\bar{o}} \cdot x\right) \mathrm{a}^{o} \overline{\mathrm{~b}}^{\bar{o}} \rightarrow_{\beta} \mathrm{a}^{o}, \tag{3.3}
\end{equation*}
$$

where we retype $w^{o \rightarrow o \rightarrow o}$ with $w^{o \rightarrow \bar{o} \rightarrow o}$, so that the whole term is well-typed. In our setting, when a term has a barred type, it means that the term should be erased during $\beta$-reduction steps, and vice versa. By eliminating those barred terms and types from (3.3), we get

$$
\begin{equation*}
\left(\lambda w^{o \rightarrow o} . w \mathbf{a}^{o}\right)\left(\lambda x^{o} \cdot x\right) \rightarrow_{\beta}\left(\lambda x^{o} \cdot x\right) \mathbf{a}^{o} \rightarrow_{\beta} \mathrm{a}^{o}, \tag{3.4}
\end{equation*}
$$

which solely consists of linear terms. Hence, the linear ACG $\mathscr{G}^{\prime}$ with the following lexical entries generates the same language as the original ACG $\mathscr{G}$.

| $\mathrm{x} \in \mathscr{C}_{0}^{\prime}$ | $\tau_{0}^{\prime}(\mathrm{x})$ | $\mathscr{L}^{\prime}(\mathrm{x})$ |
| :---: | :---: | :---: |
| $\mathrm{A}^{\prime}$ | $[p, o \rightarrow \bar{o} \rightarrow o] \rightarrow[s, o]$ | $\lambda w^{o \rightarrow o} \cdot w \mathrm{a}^{o}$ |
| $\mathrm{~B}^{\prime}$ | $[p, o \rightarrow \bar{o} \rightarrow o]$ | $\lambda x^{o} \cdot x$ |

where $[p, o \rightarrow \bar{o} \rightarrow o$ ] and $[s, o]$ are new atomic types that are mapped to $o \rightarrow o$ and $o$, respectively, and $[s, o]$ is the distinguished type. We have
 just one possible bar-decoration for $\mathscr{L}(\mathrm{A})$. For instance, $\lambda w^{\bar{o} \rightarrow o \rightarrow o} . w \overline{\mathrm{a}}^{\bar{\sigma}} \mathrm{b}^{o}$ and $\lambda w^{o \rightarrow o \rightarrow o} . w \mathrm{a}^{\circ} \mathrm{b}^{o}$ are also possible. Bars appearing in $\lambda w^{\bar{\alpha} \rightarrow o \rightarrow o} . w \overline{\mathrm{a}}^{\bar{\sigma}} \mathrm{b}^{o}$ predict that the subterm $\overline{\mathrm{a}}$ will be erased, and $\lambda w^{o \rightarrow o \rightarrow o} . w \mathrm{a}^{\circ} \mathrm{b}^{o}$ predicts that no subterm of it will disappear. Our linearization method also produces lexical entries corresponding to those bar-decorations.

### 3.4.2 Formal Definition

We first give a formal definition of the set of possible bar-decorations on a type and a term. Hereafter, we fix a given affine ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$. Define $\overline{\Sigma_{1}}=\left\langle\overline{\mathscr{A}_{1}}, \overline{\mathscr{C}_{1}}, \overline{\tau_{1}}\right\rangle$ by

$$
\overline{\mathscr{A}_{1}}=\left\{\bar{p} \mid p \in \mathscr{A}_{1}\right\}, \quad \overline{\mathscr{C}_{1}}=\left\{\overline{\mathrm{c}} \mid \mathrm{c} \in \mathscr{C}_{1}\right\}, \quad \overline{\tau_{1}}=\left\{\overline{\mathrm{c}} \rightarrow \overline{\tau_{1}(\mathrm{c})} \mid \mathrm{c} \in \mathscr{C}_{1}\right\},
$$

where $\overline{\alpha \rightarrow \beta}=\bar{\alpha} \rightarrow \bar{\beta}$. Let $\Sigma_{1}^{\prime}=\left\langle\mathscr{A}_{1}^{\prime}, \mathscr{C}_{1}^{\prime}, \tau_{1}^{\prime}\right\rangle=\left\langle\mathscr{A}_{1} \cup \overline{\mathscr{A}_{1}}, \mathscr{C}_{1} \cup \overline{\mathscr{C}_{1}}, \tau_{1} \cup \overline{\tau_{1}}\right\rangle$. Here, we have the relabeling lexicon $\widetilde{\sim}$ from $\Sigma_{1}^{\prime}$ to $\Sigma_{1}$ defined as

$$
\begin{cases}\widetilde{\bar{p}}=\widetilde{p}=p & \text { for } p \in \mathscr{A}_{1}, \\ \widetilde{\mathrm{c}} \equiv \widetilde{\mathrm{c}} \equiv \mathrm{c} & \text { for } \mathrm{c} \in \mathscr{C}_{1} .\end{cases}
$$

For $P \in \Lambda^{\text {aff }}\left(\Sigma_{1}\right)$, we let $\bar{P}$ denote the unique term such that $\bar{P} \in \Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right)$ and $\widetilde{\bar{P}} \equiv P$.

The set $\widehat{\mathcal{T}}\left(\mathscr{A}_{1}\right)$ of possible bar-decorations on types is defined by

$$
\begin{aligned}
\widehat{\mathcal{T}}\left(\mathscr{A}_{1}\right)=\left\{\alpha \in \mathcal{T}\left(\mathscr{A}_{1}^{\prime}\right) \mid \text { if } \beta_{1}\right. & \rightarrow \cdots \rightarrow \beta_{n} \rightarrow \bar{p} \text { is a subtype of } \alpha \\
& \left.\quad \text { for some } \bar{p} \in \overline{\mathscr{A}_{1}}, \text { then } \beta_{1}, \ldots, \beta_{n} \in \mathcal{T}\left(\overline{\Sigma_{1}}\right)\right\}
\end{aligned}
$$

Actually, terms in $\Lambda^{\text {aff }}\left(\Sigma_{1}^{\prime}\right)$ that we are concerned with have types in $\widehat{\mathcal{T}}\left(\mathscr{A}_{1}\right)$. The reason why we ignore types in $\mathcal{T}\left(\mathscr{A}_{1}^{\prime}\right)-\widehat{\mathcal{T}}\left(\mathscr{A}_{1}\right)$ is that if a term is bound to be erased, then so is every subterm of it. For instance, if a variable $x$ has type $o \rightarrow \bar{o} \notin \widehat{\mathcal{T}}(\{o\})$, then the term $x^{o \rightarrow \bar{o}} y^{o}$ has type $\bar{o}$, which, in our setting, means that it should disappear. But if $x^{o \rightarrow \bar{o}} y^{o}$ disappears, so does $y^{o}$, which, therefore, should have type $\bar{o}$ to be consistent with our definition.

The set $\widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$ of possible bar-decorations on terms is the subset of $\Lambda^{\text {aff }}\left(\Sigma_{1}^{\prime}\right)$ such that $Q \in \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$ iff

- every variable appearing in $Q$ has a type in $\widehat{\mathcal{T}}\left(\mathscr{A}_{1}\right)$, and
- if $\lambda x^{\alpha} . Q^{\prime}$ is a subterm of $Q$ and $x^{\alpha} \notin \operatorname{Fv}\left(Q^{\prime}\right)$, then $\alpha \in \mathcal{T}\left(\overline{\mathscr{A}_{1}}\right)$.

We are not concerned with terms in $\Lambda^{\text {aff }}\left(\Sigma_{1}^{\prime}\right)-\widehat{\Lambda}{ }^{\text {aff }}\left(\Sigma_{1}\right)$.
The following properties are easily seen:

- If $Q \in \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$, then $\tau_{1}^{\prime}(Q) \in \widehat{\mathcal{T}}\left(\mathscr{A}_{1}\right)$,
- If $Q \in \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$ has a type in $\mathcal{T}\left(\overline{\mathscr{A}_{1}}\right)$, then every subterm of $Q$ is in $\Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right)$,
- If $Q \in \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$ and $Q \rightarrow_{\beta} Q^{\prime}$, then $Q^{\prime} \in \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$.

For each $\alpha \in \mathcal{T}\left(\mathscr{A}_{1}\right)$ and $P \in \Lambda^{\text {aff }}\left(\Sigma_{1}\right)$, $\Pi$ gives the set of possible bardecorations on them:

$$
\begin{aligned}
\Pi(\alpha) & =\left\{\beta \in \widehat{\mathcal{T}}\left(\mathscr{A}_{1}\right) \mid \widetilde{\beta}=\alpha\right\} \\
\Pi(P) & =\left\{Q \in \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right) \mid \widetilde{Q} \equiv P\right\} .
\end{aligned}
$$

In other words, $\Pi$ and $\simeq$ are inverse of each other, if we disregard types in $\mathcal{T}\left(\mathscr{A}_{1}^{\prime}\right)-\widehat{\mathcal{T}}\left(\mathscr{A}_{1}\right)$ and terms in $\Lambda^{\text {aff }}\left(\Sigma_{1}^{\prime}\right)-\widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$.

Secondly, we eliminate barred subtypes from $\alpha \in \widehat{\mathcal{T}}\left(\mathscr{A}_{1}\right)-\mathcal{T}\left(\overline{\mathscr{A}_{1}}\right)$ and barred subterms from $Q \in \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)-\Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right)$. Let us define $(\alpha)^{\dagger}$ and $(Q)^{\dagger}$ as follows:

$$
\begin{aligned}
(p)^{\dagger} & =p \quad \text { for } p \in \mathscr{A}_{1}, \\
(\alpha \rightarrow \beta)^{\dagger} & = \begin{cases}(\alpha)^{\dagger} \rightarrow(\beta)^{\dagger} & \text { if } \alpha \notin \mathcal{T}\left(\overline{\mathscr{A}_{1}}\right), \\
(\beta)^{\dagger} & \text { if } \alpha \in \mathcal{T}\left(\overline{\mathscr{A}_{1}}\right),\end{cases} \\
\left(x^{\alpha}\right)^{\dagger} & \equiv x^{(\alpha)^{\dagger}}, \\
(\mathrm{c})^{\dagger} & \equiv \mathrm{c} \quad \text { for } \mathrm{c} \in \mathscr{C}_{1}, \\
\left(\lambda x^{\alpha} \cdot Q\right)^{\dagger} & \equiv \begin{cases}\lambda x^{(\alpha)^{\dagger}} .(Q)^{\dagger} & \text { if } \alpha \notin \mathcal{T}\left(\overline{\mathscr{A}_{1}}\right), \\
(Q)^{\dagger} & \text { if } \alpha \in \mathcal{T}\left(\overline{\mathscr{A}_{1}}\right),\end{cases} \\
\left(Q_{1} Q_{2}\right)^{\dagger} & \equiv \begin{cases}\left(Q_{1}\right)^{\dagger}\left(Q_{2}\right)^{\dagger} & \text { if } \tau_{1}^{\prime}\left(Q_{2}\right) \notin \mathcal{T}\left(\overline{\mathscr{A}_{1}}\right), \\
\left(Q_{1}\right)^{\dagger} & \text { if } \tau_{1}^{\prime}\left(Q_{2}\right) \in \mathcal{T}\left(\overline{\mathscr{A}_{1}}\right) .\end{cases}
\end{aligned}
$$

The following properties are easily seen $\left(\alpha \in \widehat{\mathcal{T}}\left(\mathscr{A}_{1}\right)-\mathcal{T}\left(\overline{\mathscr{A}_{1}}\right)\right.$ and $Q, Q^{\prime} \in$ $\left.\widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)-\Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right)\right):$

- $(\alpha)^{\dagger} \in \mathcal{T}\left(\mathscr{A}_{1}\right)$ and $(Q)^{\dagger} \in \Lambda^{\operatorname{lin}}\left(\Sigma_{1}\right)$,
- $\tau_{1}\left((Q)^{\dagger}\right)=\left(\tau_{1}^{\prime}(Q)\right)^{\dagger}$,
- If $Q$ is $\beta$-normal, then so is $(Q)^{\dagger}$,
- $Q={ }_{\beta} Q^{\prime}$ implies $(Q)^{\dagger}={ }_{\beta}\left(Q^{\prime}\right)^{\dagger}$.

Lemma 3.4. Let $Q \in \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$ be given. If $Q$ and its free variables have types in $\mathcal{T}\left(\mathscr{A}_{1}\right)$, then $(Q)^{\dagger}={ }_{\beta} Q={ }_{\beta} \widetilde{Q}$.

Proof. Since both functions $\widetilde{\sim}$ and $\dagger$ preserve the $\beta$-equality, we can assume that $Q$ is $\beta$-normal. We show that $(Q)^{\dagger} \equiv Q \equiv \widetilde{Q}$ by induction on $Q$.

If $Q \equiv x^{\alpha} Q_{1} \ldots Q_{m}$, by $\alpha \in \mathcal{T}\left(\mathscr{A}_{1}\right)$, we have $\tau_{1}^{\prime}\left(Q_{i}\right) \in \mathcal{T}\left(\mathscr{A}_{1}\right)$ for each $i \in\{1, \ldots, m\}$. Applying the induction hypothesis to each $Q_{i}$, we get the conclusion $(Q)^{\dagger} \equiv Q \equiv \widetilde{Q}$.

If $Q \equiv \mathrm{c} Q_{1} \ldots Q_{m}$ for $\mathrm{c} \in \mathscr{C}_{1}^{\prime}$, by $\tau_{1}^{\prime}(Q) \in \mathcal{T}\left(\mathscr{A}_{1}\right)$, we have $\mathrm{c} \in \mathscr{C}_{1}$. By the type of c , we can apply the induction hypothesis to each $Q_{i}$, and get the conclusion $(Q)^{\dagger} \equiv Q \equiv \widehat{Q}$.

If $Q \equiv \lambda x^{\alpha} \cdot Q^{\prime}$, by $\tau_{1}^{\prime}(Q)=\alpha \rightarrow \tau_{1}^{\prime}\left(Q^{\prime}\right) \in \mathcal{T}\left(\mathscr{A}_{1}\right)$, we can apply the induction hypothesis to $Q^{\prime}$. Thus $(Q)^{\dagger} \equiv Q \equiv \widetilde{Q}$.

Lemma 3.5. For every closed term $P \in \Lambda^{\text {aff }}\left(\Sigma_{1}\right),|P|_{\beta}$ is linear iff there is $Q \in \Pi(P)$ whose type is in $\mathcal{T}\left(\mathscr{A}_{1}\right)$.

Proof. First we show the "if" direction. If there is $Q \in \Pi(P)$ with $\tau_{1}^{\prime}(Q) \in$ $\mathcal{T}\left(\mathscr{A}_{1}\right)$, then $(Q)^{\dagger}={ }_{\beta} \widetilde{Q} \equiv P$ by Lemma 3.4. Since $(Q)^{\dagger}$ is linear, so is $|P|_{\beta}$.

Second we show the "only if" direction by induction on the maximum number of $\beta$-reduction steps from $P$ to $|P|_{\beta}$. If $P$ is $\beta$-normal and linear, then $P \in \Pi(P)$. Otherwise, suppose that $P$ has a $\beta$-redex as $P \equiv P_{0}[z:=$ $\left(\lambda x^{\alpha} . P_{1}\right) P_{2}$ ], where $z$ occurs free in $P_{0}$.

Case 1. $x^{\alpha} \in \operatorname{Fv}\left(P_{1}\right)$ and

$$
P \equiv P_{0}\left[z:=\left(\lambda x^{\alpha} \cdot P_{1}\right) P_{2}\right] \rightarrow_{\beta} P_{0}\left[z:=P_{1}\left[P_{2} / x^{\alpha}\right]\right] \rightarrow_{\beta}|P|_{\beta} .
$$

By the induction hypothesis, there is $Q^{\prime} \in \Pi\left(P_{0}\left[z:=P_{1}\left[P_{2} / x^{\alpha}\right]\right]\right)$ such that $\tau_{1}^{\prime}\left(Q^{\prime}\right) \in \mathcal{T}\left(\mathscr{A}_{1}\right) . Q^{\prime}$ has the form $Q^{\prime} \equiv Q_{0}\left[z:=Q_{1}\left[Q_{2} / x^{\beta}\right]\right]$ for $Q_{i} \in \Pi\left(P_{i}\right)$ for $i=0,1,2$, and $\beta \in \Pi(\alpha)$. Let $Q \equiv Q_{0}\left[z:=\left(\lambda x^{\beta} . Q_{1}\right) Q_{2}\right]$. Clearly $Q \in \Pi(P)$ and $\tau_{1}^{\prime}(Q)=\tau_{1}^{\prime}\left(Q^{\prime}\right) \in \mathcal{T}\left(\mathscr{A}_{1}\right)$.

Case 2. $x^{\alpha} \notin \operatorname{Fv}\left(P_{1}\right)$ and

$$
P \equiv P_{0}\left[z:=\left(\lambda x^{\alpha} . P_{1}\right) P_{2}\right] \rightarrow_{\beta} P_{0}\left[z:=P_{1}\right] \rightarrow_{\beta}|P|_{\beta} .
$$

By the induction hypothesis, there is $Q^{\prime} \in \Pi\left(P_{0}\left[z:=P_{1}\right]\right)$ such that $\tau_{1}^{\prime}\left(Q^{\prime}\right) \in$ $\mathcal{T}\left(\mathscr{A}_{1}\right)$. $Q^{\prime}$ has the form $Q^{\prime} \equiv Q_{0}\left[z:=Q_{1}\right]$ for $Q_{i} \in \Pi\left(P_{i}\right)$ for $i=0,1$. Suppose that $y^{\gamma} \in \operatorname{Fv}\left(P_{2}\right)$. Since $P_{0}\left[z:=\left(\lambda x^{\alpha} . P_{1}\right) P_{2}\right]$ is closed, $P_{0}$ contains an occurrence of the $\lambda$-abstraction $\lambda y^{\gamma}$ that binds the free occurrence of $y^{\gamma}$
in $P_{2}$. The occurrence of the $\lambda$-abstraction $\lambda y^{\gamma}$ becomes vacuous in $P_{0}[z:=$ $\left.P_{1}\right]$. By $Q^{\prime} \in \Pi\left(P_{0}\left[z:=P_{1}\right]\right) \subseteq \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$, the corresponding occurrence of the $\lambda$-abstraction in $Q_{0}$ is $\lambda y^{\bar{\gamma}}$. Let $Q \equiv Q_{0}\left[z:=\left(\lambda x^{\bar{\alpha}} \cdot Q_{1}\right) \overline{P_{2}}\right]$. By the above observation, $Q$ is well-typed (the types of bound variables and their abstractions are consistent). Clearly $Q \in \Pi(P)$ and $\tau_{1}^{\prime}(Q)=\tau_{1}^{\prime}\left(Q^{\prime}\right) \in \mathcal{T}\left(\mathscr{A}_{1}\right)$.

## Second-Order Case

We say that an abstract atomic type $p \in \mathscr{A}_{0}$ is useless if there is no $M \in$ $\mathcal{A}^{\operatorname{lin}}(\mathscr{G})$ that has a subterm whose type contains $p$. An abstract constant a is useless if there is no $M \in \mathcal{A}^{\operatorname{lin}}(\mathscr{G})$ containing a. If an ACG is second-order, it is easy to check whether the abstract vocabulary contains useless atomic types or constants, and if so, we easily eliminate useless abstract atomic types and constants. This can be done as elimination of useless nonterminals and productions from a context-free grammar.

Definition 3.6. Let $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ be a second-order ACG that has no useless abstract atomic types or useless abstract constants. We define $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L}^{\prime},[s, \mathscr{L}(s)]\right\rangle$ as follows: define $\Sigma_{0}^{\prime}$ by

$$
\begin{aligned}
& \mathscr{A}_{0}^{\prime}=\left\{[p, \beta] \mid p \in \mathscr{A}_{0}, \beta \in \Pi(\mathscr{L}(p))-\mathcal{T}\left(\overline{\mathscr{A}_{1}}\right)\right\} \\
& \mathscr{C}_{0}^{\prime}=\left\{\llbracket \mathrm{a}, Q \rrbracket \mid \mathrm{a} \in \mathscr{C}_{0}, Q \in \Pi(\mathscr{L}(\mathrm{a}))-\Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right)\right\} \\
& \tau_{0}^{\prime}=\left\{\llbracket \mathrm{a}, Q \rrbracket \mapsto\left(\left[\tau_{0}(\mathrm{a}), \tau_{1}^{\prime}(Q)\right]\right)^{\ddagger}\right\} \text { where } \\
& \quad([p, \beta])^{\ddagger}=[p, \beta], \\
& \quad([\alpha \rightarrow \gamma, \beta \rightarrow \delta])^{\ddagger}= \begin{cases}([\alpha, \beta])^{\ddagger} \rightarrow([\gamma, \delta])^{\ddagger} & \text { if } \beta \notin \mathcal{T}\left(\overline{\mathscr{A}_{1}}\right), \\
([\gamma, \delta])^{\ddagger} & \text { if } \beta \in \mathcal{T}\left(\mathscr{A}_{1}\right),\end{cases}
\end{aligned}
$$

and $\mathscr{L}^{\prime}$ by

$$
\mathscr{L}^{\prime}([p, \beta])=(\beta)^{\dagger}, \quad \mathscr{L}^{\prime}(\llbracket \mathrm{a}, Q \rrbracket)=(Q)^{\dagger}
$$

$\mathscr{G}^{\prime}$ is linear, but it may contain useless abstract atomic types or constants. The linearized $A C G \mathscr{G}^{l}$ for $\mathscr{G}$ is the result of eliminating all the useless abstract atomic types and constants from $\mathscr{G}^{\prime}$.
Lemma 3.7. Let $\mathscr{G}$ and $\mathscr{G}^{\prime}$ be as in Definition 3.6.
For every variable-free $M \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}\right)$ of an atomic type and every $Q \in$ $\Pi(\mathscr{L}(M))-\Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right)$, there is $N \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}(N)=\left[\tau_{0}(M), \tau_{1}^{\prime}(Q)\right]$ and $\mathscr{L}^{\prime}(N) \equiv(Q)^{\dagger}$.

Conversely, for every variable-free $N \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}^{\prime}\right)$ of an atomic type, there are $M \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}\right)$ and $Q \in \Pi(\mathscr{L}(M))-\Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right)$ such that $\tau_{0}^{\prime}(N)=\left[\tau_{0}(M)\right.$, $\left.\tau_{1}^{\prime}(Q)\right]$ and $\mathscr{L}^{\prime}(N) \equiv(Q)^{\dagger}$.

Proof. Induction on $M$ and $N$ respectively.
Let $M=\mathrm{a} M_{1} \ldots M_{m} \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}\right) . Q \in \Pi(\mathscr{L}(M))$ is written as

$$
Q \equiv Q_{0} Q_{1} \ldots Q_{m}
$$

where $Q_{0} \in \Pi(\mathscr{L}(\mathrm{a}))$ and $Q_{i} \in \Pi\left(\mathscr{L}\left(M_{i}\right)\right)$ for $i \in\{1, \ldots, m\}$. Let $I=\{i \mid$ $\left.1 \leq i \leq m, \tau_{1}^{\prime}\left(Q_{i}\right) \notin \mathcal{T}\left(\overline{\mathscr{A}_{1}}\right)\right\}$. Then,

$$
(Q)^{\dagger} \equiv\left(Q_{0}\right)^{\dagger}\left(Q_{i_{1}}\right)^{\dagger} \ldots\left(Q_{i_{k}}\right)^{\dagger},
$$

where $\left\{i_{1}, \ldots, i_{k}\right\}=I$. By the definition of $\mathscr{G}^{\prime}$, there is $\llbracket \mathrm{a}, Q_{0} \rrbracket \in \mathscr{C}_{0}^{\prime}$ such that

$$
\begin{aligned}
\tau_{0}^{\prime}\left(\llbracket \mathrm{a}, Q_{0} \rrbracket\right) & =\left[\tau_{0}\left(M_{i_{1}}\right), \tau_{1}^{\prime}\left(Q_{i_{1}}\right)\right] \rightarrow \cdots \rightarrow\left[\tau_{0}\left(M_{i_{k}}\right), \tau_{1}^{\prime}\left(Q_{i_{k}}\right)\right] \rightarrow\left[\tau_{0}(M), \tau_{1}^{\prime}(Q)\right] \\
\mathscr{L}^{\prime}\left(\llbracket \mathrm{a}, Q_{0} \rrbracket\right) & \equiv\left(Q_{0}\right)^{\dagger}
\end{aligned}
$$

For each $i \in I$, the induction hypothesis gives $N_{i} \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}\left(N_{i}\right)=\left[\tau_{0}\left(M_{i}\right), \tau_{1}^{\prime}\left(Q_{i}\right)\right]$ and $\mathscr{L}^{\prime}\left(N_{i}\right) \equiv\left(Q_{i}\right)^{\dagger}$. Clearly for

$$
N \equiv \llbracket \mathrm{a}, Q_{0} \rrbracket N_{i_{1}} \ldots N_{i_{k}},
$$

we have $\tau_{0}^{\prime}(N)=\left[\tau_{0}(M), \tau_{1}^{\prime}(Q)\right]$ and $\mathscr{L}^{\prime}(N) \equiv(Q)^{\dagger}$.
Conversely, suppose that a variable free term $N \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}^{\prime}\right)$ of an atomic type is given. For the head $\llbracket \mathrm{a}, Q_{0} \rrbracket$ of $N$, let

$$
\begin{aligned}
\tau_{0}(\mathrm{a}) & =p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow p_{0} \\
\tau_{1}^{\prime}\left(Q_{0}\right) & =\alpha_{1} \rightarrow \cdots \rightarrow \alpha_{m} \rightarrow \alpha_{0}
\end{aligned}
$$

where $\alpha_{i} \in \Pi\left(\mathscr{L}\left(p_{i}\right)\right)$ for $i \in\{0,1, \ldots, m\}$. Let $I=\left\{i \mid 1 \leq i \leq m, \alpha_{i} \notin\right.$ $\left.\mathcal{T}\left(\overline{\mathscr{A}_{1}}\right)\right\}$. Then, we have

$$
\begin{aligned}
\tau_{0}^{\prime}\left(\llbracket \mathrm{a}, Q_{0} \rrbracket\right) & =\left[p_{i_{1}}, \alpha_{i_{1}}\right] \rightarrow \cdots \rightarrow\left[p_{i_{k}}, \alpha_{i_{k}}\right] \rightarrow\left[p_{0}, \alpha_{0}\right], \\
N & \equiv \llbracket \mathrm{a}, Q_{0} \rrbracket N_{i_{1}} \cdots N_{i_{k}},
\end{aligned}
$$

where $\left\{i_{1}, \ldots, i_{k}\right\}=I$. For each $i \in I$, the induction hypothesis gives $M_{i} \in$ $\Lambda^{\operatorname{lin}}\left(\Sigma_{0}\right)$ and $Q_{i} \in \Pi\left(\mathscr{L}\left(M_{i}\right)\right)-\Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right)$ such that $\tau_{0}^{\prime}\left(N_{i}\right)=\left[\tau_{0}\left(M_{i}\right), \tau_{1}^{\prime}\left(Q_{i}\right)\right]=$ [ $p_{i}, \alpha_{i}$ ] and $\mathscr{L}^{\prime}\left(N_{i}\right) \equiv\left(Q_{i}\right)^{\dagger}$. Recall that $\mathscr{G}$ has neither useless atomic types nor useless constants. Thus for each $j \in\{1, \ldots, m\}-I$, we can find a variablefree term $M_{j} \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}\right)$ of type $p_{j}$. Let $Q_{j} \equiv \overline{\mathscr{L}\left(M_{j}\right)}$ for $j \in\{1, \ldots, m\}-I$. Let

$$
\begin{aligned}
M & \equiv \mathrm{a} M_{1} \ldots M_{m} \\
Q & \equiv Q_{0} Q_{1} \ldots Q_{m}
\end{aligned}
$$

## Clearly

$$
\begin{aligned}
Q & \in \Pi(\mathscr{L}(M))-\Lambda^{\text {aff }}\left(\overline{\Sigma_{0}}\right), \\
\tau_{0}^{\prime}(N) & =\left[p_{0}, \alpha_{0}\right]=\left[\tau_{0}(M), \tau_{1}^{\prime}(Q)\right], \\
\mathscr{L}^{\prime}(N) & \equiv \mathscr{L}^{\prime}\left(\llbracket \mathrm{a}, Q_{0} \rrbracket N_{i_{1}} \ldots N_{i_{k}}\right) \\
& \equiv\left(Q_{0}\right)^{\dagger}\left(Q_{i_{1}}\right)^{\dagger} \ldots\left(Q_{i_{k}}\right)^{\dagger} \\
& \equiv(Q)^{\dagger} .
\end{aligned}
$$

Theorem 3.8. For every affine $A C G \mathscr{G} \in \mathbf{G}^{\text {aff }}(2, n)$, there is a linear $A C G$ $\mathscr{G}^{l} \in \mathbf{G}^{\operatorname{lin}}(2, n)$ such that $\mathcal{O}^{\operatorname{lin}}\left(\mathscr{G}^{l}\right)=\left\{P \in \mathcal{O}^{\operatorname{lin}}(\mathscr{G}) \mid P\right.$ is linear $\}$.

Proof. It is enough to show that $\mathcal{O}^{\operatorname{lin}}\left(\mathscr{G}^{\prime}\right)=\left\{P \in \mathcal{O}^{\operatorname{lin}}(\mathscr{G}) \mid P\right.$ is linear $\}$.
Let $M \in \mathcal{A}^{\operatorname{lin}}(\mathscr{G})$ be such that $|\mathscr{L}(M)|_{\beta \eta}$ is linear. By Lemma 3.5, there is $Q \in \Pi(\mathscr{L}(M))$ of type $\mathscr{L}(s)$. By applying Lemma 3.7 to $M$ and $Q$, we have $N \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}^{\prime}\right)$ with $\tau_{0}^{\prime}(N)=[s, \mathscr{L}(s)]$ and $\mathscr{L}^{\prime}(N) \equiv(Q)^{\dagger}$. We have $N \underset{\sim}{N} \in \mathcal{A}^{\operatorname{lin}}\left(\mathscr{G}^{\prime}\right)$. Together with Lemma 3.4, it holds that $\mathscr{L}^{\prime}(N) \equiv(Q)^{\dagger}={ }_{\beta}$ $\widetilde{Q} \equiv \mathscr{L}(M)$.

Let $N \in \mathcal{A}^{\operatorname{lin}}\left(\mathscr{G}^{\prime}\right)$. By applying Lemma 3.7 to $N$, we have $M \in \Lambda^{\operatorname{lin}\left(\Sigma_{0}\right)}$ and $Q \in \Pi(\mathscr{L}(M))-\Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right)$ such that $\tau_{0}(M)=s, \tau_{1}^{\prime}(Q)=\mathscr{L}(s)$, and $\mathscr{L}^{\prime}(N) \equiv(Q)^{\dagger}$. We have $M \in \mathcal{A}^{\operatorname{lin}}(\mathscr{G})$. Together with Lemma 3.4, it holds that $\mathscr{L}^{\prime}(N) \equiv(Q)^{\dagger}={ }_{\beta} \widetilde{Q} \equiv \mathscr{L}(M)$.

De Groote and Pogodalla $[18,19]$ have presented encoding methods for linear CFTGs and LCFRSs by linear ACGs. Their methods can also be applied to non-duplicating CFTGs and MCFGs.

Example 3.9. Let a non-duplicating CFTG $G$ consist of the following productions:

$$
S \rightarrow P \mathrm{ab}, \quad P\left[x_{1}, x_{2}\right] \rightarrow P\left(\mathrm{c} x_{1}\right)(\mathrm{c} S) \mid \mathrm{d} x_{1} x_{2},
$$

where the ranks of $S, P, \mathrm{a}, \mathrm{b}, \mathrm{c}, \mathrm{d}$ are $0,2,0,0,1,2$, respectively. De Groote and Pogodalla's method transforms $G$ into the following affine ACG $\mathscr{G}$ :

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}(\mathrm{x})$ |
| :---: | :---: | :---: |
| A | $p \rightarrow s$ | $\lambda y_{p}^{o^{2} \rightarrow o} \cdot y_{p} \mathrm{a}^{o} \mathrm{~b}^{o}$ |
| B | $s \rightarrow p \rightarrow p$ | $\lambda y_{s}^{o} y_{p}^{o^{2} \rightarrow o} x_{1}^{o} x_{2}^{o} \cdot y_{p}\left(\mathrm{c}^{o \rightarrow o} x_{1}\right)\left(\mathrm{c}^{o \rightarrow o} y_{s}\right)$ |
| C | $p$ | $\lambda x_{1}^{o} x_{2}^{o} . \mathrm{d}^{\mathrm{o}^{2} \rightarrow o} x_{1} x_{2}$ |

When we apply the linearization method given in Definition 3.6 to $\mathscr{G}$, we get the following linear ACG $\mathscr{G}^{l}$ whose distinguished type is $[s, o]$ :

| $x \in \mathscr{C}_{0}^{l}$ | $\mathscr{L}^{\prime}(\mathrm{x})$ |
| :---: | :---: |
| $\tau_{0}^{l}(\mathrm{x})$ |  |
| 【A, $\lambda y_{p}^{o \rightarrow o \rightarrow o} . y_{p} \mathrm{ab} \rrbracket$ | $\lambda y_{p}^{o^{2} \rightarrow o} . y_{p} \mathrm{ab}$ |
| $[p, o \rightarrow o \rightarrow o] \rightarrow[s, o]$ |  |
| 【A,$\lambda y_{p}^{o \rightarrow \bar{\sigma} \rightarrow o} . y_{p} \mathrm{a} \overline{\mathrm{b}} \rrbracket$ | $\lambda y_{p}^{o \rightarrow o} . y_{p} \mathrm{a}$ |
| $[p, o \rightarrow \bar{o} \rightarrow o] \rightarrow[s, o]$ |  |
| [ $\left.\mathrm{B}, \lambda y_{s}^{o} y_{p}^{o \rightarrow o \rightarrow o} x_{1}^{o} x_{2}^{\bar{\sigma}} \cdot y_{p}\left(\mathrm{c} x_{1}\right)\left(\mathrm{c} y_{s}\right)\right]$ | $\lambda y_{s}^{o} y_{p}^{o^{2} \rightarrow o} x_{1}^{o} \cdot y_{p}\left(\mathrm{c} x_{1}\right)\left(\mathrm{c} y_{s}\right)$ |
| $[s, o] \rightarrow[p, o \rightarrow o \rightarrow o] \rightarrow[p, o \rightarrow \bar{o} \rightarrow o]$ |  |
| [ $\left.\mathrm{B}, \lambda y_{s}^{\bar{\sigma}} y_{p}^{o \rightarrow \bar{\sigma} \rightarrow o} x_{1}^{o} x_{2}^{\bar{\sigma}} \cdot y_{p}\left(\mathrm{c} x_{1}\right)\left(\overline{\mathrm{c}} y_{s}\right)\right]$ | $\lambda y_{p}^{o \rightarrow o} x_{1}^{o} \cdot y_{p}\left(\mathrm{c} x_{1}\right)$ |
| $[p, o \rightarrow \bar{o} \rightarrow o] \rightarrow[p, o \rightarrow \bar{o} \rightarrow o]$ |  |
| [C, $\lambda x_{1}^{o} x_{2}^{o} . \mathrm{d} x_{1} x_{2} \rrbracket$ | $\lambda x_{1}^{o} x_{2}^{o} . \mathrm{d} x_{1} x_{2}$ |
| $[p, o \rightarrow o \rightarrow o]$ |  |

The linearized ACG $\mathscr{G}^{l}$ is actually the encoding of the linear CFTG $G^{\prime}$ consisting of the following productions:

$$
\begin{gathered}
S \rightarrow P(\mathrm{a}, \mathrm{~b})\left|P^{\prime}(\mathrm{a}), \quad P^{\prime}\left(x_{1}\right) \rightarrow P\left(\mathrm{c}\left(x_{1}\right), \mathrm{c}(S)\right)\right| P^{\prime}\left(\mathrm{c}\left(x_{1}\right)\right), \\
P\left(x_{1}, x_{2}\right) \rightarrow \mathrm{d}\left(x_{1}, x_{2}\right),
\end{gathered}
$$

where the ranks of nonterminals $S, P, P^{\prime}$ are $0,2,1$, respectively. $G, \mathscr{G}, \mathscr{G}^{l}$, and $G^{\prime}$ generate the same tree language.

The following corollary generalizes the result by Fujiyoshi [10], which covers the monadic case only.

Corollary 3.10. For every non-duplicating $\operatorname{CFTG} G$, there is a linear CFTG $G^{\prime}$ such that $G$ and $G^{\prime}$ generate the same tree language.

De Groote and Pogodalla [19]'s encoding method of LCFRSs by linear ACGs described in Section 2.2.1 can be applied to MCFGs also. Let $\mathscr{G}$ be the affine ACG that encodes an MCFG $G$. The linearized ACG $\mathscr{G}^{l}$ is indeed in the form that is the encoding of an LCFRS, while $\mathscr{G}^{\prime}$ is not. We here explain that lexical entries of $\mathscr{G}^{\prime}$ that cannot be interpreted as the encodings of productions of an LCFRS are useless. For instance, for the production rule $S \rightarrow f()$ with $f()=\langle\mathrm{a}\rangle$ of an MCFG, $\mathscr{G}$ has the corresponding abstract constant A of type $S$ that is mapped to the term

$$
\lambda z^{s t r \rightarrow s t r} . z \mathrm{a} .
$$

The linearization method in Definition 3.6 lets $\mathscr{G}^{\prime}$ have the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}^{\prime}$ | $\tau_{0}^{\prime}(\mathrm{x})$ | $\mathscr{L}^{\prime}(\mathrm{x})$ |
| :---: | :---: | :---: |
| 【A，$\lambda z^{\text {str } \rightarrow s t r} . z \mathrm{a} \rrbracket$ | $[S,(s t r \rightarrow s t r) \rightarrow s t r]$ | $\lambda z^{\text {str } \rightarrow \text { str }} . z \mathrm{a}$ |
|  | $[S,(s t r \rightarrow \bar{o} \rightarrow o) \rightarrow \bar{o} \rightarrow o]$ | $\lambda z^{s t r \rightarrow o} . z \mathrm{a}$ |
| 【A，$\lambda z^{s t r r} \rightarrow s t r . z \bar{a} \rrbracket$ | $[S,(\overline{s t r} \rightarrow s t r) \rightarrow s t r]$ | $\lambda z^{\text {str }} . z$ |
| 【A,$\left.\lambda z^{\text {str }}+\bar{o} \rightarrow o . ~ z \bar{a} \rrbracket\right]$ | $[S,(\overline{s t r} \rightarrow \bar{o} \rightarrow o) \rightarrow \bar{o} \rightarrow o]$ | $\lambda z^{o} . z$ |

Recall that a nonterminal symbol of rank $k$ in an LCFRS is encoded as an abstract atomic type mapped to $\left(s t r^{k} \rightarrow s t r\right) \rightarrow s t r$ by the lexicon in the corresponding linear ACG．In the above table，the new atomic types $[S,(\operatorname{str} \rightarrow \bar{o} \rightarrow o) \rightarrow \bar{o} \rightarrow o]$ and $[S,(\overline{s t r} \rightarrow \bar{o} \rightarrow o) \rightarrow \bar{o} \rightarrow o]$ are，how－ ever，mapped to $(s t r \rightarrow o) \rightarrow o$ and $o \rightarrow o$ ，respectively．Those atomic types cannot be interpreted as the encodings of nonterminals of an LCFRS． Consequently，the abstract constants $\llbracket \mathrm{A}, \lambda z^{s t r \rightarrow \bar{o} \rightarrow o} . z a \rrbracket$ and $\llbracket \mathrm{A}, \lambda z^{\overline{s t r} \rightarrow \bar{o} \rightarrow o} . z \overline{\mathrm{a}} \rrbracket$ cannot be interpreted as the encodings of some productions of an LCFRS． We show that indeed those abstract constants are useless．

Let

$$
\Gamma=\left\{\left(\alpha_{1} \rightarrow \cdots \rightarrow \alpha_{k} \rightarrow \operatorname{str}\right) \rightarrow \operatorname{str} \mid \alpha_{i} \in\{s t r, \overline{s t r}\}\right\} .
$$

We prove that for every new abstract atomic type $[p, \alpha]$ with $p \neq s^{\prime}$ ，where $s^{\prime}$ is the distinguished type of the affine $\operatorname{ACG} \mathscr{G}$ ，if $\alpha \notin \Gamma$ ，then $[p, \alpha]$ is useless in $\mathscr{G}^{\prime}$ ．Let

$$
\Delta=\left\{\left(\alpha_{1} \rightarrow \cdots \rightarrow \alpha_{k} \rightarrow \bar{o} \rightarrow o\right) \rightarrow \bar{o} \rightarrow o \mid \alpha_{i} \in\{s t r, \overline{s t r}\}\right\}
$$

First we show that if a variable－free term $N \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}^{\prime}\right)$ has an atomic type $[p, \gamma]$ with $p \neq s^{\prime}$ ，then $\gamma \in \Gamma \cup \Delta$ holds，and moreover，if a subterm $N^{\prime}$ of $N$ has a type $\left[q, \gamma^{\prime}\right]$ ，then either $\gamma, \gamma^{\prime} \in \Gamma$ or $\gamma, \gamma^{\prime} \in \Delta$ ．Let $\llbracket \mathrm{A}, Q \rrbracket$ be the head of $N, \tau_{0}(\mathrm{~A})=p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow q, \tau_{1}^{\prime}(Q)=\alpha_{1} \rightarrow \cdots \rightarrow \alpha_{m} \rightarrow \beta$ ．By $\tau_{0}^{\prime}(\llbracket \mathrm{A}, Q \rrbracket)=\left(\left[p_{1}, \alpha_{1}\right] \rightarrow \cdots \rightarrow\left[p_{m}, \alpha_{m}\right] \rightarrow[q, \beta]\right)^{\ddagger}, N$ has the form

$$
N \equiv \llbracket \mathrm{~A}, Q \rrbracket N_{i_{1}} \ldots N_{i_{n}}
$$

where $\left\{i_{1}, \ldots, i_{n}\right\}=\left\{i \mid \alpha_{i} \notin \mathcal{T}\left(\overline{\mathscr{A}_{1}}\right)\right\}$ ．Let

$$
Q \equiv \lambda y_{1}^{\alpha_{1}} \ldots y_{m}^{\alpha_{m}} z^{\beta_{0}} \cdot y_{1}\left(\lambda \vec{x}_{1} \ldots y_{m}\left(\lambda \vec{x}_{m} . z Q_{1} \ldots Q_{k}\right) \ldots\right)
$$

where $\beta=\beta_{0} \rightarrow \gamma$ for $\gamma \in\{$ str， $\bar{o} \rightarrow o\}$ ．Then，$\alpha_{1}$ has the form $\alpha_{1}^{\prime} \rightarrow \gamma \notin$ $\mathcal{T}\left(\overline{\mathscr{A}_{1}}\right)$ ．Thus，$i_{1}=1$ ．By applying the induction hypothesis to $N_{1}$ ，we see $\alpha_{1} \in \Gamma \cup \Delta$ and $\alpha_{1}=\left(\tau_{1}^{\prime}\left(\vec{x}_{1}\right) \rightarrow \gamma\right) \rightarrow \gamma$ ．Applying the same discussion to $\alpha_{2}, \ldots, \alpha_{m}$ repeatedly，we have $n=m, i_{j}=j$ ，and $\alpha_{i}=\left(\tau_{1}^{\prime}\left(\vec{x}_{i}\right) \rightarrow \gamma\right) \rightarrow \gamma \in$ $\Gamma \cup \Delta . \alpha_{i} \in \Gamma \cup \Delta$ implies that the type of each variable in $\vec{x}_{i}$ is either str or $\overline{s t r}$ ．Thus，$Q_{j}$ consists of atomic terms of types in $\{s t r, \overline{s t r}\}$ and hence the type of each $Q_{j}$ is either str or $\overline{s t r}$ ．Therefore，

$$
\beta=\beta_{0} \rightarrow \gamma=\left(\tau_{1}^{\prime}\left(Q_{1}\right) \rightarrow \cdots \rightarrow \tau_{1}^{\prime}\left(Q_{m}\right) \rightarrow \gamma\right) \rightarrow \gamma \in \Gamma \cup \Delta .
$$

Let $N \in \mathcal{A}^{\operatorname{lin}}\left(\mathscr{G}^{\prime}\right) . N$ has the form $\llbracket \mathrm{S}, Q \rrbracket N^{\prime}$, where $\tau_{0}^{\prime}(\llbracket \mathrm{S}, Q \rrbracket)=[S, \alpha] \rightarrow$ [ $\left.s^{\prime}, s t r\right]$. By applying the above claim to $N^{\prime}$, we have $\alpha \in \Gamma \cup \Delta$. Let

$$
Q \equiv \lambda y^{\left(\alpha_{1} \rightarrow \alpha_{2}\right) \rightarrow \alpha_{3}} \cdot y\left(\lambda x^{\alpha_{4}} \cdot x\right) \in \Pi(\mathscr{L}(\mathrm{S}))=\Pi\left(\lambda y^{(s t r \rightarrow s t r) \rightarrow s t r} \cdot y\left(\lambda x^{s t r} \cdot x\right)\right),
$$

where $\alpha=\left(\alpha_{1} \rightarrow \alpha_{2}\right) \rightarrow \alpha_{3}$. We have $\alpha_{3}=\operatorname{str}$ by $\tau_{1}^{\prime}(Q)=\alpha \rightarrow \operatorname{str}$, $\alpha_{1} \rightarrow \alpha_{2}=\alpha_{4} \rightarrow \alpha_{4}$ by the well-typedness of $Q$, and $\alpha_{2}=\alpha_{3}$ by $\alpha \in \Gamma \cup \Delta$. Therefore, $Q \equiv \mathscr{L}(\mathrm{~S})$ and $\alpha=(s t r \rightarrow \operatorname{str}) \rightarrow \operatorname{str} \in \Gamma$. The above claim entails that if a subterm of $N \in \mathcal{A}\left(\mathscr{G}^{\prime}\right)$ has a type $[q, \gamma]$ with $q \neq s^{\prime}$, then $\gamma \in \Gamma$. Therefore, every useful abstract atomic type of $\mathscr{G}^{\prime}$ is mapped to $\left(s t r^{k} \rightarrow s t r\right) \rightarrow s t r$ for some non-negative integer $k$, which is interpreted as the encoding of a nonterminal of rank $k$ in an LCFRS.

Note that the LCFRS obtained from an MCFG through our linearization method may have nonterminals of rank 0 , though usual definitions of an LCFRS do not allow nonterminals to have rank 0 . Mathematically speaking, allowing or disallowing nonterminals of rank 0 does not matter at all. The reason why usual definitions of an LCFRS do not allow nonterminals to have rank 0 is just to avoid redundancy. If a nonterminal $A_{k}$ has rank 0 , and a production $A_{k} \rightarrow f\left(B_{1}, \ldots, B_{m}\right)$ is in an LCFRS, then $f$ is the function that maps the empty sequence to the empty sequence and each $B_{i}$ has rank 0 . If $A_{k}$ appears in the right-hand side of a production as $C \rightarrow g\left(A_{1}, \ldots, A_{k}, \ldots, A_{n}\right)$, we can eliminate $A_{k}$ from the right-hand side without modifying the function $g$, which is a function from $\prod_{1 \leq i \leq n}\left(T^{*}\right)^{\rho\left(A_{i}\right)}=\prod_{\substack{1 \leq i \leq n \\ i \neq k}}\left(T^{*}\right)^{\rho\left(A_{i}\right)}$ to $\left(T^{*}\right)^{\rho(C)}$. Note that here we work modulo the associativity of the cartesian product, i.e., we identify $\left(T^{*}\right)^{n} \times\left(T^{*}\right)^{m}$ with $\left(T^{*}\right)^{n+m}$. Thus, we can simply ignore nonterminals of rank 0 if they occur in an LCFRS.

Therefore, our result covers the following theorem shown by Seki et al. [49]
Corollary 3.11. For every $M C F G G$, there is an $L C F R S G^{\prime}$ such that the languages generated by $G$ and $G^{\prime}$ coincide.

## Third or Higher-Order Case

Definition 3.6 itself does not depend on the order of the given affine ACG except for checking whether useless abstract atomic types or useless atomic constants exist. For the general case, however, the linearized ACG given in Definition 3.6 may generate a strictly larger language than the original affine ACG. In the remainder of this chapter, we present a linearization method for general affine ACGs.

Example 3.12. Suppose that an affine ACG $\mathscr{G} \in \mathbf{G}^{\text {aff }}(3,1)$ consists of the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}(\mathrm{x})$ |
| :---: | :---: | :---: |
| A | $q$ | $\#$ |
| B | $p \rightarrow q \rightarrow q$ | $\lambda y^{o} z^{o} \cdot \mathrm{~b} z$ |
| C | $q \rightarrow s$ | $\lambda z^{o} . z$ |
| D | $(p \rightarrow s) \rightarrow s$ | $\lambda x^{o \rightarrow o} . \mathrm{a}(x \mathrm{e})$ |

We have

$$
\begin{aligned}
& \mathcal{A}^{\operatorname{lin}(\mathscr{G})}=\left\{\begin{array}{l}
\mathrm{D}\left(\lambda y_{1}^{p} \cdot \mathrm{D}\left(\lambda y_{2}^{p} \ldots . \mathrm{D}\left(\lambda y_{n}^{p} \cdot \mathrm{C}\left(\mathrm{~B} y_{i_{1}}\left(\mathrm{~B} y_{i_{2}}\left(\ldots\left(\mathrm{~B} y_{i_{n}} \mathrm{~A}\right) \ldots\right)\right)\right)\right) \ldots\right)\right) \\
\left.\quad \mid n \geq 0,\left\{i_{1}, \ldots, i_{n}\right\}=\{1, \ldots, n\}\right\}
\end{array}\right. \\
& \mathcal{O}^{\operatorname{lin}(\mathscr{G})}=\{\overbrace{\mathrm{a}(\ldots(\mathrm{a}(\overbrace{\mathrm{~b}(\ldots(\mathrm{~b}}^{n \text { times }} \#) \ldots)) \ldots) \mid n \geq 0\} .}^{n \text { times }} .
\end{aligned}
$$

$\mathscr{G}^{\prime}$ consists of the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}^{\prime}$ | $\tau_{0}^{\prime}(\mathrm{x})$ | $\mathscr{L}^{\prime}(\mathrm{x})$ |
| :---: | :---: | :---: |
| $\llbracket \mathrm{A}, \# \rrbracket$ | $[q, o]$ | $\#$ |
| $\llbracket \mathrm{~B}, \lambda y^{\bar{o}} z^{\circ} \cdot \mathrm{b} z \rrbracket$ | $[q, o] \rightarrow[q, o]$ | $\lambda z^{o} \cdot \mathrm{~b} z$ |
| $\llbracket \mathrm{C}, \lambda z^{o} . z \rrbracket$ | $[q, o] \rightarrow[s, o]$ | $\lambda z^{o} . z$ |
| $\llbracket \mathrm{D}, \lambda x^{\bar{\sigma} \rightarrow o} . \mathrm{a}(x \overline{\mathrm{e}}) \rrbracket$ | $[s, o] \rightarrow[s, o]$ | $\lambda x^{o} . \mathrm{a} x$ |
| $\llbracket \mathrm{D}, \lambda x^{o \rightarrow o} . \mathrm{a}(x \mathrm{e}) \rrbracket$ | $([p, o] \rightarrow[s, o]) \rightarrow[s, o]$ | $\lambda x^{o \rightarrow o} . \mathrm{a}(x \mathrm{e})$ |

The last lexical entry is useless. We have

$$
\begin{aligned}
& \mathcal{A}^{\operatorname{lin}\left(\mathscr{G}^{\prime}\right)}=\{\overbrace{\mathrm{D}^{\prime}\left(\ldots \left(\mathrm{D}^{\prime}\right.\right.}^{m \text { times }}(\mathrm{C}^{\prime}(\overbrace{\left.\left.\left.\mathrm{B}^{\prime}\left(\ldots\left(\mathrm{B}^{\prime} \mathrm{A}^{\prime}\right) \ldots\right)\right) \ldots\right) \mid m, n \geq 0\right\}}^{n \text {-times }} \\
& \mathcal{O}^{\operatorname{lin}}\left(\mathscr{G}^{\prime}\right)=\{\overbrace{\mathrm{a}(\ldots(\mathrm{a}(\mathrm{~b}(\ldots(\mathrm{~b}}^{m \text {-imes }} \overbrace{\text {-times }}^{n} \ldots)) \ldots) \mid m, n \geq 0\} \supsetneq \mathcal{O}^{\operatorname{lin}}(\mathscr{G})
\end{aligned}
$$

where $\mathrm{A}^{\prime}, \mathrm{B}^{\prime}, \mathrm{C}^{\prime}, \mathrm{D}^{\prime}$ are, respectively, the first, second, third, fourth abstract constants of $\mathscr{G}^{\prime}$ shown in the above table. Though any term of type $p$ that is the first argument of an occurrence of B is bound to be erased in the original ACG $\mathscr{G}$, we cannot ignore the occurrence of the type $p$, because that occurrence of $p$ balances the numbers of occurrences of B and D in a term in $\mathcal{A}^{\operatorname{lin}}(\mathscr{G})$.

Our alternative linearization method presented later gives the linear ACG $\mathscr{G}^{\prime \prime}$ consisting of the following lexical entries (useless lexical entries are sup-
pressed):

| $\mathrm{x} \in \mathscr{C}_{0}^{\prime \prime}$ | $\tau_{0}^{\prime \prime}(\mathrm{x})$ | $\mathscr{L}^{\prime \prime}(\mathrm{x})$ |
| :---: | :---: | :---: |
| $\llbracket \mathrm{A}, \# \rrbracket$ | $[q, o]$ | $\#$ |
| $\llbracket \mathrm{~B}, \lambda y^{\bar{\sigma}} z^{o} . \mathrm{b} z \rrbracket$ | $[p, \bar{o}] \rightarrow[q, o] \rightarrow[q, o]$ | $\lambda y^{o \rightarrow o} z^{o} . y(\mathrm{~b} z)$ |
| $\llbracket \mathrm{C}, \lambda z^{o} . z \rrbracket$ | $[q, o] \rightarrow[s, o]$ | $\lambda z^{o} . z$ |
| $\llbracket \mathrm{D}, \lambda x^{\bar{o} \rightarrow o} . \mathrm{a}(x \overline{\mathrm{e}}) \rrbracket$ | $([p, \bar{o}] \rightarrow[s, o]) \rightarrow[s, o]$ | $\lambda x^{(o \rightarrow o) \rightarrow o} . \mathrm{a}\left(x\left(\lambda z^{o} . z\right)\right)$ |

where $[p, \bar{o}]$ is mapped to $o \rightarrow o$. For

$$
M \equiv \mathrm{D}\left(\lambda y^{p} \cdot \mathrm{C}(\mathrm{~B} y \mathrm{~A})\right) \in \mathcal{A}^{\operatorname{lin}}(\mathscr{G})
$$

we have

$$
N \equiv \mathrm{D}^{\prime \prime}\left(\lambda y^{[p, \bar{\sigma}]} \cdot \mathrm{C}^{\prime \prime}\left(\mathrm{B}^{\prime \prime} y \mathrm{~A}^{\prime \prime}\right)\right) \in \mathcal{A}^{\operatorname{lin}}\left(\mathscr{G}^{\prime \prime}\right)
$$

where $A^{\prime \prime}, B^{\prime \prime}, C^{\prime \prime}, D^{\prime \prime}$ are, respectively, the first, second, third, fourth abstract constants of $\mathscr{G}^{\prime \prime}$ shown in the above table. It is easy to check that $\mathscr{L}(M)=$ $\mathscr{L}^{\prime \prime}(N)=\mathrm{a}(\mathrm{b} \#)$. We have $\mathcal{O}(\mathscr{G})=\mathcal{O}\left(\mathscr{G}^{\prime \prime}\right)$.

Now, we give the formal definition of our new linearization method for general affine ACGs. For simplicity, we assume that $\mathscr{A}_{1}=\{o\}$ here, but it is possible to lift this assumption. The new linearized ACG $\mathscr{G}^{\prime \prime}$ has the form $\mathscr{G}^{\prime \prime}=\left\langle\Sigma_{0}^{\prime \prime}, \Sigma_{1}, \mathscr{L}^{\prime \prime},[s, \mathscr{L}(s)]\right\rangle$, where $\Sigma_{0}^{\prime \prime}$ is defined by

$$
\begin{aligned}
\mathscr{A}_{0}^{\prime \prime} & =\left\{[p, \beta] \mid p \in \mathscr{A}_{0}, \beta \in \Pi(\mathscr{L}(p))\right\} \\
\mathscr{C}_{0}^{\prime \prime} & =\left\{\llbracket \mathrm{a}, Q \rrbracket \mid \mathrm{a} \in \mathscr{C}_{0}, Q \in \Pi(\mathscr{L}(\mathrm{a}))\right\} \\
\tau_{0}^{\prime \prime} & =\left\{\llbracket \mathrm{a}, Q \rrbracket \mapsto\left[\tau_{0}(\mathrm{a}), \tau_{1}^{\prime}(Q)\right]\right\} \text { where }[\alpha \rightarrow \gamma, \beta \rightarrow \delta]=[\alpha, \beta] \rightarrow[\gamma, \delta] .
\end{aligned}
$$

Here we have two simple lexicons $\mathscr{L}_{0}: \Sigma_{0}^{\prime \prime} \rightarrow \Sigma_{0}$ and $\mathscr{L}_{1}: \Sigma_{0}^{\prime \prime} \rightarrow \Sigma_{1}^{\prime}$ such that

$$
\begin{array}{ll}
\mathscr{L}_{0}([p, \beta])=p, & \mathscr{L}_{0}(\llbracket \mathrm{a}, Q \rrbracket)=\mathrm{a} \\
\mathscr{L}_{1}([p, \beta])=\beta, & \mathscr{L}_{1}(\llbracket \mathrm{a}, Q \rrbracket)=Q .
\end{array}
$$

$\mathscr{L}_{0}$ is relabeling. We have $\widetilde{\mathscr{L}_{1}(N)} \equiv \mathscr{L} \circ \mathscr{L}_{0}(N)$ for $N \in \Lambda^{\operatorname{lin}\left(\Sigma_{0}^{\prime \prime}\right)}$.
Lemma 3.13. For every $Q \in \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$ and $\alpha \in \mathcal{T}\left(\mathscr{A}_{0}\right)$, the following statements are equivalent:
(i) There is $M \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}\right)$ of type $\alpha$ such that $\mathscr{L}(M) \equiv \widetilde{Q}$.
(ii) There is $N \in \Lambda^{\operatorname{lin}}\left(\Sigma_{0}^{\prime \prime}\right)$ of type $\left[\alpha, \tau_{1}^{\prime}(Q)\right]$ such that $\mathscr{L}_{1}(N) \equiv Q$.

Proof. The direction $[(\mathrm{ii}) \Rightarrow(\mathrm{i})]$ can be shown by letting $M \equiv \mathscr{L}_{0}(N)$.
The opposite direction $[(\mathrm{i}) \Rightarrow$ (ii)] can be shown by induction on $M$. Let $N \equiv \chi(M, Q)$ where

- $\chi(\mathrm{a}, Q) \equiv \llbracket \mathrm{a}, Q \rrbracket$,
- $\chi\left(x^{\alpha}, x^{\beta}\right) \equiv x^{[\alpha, \beta]}$,
- $\chi\left(M_{1} M_{2}, Q_{1} Q_{2}\right) \equiv \chi\left(M_{1}, Q_{1}\right) \chi\left(M_{2}, Q_{2}\right)$,
- $\chi\left(\lambda x^{\alpha} \cdot M^{\prime}, \lambda x^{\beta} \cdot Q^{\prime}\right) \equiv \lambda x^{[\alpha, \beta]} \cdot \chi\left(M^{\prime}, Q^{\prime}\right)$.

Here, $\mathscr{L}_{0}(\chi(M, Q)) \equiv M, \mathscr{L}_{1}(\chi(M, Q)) \equiv Q$, and $\chi\left(\mathscr{L}_{0}(N), \mathscr{L}_{1}(N)\right) \equiv N$ hold.

Lemmas 3.5 and 3.13 imply

$$
\begin{equation*}
\left\{M \in \mathcal{A}(\mathscr{G})\left||\mathscr{L}(M)|_{\beta} \text { is linear }\right\}=\left\{\mathscr{L}_{0}(N) \mid N \in \mathcal{A}\left(\mathscr{G}^{\prime \prime}\right)\right\} .\right. \tag{3.5}
\end{equation*}
$$

Since $\left(\mathscr{L}_{1}(N)\right)^{\dagger}={ }_{\beta} \widetilde{\mathscr{L}_{1}(N)} \equiv \mathscr{L} \circ \mathscr{L}_{0}(N)$ for every $N \in \mathcal{A}\left(\mathscr{G}^{\prime \prime}\right)$ by Lemma 3.4, it is enough to define a new lexicon $\mathscr{L}^{\prime \prime}$ so that

$$
\begin{equation*}
\mathscr{L}^{\prime \prime}(N)={ }_{\beta \eta}\left(\mathscr{L}_{1}(N)\right)^{\dagger} \tag{3.6}
\end{equation*}
$$

for every $N \in \mathcal{A}\left(\mathscr{G}^{\prime \prime}\right)$.
We define the type substitution $\sigma: \mathscr{A}_{0}^{\prime \prime} \rightarrow \mathcal{T}(\{o\})$ of $\mathscr{L}^{\prime \prime}=\langle\sigma, \theta\rangle$ as

$$
\sigma([p, \beta])= \begin{cases}(\beta)^{\dagger} & \text { if } \beta \notin \mathcal{T}(\{\bar{o}\}) \\ o \rightarrow o & \text { if } \beta \in \mathcal{T}(\{\bar{o}\})\end{cases}
$$

Here we identify $\sigma$ with its homomorphic extension. For each $[\alpha, \beta] \in \mathcal{T}\left(\mathscr{A}_{0}^{\prime \prime}\right)$ such that $\beta \in \widehat{\mathcal{T}}(\{o\})-\mathcal{T}(\{\bar{o}\})$, we define two linear combinators $X_{\alpha}^{\beta}$ of type $\sigma([\alpha, \beta]) \rightarrow(\beta)^{\dagger}$ and $Y_{\alpha}^{\beta}$ of type $(\beta)^{\dagger} \rightarrow \sigma([\alpha, \beta])$ by mutual induction. Let $[\alpha, \beta]=\left[\alpha_{1}, \beta_{1}\right] \rightarrow \cdots \rightarrow\left[\alpha_{m}, \beta_{m}\right] \rightarrow\left[p, \beta_{0}\right]$ with $\left[p, \beta_{0}\right] \in \mathscr{A}_{0}^{\prime \prime}$ and the set $\{1, \ldots, m\}$ be partitioned into two subsets $I$ and $J$ so that $\beta_{i} \notin \mathcal{T}(\{\bar{o}\})$ iff $i \in$ I. Let $I=\left\{i_{1}, \ldots, i_{k}\right\}$ with $i_{j}<i_{j+1}$ for each $1 \leq j<k$ and $J=\left\{j_{1}, \ldots, j_{l}\right\}$. If $i \in J$, then the linear combinators $Z^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right)}$ and $Z^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right) \rightarrow o \rightarrow o}$ (given in Lemma 2.12) are well-defined by the definition of $\sigma$. We define

$$
\begin{aligned}
X_{\alpha}^{\beta} & \equiv \lambda y^{\sigma([\alpha, \beta])} x_{i_{1}}^{\left(\beta_{i_{1}}\right)^{\dagger}} \ldots x_{i_{k}}^{\left(\beta_{i_{k}}\right)^{\dagger}} \cdot y^{\sigma([\alpha, \beta])} R_{1} \ldots R_{m} \\
\text { where } \quad R_{i} & \equiv \begin{cases}Y_{\alpha_{i}}^{\beta_{i}} x_{i}^{\left(\beta_{i}\right)^{\dagger}} & \text { if } i \in I, \\
Z^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right)} & \text { if } i \in J,\end{cases}
\end{aligned}
$$

and

$$
Y_{\alpha}^{\beta} \equiv \lambda x^{(\beta)^{\dagger}} y_{1}^{\sigma\left(\left[\alpha_{1}, \beta_{1}\right]\right)} \ldots y_{m}^{\sigma\left(\left[\alpha_{m}, \beta_{m}\right]\right)} \vec{z} \cdot M_{j_{1}}\left(\ldots\left(M_{j_{l}}\left(x^{(\beta)^{\dagger}} L_{i_{1}} \ldots L_{i_{k}} \vec{z}\right)\right) \ldots\right)
$$

where $\vec{z}$ is short for $z_{1}^{\gamma_{1}} \ldots z_{n}^{\gamma_{n}}$ for $\left(\beta_{0}\right)^{\dagger}=\gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n} \rightarrow o$, and

$$
\begin{cases}L_{i} \equiv X_{\alpha_{i}}^{\beta_{i}} y_{i}^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right)} & \text { for } i \in I \\ M_{i} \equiv Z^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right) \rightarrow o \rightarrow o} y_{i}^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right)} & \text { for } i \in J .\end{cases}
$$

Note that if $[\alpha, \beta]=\left[p, \beta_{0}\right] \in \mathscr{A}_{0}^{\prime \prime}$, then $X_{p}^{\beta_{0}}={ }_{\beta \eta} Y_{p}^{\beta_{0}}={ }_{\beta_{\eta}} \lambda z^{\left(\beta_{0}\right)^{\dagger}} . z$.
Now, we give a new linearization method as follows.
Definition 3.14. For a given affine ACG $\mathscr{G}$, we define a new linear ACG as $\mathscr{G}^{\prime \prime}=\left\langle\Sigma_{0}^{\prime \prime}, \Sigma_{1}, \mathscr{L}^{\prime \prime},[s, \mathscr{L}(s)]\right\rangle$, where $\mathscr{L}^{\prime \prime}=\langle\sigma, \theta\rangle$ for $\sigma$ as above and

$$
\theta(\llbracket \mathrm{a}, Q \rrbracket) \equiv \begin{cases}\left|Y_{\tau_{0}}^{\tau_{1}^{\prime}(Q)}(Q)(Q)^{\dagger}\right|_{\beta} & \text { if } Q \notin \Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right), \\ Z^{\sigma\left(\tau_{0}^{\prime \prime}(\llbracket a, Q \rrbracket)\right)} & \text { if } Q \in \Lambda^{\text {aff }}\left(\overline{\Sigma_{1}}\right) .\end{cases}
$$

Thus, if the given affine ACG $\mathscr{G}$ belongs to $\mathbf{G}^{\text {aff }}(m, n)$, then $\mathscr{G}^{\prime \prime}$ belongs to $\mathbf{G}^{\operatorname{lin}}(m, \max \{2, n\})$.
Lemma 3.15. Given $N \in \Lambda\left(\Sigma_{0}^{\prime \prime}\right)$ of type $[\alpha, \beta]$ such that $\beta \notin \mathcal{T}(\{\bar{o}\})$ and $\mathscr{L}_{1}(N) \in \widehat{\Lambda}^{\text {aff }}\left(\Sigma_{1}\right)$, we have

$$
\left(\mathscr{L}_{1}(N)\right)^{\dagger}={ }_{\beta \eta} X_{\alpha}^{\beta} \mathscr{L}^{\prime \prime}(N) \phi_{N}
$$

where $\phi_{N}$ denotes the substitution on the free variables of $\mathscr{L}^{\prime \prime}(N)$ such that

$$
x^{\sigma([\alpha, \beta])} \phi_{N}= \begin{cases}Y_{\alpha}^{\beta} x^{(\beta)^{\dagger}} & \text { if } x \text { has the type }[\alpha, \beta] \text { in } N \text { and } \beta \notin \mathcal{T}(\{\bar{o}\}), \\ Z^{\sigma([\alpha, \beta])} & \text { otherwise. }\end{cases}
$$

Proof. We assume that $N$ is in long normal form for simplicity. The proof is done by induction on $N$.

Case 1. Suppose that $N=\lambda x_{1}^{\left[\alpha_{1}, \beta_{1}\right]} \ldots x_{m}^{\left[\alpha_{m}, \beta_{m}\right]} . N^{\prime}$ have type $[\alpha, \beta]=$ $\left[\alpha_{1}, \beta_{1}\right] \rightarrow \cdots \rightarrow\left[\alpha_{m}, \beta_{m}\right] \rightarrow\left[p, \beta_{0}\right]$ with $m \geq 1$. Let us partition the set $\{1, \ldots, m\}$ into two subsets $I$ and $J$ so that $\beta_{i} \notin \mathcal{T}(\{\bar{o}\})$ iff $i \in I$ and $\beta_{i} \in \mathcal{T}(\{\bar{o}\})$ iff $i \in J$. Let the elements of $I$ be $i_{1}, \ldots, i_{k}$ with $i_{j}<i_{j+1}$ for $1 \leq j<k$. Since $\operatorname{Fv}\left(N^{\prime}\right)=\operatorname{Fv}(N) \cup\left\{x_{1}, \ldots, x_{m}\right\}$, we have

$$
\begin{gathered}
\phi_{N^{\prime}}=\phi_{N} \cup\left\{x_{i}^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right)} \mapsto R_{i} \mid 1 \leq i \leq m\right\} \\
\text { where } R_{i}= \begin{cases}Y_{\alpha_{i}}^{\beta_{i}} x_{i} & \text { if } i \in I, \\
Z^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right)} & \text { if } i \in J .\end{cases}
\end{gathered}
$$

Thus, together with the induction hypothesis and $X_{p}^{\left.\beta_{0}\right]}=\lambda z^{\left(\beta_{0}\right)^{\dagger}} . . z$,

$$
\begin{align*}
\left(\mathscr{L}_{1}(N)\right)^{\dagger} & \equiv \lambda x_{i_{1}}^{\left(\beta_{i_{1}}\right)^{\dagger}} \ldots x_{i_{k}}^{\left(\beta_{i_{k}}\right)^{\dagger}} \cdot\left(\mathscr{L}_{1}\left(N^{\prime}\right)\right)^{\dagger} \\
& ={ }_{\beta \eta} \lambda x_{i_{1}}^{\left(\beta_{i_{1}}\right)^{\dagger}} \ldots x_{i_{k}}^{\left(\beta_{i_{k}}\right)^{\dagger}} \cdot\left(X_{p}^{\beta_{0}} \mathscr{L}^{\prime \prime}\left(N^{\prime}\right) \phi_{N^{\prime}}\right) \\
& \rightarrow_{\beta} \lambda x_{i_{1}}^{\left(\beta_{i_{1}}\right)^{\dagger}} \ldots x_{i_{k}}^{\left(\beta_{i_{k}}\right)^{\dagger}} \cdot\left(\mathscr{L}^{\prime \prime}\left(N^{\prime}\right)\left[R_{i} / x_{i}^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right)}\right]_{1 \leq i \leq m}\right) \phi_{N} . \tag{3.7}
\end{align*}
$$

On the other hand,

$$
\begin{align*}
& X_{\alpha}^{\beta} \mathscr{L}^{\prime \prime}(N) \phi_{N} \\
& \equiv\left(\lambda y^{\sigma([\alpha, \beta])} x_{i_{1}}^{\left(\beta_{i_{1}}\right)^{\dagger}} \ldots x_{i_{k}}^{\left(\beta_{i_{k}}\right)^{\dagger}} \cdot y R_{1} \ldots R_{m}\right) \mathscr{L}^{\prime \prime}\left(\lambda x_{1}^{\left[\alpha_{1}, \beta_{1}\right]} \ldots x_{m}^{\left[\alpha_{m}, \beta_{m}\right]} \cdot N^{\prime}\right) \phi_{N} \\
& \rightarrow_{\beta} \lambda x_{i_{1}}^{\left(\beta_{i_{1}}\right)^{\prime}} \ldots x_{i_{k}}^{\left(\beta_{i_{k}}\right)^{\dagger}} \cdot \mathscr{L}^{\prime \prime}\left(\lambda x_{1}^{\left[\alpha_{1}, \beta_{1}\right]} \ldots x_{m}^{\left[\alpha_{m}, \beta_{m}\right]} \cdot N^{\prime}\right) R_{1} \ldots R_{m} \phi_{N} \\
& \rightarrow_{\beta} \lambda x_{i_{1}}^{\left(\beta_{i_{1}}\right)^{\dagger}} \ldots x_{i_{k}}^{\left(\beta_{i_{k}}\right)^{\dagger}} \cdot\left(\mathscr{L}^{\prime \prime}\left(N^{\prime}\right)\left[R_{i} / x_{i}^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right)}\right]_{1 \leq i \leq m}\right) \phi_{N} . \tag{3.8}
\end{align*}
$$

(3.7) and (3.8) coincide.

Case 2. Suppose that $N=x^{[\alpha, \beta]} N_{1} \ldots N_{m}$ where $\alpha=\alpha_{1} \rightarrow \cdots \rightarrow$ $\alpha_{m} \rightarrow p, \beta=\beta_{1} \rightarrow \cdots \rightarrow \beta_{m} \rightarrow \beta_{0}$. Let us partition the set $\{1, \ldots, m\}$ into two subsets $I$ and $J$ so that $\beta_{i} \notin \mathcal{T}(\{\bar{o}\})$ iff $i \in I$ and $\beta_{i} \in \mathcal{T}(\{\bar{o}\})$ iff $i \in J$. Let the elements of $I$ be $i_{1}, \ldots, i_{k}$ with $i_{j}<i_{j+1}$ for $1 \leq j<k$ and $J=\left\{j_{1}, \ldots, j_{l}\right\}$. By the induction hypothesis,

$$
\begin{align*}
\left(\mathscr{L}_{1}(N)\right)^{\dagger} & \equiv x^{(\beta)^{\dagger}}\left(\mathscr{L}_{1}\left(N_{i_{1}}\right)\right)^{\dagger} \ldots\left(\mathscr{L}_{1}\left(N_{i_{k}}\right)\right)^{\dagger} \\
& ={ }_{\beta \eta} x^{(\beta)^{\dagger}}\left(X_{i_{1}} \mathscr{L}^{\prime \prime}\left(N_{i_{1}}\right) \phi_{N_{i_{1}}}\right) \ldots\left(X_{i_{k}} \mathscr{L}^{\prime \prime}\left(N_{i_{k}}\right) \phi_{N_{i_{k}}}\right) \tag{3.9}
\end{align*}
$$

where $X_{i}=X_{\alpha_{i}}^{\beta_{i}}$.
On the other hand, since $X_{p}^{\beta_{0}}$ is the identity,

$$
\begin{aligned}
& X_{p}^{\beta_{0}} \mathscr{L}^{\prime \prime}(N) \phi_{N} \rightarrow_{\beta} \mathscr{L}^{\prime \prime}(N) \phi_{N} \\
& \equiv\left(Y_{\alpha}^{\beta} x^{(\beta)^{\dagger}}\right)\left(\mathscr{L}^{\prime \prime}\left(N_{1}\right) \phi_{N_{1}}\right) \ldots\left(\mathscr{L}^{\prime \prime}\left(N_{m}\right) \phi_{N_{m}}\right) \\
& \rightarrow_{\beta}\left(\lambda y_{1}^{\sigma\left(\left[\alpha_{1}, \beta_{1}\right]\right)} \ldots y_{m}^{\sigma\left(\left[\alpha_{m}, \beta_{m}\right]\right)} \vec{z} \cdot M_{j_{1}}\left(\ldots\left(M_{j_{l}}\left(x^{(\beta)^{\dagger}} L_{i_{1}} \ldots L_{i_{k}} \vec{z}\right)\right) \ldots\right)\right) \\
& \left(\mathscr{L}^{\prime \prime}\left(N_{1}\right) \phi_{N_{1}}\right) \ldots\left(\mathscr{L}^{\prime \prime}\left(N_{m}\right) \phi_{N_{m}}\right)
\end{aligned}
$$

where

$$
\begin{cases}L_{i}=X_{\alpha_{i}}^{\beta_{i}} y_{i} & \text { for } i \in I \\ M_{i}=Z^{\sigma\left(\left[\alpha_{i}, \beta_{i}\right]\right) \rightarrow o \rightarrow o} y_{i} & \text { for } i \in J .\end{cases}
$$

For $j \in J$, it is easy to see that if a subterm (a free variable in particular) of $N_{j}$ has a type $[\gamma, \delta]$, then $\delta=\overline{\mathscr{L}(\gamma)} \in \mathcal{T}(\{\bar{o}\})$. Thus, $\phi_{N_{j}}$ substitutes linear combinators for all the free variables of $\mathscr{L}^{\prime \prime}\left(N_{j}\right)$. Therefore, $\mathscr{L}^{\prime \prime}\left(N_{j}\right) \phi_{N_{j}}$ and $M_{j}\left[\mathscr{L}^{\prime \prime}\left(N_{j}\right) \phi_{N_{j}} / y_{j}\right]$ are linear combinators. By the type $o \rightarrow o$ of $M_{j}\left[\mathscr{L}^{\prime \prime}\left(N_{j}\right) \phi_{N_{j}} / y_{j}\right]$, we see $M_{j}\left[\mathscr{L}^{\prime \prime}\left(N_{j}\right) \phi_{N_{j}} / y_{j}\right] \rightarrow_{\beta} \lambda z^{o} . z$. Therefore

$$
\begin{align*}
\mathscr{L}^{\prime \prime}(N) \phi_{N} & \rightarrow_{\beta} \lambda \vec{z} \cdot x^{(\beta)^{\dagger}}\left(X_{i_{1}} \mathscr{L}^{\prime \prime}\left(N_{i_{1}}\right) \phi_{N_{i_{1}}}\right) \ldots\left(X_{i_{k}} \mathscr{L}^{\prime \prime}\left(N_{i_{k}}\right) \phi_{N_{i_{k}}}\right) \vec{z} \\
& \rightarrow_{\eta} x^{(\beta)^{\dagger}}\left(X_{i_{1}} \mathscr{L}^{\prime \prime}\left(N_{i_{1}}\right) \phi_{N_{i_{1}}}\right) \ldots\left(X_{i_{k}} \mathscr{L}^{\prime \prime}\left(N_{i_{k}}\right) \phi_{N_{i_{k}}}\right) . \tag{3.10}
\end{align*}
$$

(3.9) and (3.10) coincide.

Case 3. Suppose that $N=\llbracket \mathrm{a}, Q \rrbracket N_{1} \ldots N_{m}$. This case can be treated by replacing the head variable $x^{[\alpha, \beta]}$ in the previous case with $\llbracket \mathrm{a}, Q \rrbracket$.

Theorem 3.16. For every affine $A C G \mathscr{G} \in \mathbf{G}^{\mathrm{aff}}(m, n)$, there is a linear $A C G$ $\mathscr{G}^{\prime \prime} \in \mathbf{G}^{\operatorname{lin}}(m, \max \{2, n\})$ such that $\mathcal{O}^{\operatorname{lin}}\left(\mathscr{G}^{\prime \prime}\right)=\left\{P \in \mathcal{O}^{\operatorname{lin}}(\mathscr{G}) \mid P\right.$ is linear $\}$.

Proof. Applying Lemma 3.15 to $N \in \mathcal{A}^{\operatorname{lin}\left(\mathscr{G}^{\prime \prime}\right) \text {, we get the equation (3.6), }}$ since $\phi_{N}$ is the empty substitution and $X_{s}^{\mathscr{L}(s)}$ is the identity.

### 3.4.3 Open Issues

In this section, we have presented how vacuous lambda abstraction can be eliminated from affine ACGs. It is natural to ask whether or not we can transform a $\lambda \mathrm{K}$-ACG $\mathscr{G}$ into a $\lambda \mathrm{I}$-ACGs $\mathscr{G}^{\prime}$ such that $\mathcal{O}^{\text {lin }}\left(\mathscr{G}^{\prime}\right)=\{P \in$
 $\lambda x^{(o \rightarrow o \rightarrow o) \rightarrow o} . x\left(\lambda z_{1}^{o} z_{2}^{o} . z_{1}\right)$ is retyped as $\lambda x^{(o \rightarrow \bar{o} \rightarrow o) \rightarrow o} . x\left(\lambda z_{1}^{o} z_{2}^{\bar{o}} . z_{1}\right)$ by our method. However, we cannot retype the non-affine term

$$
\lambda x^{(o \rightarrow o) \rightarrow o} . \mathrm{f}\left(x^{(o \rightarrow o) \rightarrow o}\left(\lambda z^{o} . \mathrm{a}\right)\right)\left(x^{(o \rightarrow o) \rightarrow o}\left(\lambda z^{o} \cdot z^{o}\right)\right)
$$

as

$$
\lambda x^{(? \rightarrow o) \rightarrow o} . \mathrm{f}\left(x^{(\bar{o} \rightarrow o) \rightarrow o}\left(\lambda z^{\bar{o}} \cdot \mathrm{a}\right)\right)\left(x^{(o \rightarrow o) \rightarrow o}\left(\lambda z^{o} \cdot z^{o}\right)\right) .
$$

The author conjectures that one can eliminate vacuous $\lambda$-abstraction from semi-affine ACGs, where a $\lambda \mathrm{K}$-term is semi-affine if for every free variable $x$ of any subterm, either

- $x$ occurs exactly once, or
- $x$ has an atomic or second-order type.

For instance, for the semi-affine term

$$
\left(\lambda u^{o \rightarrow o} v^{o \rightarrow o} z^{o} \cdot \mathbf{f}\left(u^{o \rightarrow o} z^{o}\right)\left(v^{o \rightarrow o} z^{o}\right)\right)\left(\lambda x^{o} \cdot x^{o}\right)\left(\lambda y^{o} \cdot \mathrm{a}^{o}\right),
$$

although

$$
\left(\lambda u^{o \rightarrow o} v^{\bar{o} \rightarrow o} z^{?} \cdot \mathrm{f}\left(u^{o \rightarrow o} z^{o}\right)\left(v^{\bar{o} \rightarrow o} z^{\bar{o}}\right)\right)\left(\lambda x^{o} . x^{o}\right)\left(\lambda y^{\bar{o}} \cdot \mathrm{a}^{o}\right)
$$

is not well-typed, the term

$$
\left(\lambda u^{o \rightarrow o} v^{o} z^{o} . \mathrm{f}\left(u^{o \rightarrow o} z^{o}\right) v^{o}\right)\left(\lambda x^{o} . x^{o}\right) a^{o}
$$

obtained by eliminating barred terms and types from the previous one is welltyped. If the construction based on this idea is correct, it implies that every

CFTG has a corresponding non-deleting CFTG whose IO-tree language is equivalent, because every CFTG can be encoded with respect to the IOtree language by an ACG belonging to $\mathbf{G}_{\text {tree }}^{\mathrm{K}}(2,2)$ such that only variables of atomic types occur more than once in its lexical entries, and vice versa. This also entails Fisher's result [8, 9] that every CFTG has a corresponding non-deleting CFTG whose IO-string language is equivalent.

On the other hand, the author conjectures that it is impossible to linearize $\lambda \mathrm{I}$-ACGs, i.e., there is a $\lambda \mathrm{I}$-ACG $\mathscr{G}$ such that for any linear ACG $\mathscr{G}^{\prime}$,

$$
\left\{P \in \mathcal{O}^{\operatorname{lin}}(\mathscr{G}) \mid P \text { is linear }\right\} \neq \mathcal{O}^{\operatorname{lin}}\left(\mathscr{G}^{\prime}\right) .
$$

However, it seems difficult to prove it, as we have not yet found any recursively enumerable language that cannot be generated by any linear ACG. At least, we know that the above statement is true when we consider only secondorder ACGs. The following second-order $\lambda \mathrm{I}$-ACG $\mathscr{G} \in \mathbf{G}_{\text {string }}^{\mathrm{I}}(2,2)$ generates a non-semilinear language, which cannot be generated by any second-order linear ACG by Proposition 2.20.

| $\mathrm{x} \in \mathscr{C}_{0}$ $\tau_{0}(\mathrm{x})$ $\mathscr{L}_{1}(\mathrm{x})$ <br> A $s$ $/ \mathrm{c} /$ <br> B $s \rightarrow s$ $\lambda x^{s t r} \cdot x+x$ |
| :---: |
| $\mathcal{O}^{\operatorname{lin}}(\mathscr{G})=\left\{/ \mathrm{c}^{2^{n}} / \mid n \in \mathbb{N}\right\}$ |

### 3.5 Summary

This chapter has discussed non-linear extensions of ACGs. The original ACGs have two kinds of linearity constraint: one on the lexicons, one on the abstract languages. Section 3.3 shows that relaxing the linearity constraint on the abstract language does not enlarge the class of ACLs as much as allowing non-linear lexicons.

The main issue of this chapter is comparison of generative capacity of original linear ACGs and affine ACGs, which may have vacuous lambda abstraction. In Section 3.4, we have established the equivalence between affine ACGs and linear ACGs by presenting two procedures transforming affine ACGs into linear ACGs. The first one is for second-order affine ACGs and the second one, which is an elaboration of the first one, is for third or higher-order ACGs.

The first linearization method covers the results obtained by Seki et al. [49] and Fujiyoshi [10]. Seki et al. have shown the equivalence between MCFGs and LCFRSs, and Fujiyoshi has shown the equivalence between
monadic non-duplicating CFTG and monadic linear CFTG with respect to tree languages. Their and our common basic idea is to eliminate deleting operations and some strings or subtrees in productions from the given grammar that would be erased by deleting operations during derivations. Their techniques create new productions from an original production by replacing nonterminals by new nonterminals of less ranks and fewer arguments. In this sense, their techniques work only on "second-order" settings. On the other hand, our linearization method generalizes their techniques for higherorder settings. From a type appearing in the given affine ACG, our method constructs types by eliminating subtypes of the original.

Moreover, our method entails the new result as a corollary that nonduplicating CFTGs and linear CFTGs are equivalent with respect to tree languages. This demonstrates that studying ACGs would be investigation of mildly context-sensitive grammars. Our result strengthens de Groote and Pogodalla's view that ACGs can be the kernel of a grammatical framework; not only well known grammar formalisms themselves, but also a transformation of existing grammar formalisms are encoded in ACGs.

The second linearization method can also be applied to second-order affine ACGs actually and it transforms non-duplicating CFTGs into linear CFTGs, but it does not give a conversion from MCFGs to LCFRS. The advantage of the second method is not only the fact that it is applicable to every affine ACGs, but also it preserves the original derivation structures in the sense that the original abstract language is the image of the new abstract language under a relabeling lexicon, namely $\mathscr{L}_{0}$ on page 40 . As we will see in Section 5.6, due to this property, the second linearization method is still valid for multidimensional extensions of affine ACGs.

Though every affine ACG admits linearization, our both constructions increase the size of the given grammar exponentially due to the definition of $\Pi$, so there still exists an advantage of allowing deleting operations in the ACG formalism. For instance, the atomic type $n p$ of the abstract vocabulary of the ACG in Example 1 will be divided up into three new atomic types (not counting other useless types) which correspond to noun phrases as third person singular subjects, plural subjects, and objects respectively.

## Chapter 4

## Lexicalized Abstract Categorial Grammars

### 4.1 Introduction

A formalization of the notion of lexicalized grammars is given by Schabes et al. $[47,48]$ A grammar is called lexicalized if each of its lexical entries that contribute to deriving an element of the language contains an item, called a lexical item, that explicitly appears on the surface of the derived structure. From the point of lexicalists' view, which thinks that linguistic phenomena should be accounted for by the inherent information in the lexical entries, to be lexicalized is a natural requirement. Moreover, not only from the lexicalists' point of view, but also from the point of view of the computational complexity, to be lexicalized is often thought to be desirable. At least, the decidability of the universal membership problem for lexicalized grammar is guaranteed in general.

Preceding research showed that some grammar formalisms admit lexicalization. Greibach [13] showed that every CFG has an equivalent (modulo the empty string) CFG in Greibach normal form, which can be thought of as a lexicalized grammar, because every production contains a terminal symbol in its right hand side. Schabes et al. [47, 48] presented a lexicalization method that converts finitely ambiguous TAGs into lexicalized TAGs.

This chapter is devoted to discussion of lexicalized ACGs. We say that an ACG is lexicalized iff every abstract constant is mapped to an object term containing an object constant. Although the origin of ACGs is located in a history of categorial grammars, ACGs are not necessarily lexicalized grammars by definition unlike usual categorial grammars.

In Section 4.2, which is based on Yoshinaka and Kanazawa's paper [61],
the generative capacity and complexity of lexicalized ACGs is discussed. We see that the universal membership problem for lexicalized ACG is NPcomplete. This result contrasts with the one for general ACGs and suggests that the restriction to lexicalized ACGs may be preferable from the point of view of computational complexity as well as of lexicalism.

Although the paper [61] claims that every second-order ACG admits lexicalization, the proof is in error. Kanazawa and Yoshinaka [27] give a correct proof. The main result of this chapter is a generalization of that claim. In Section 4.3, we show that semilexicalized ACGs, which form a superclass of second-order ACGs, admit lexicalization.

### 4.2 Lexicalized ACGs

Definition 4.1. For an ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$, an abstract constant a $\in \mathscr{C}_{0}$ is said to be lexical with respect to $\mathscr{L}$ if $\mathscr{L}($ a) contains a constant; otherwise, i.e., if $\mathscr{L}(\mathrm{a})$ is a combinator, a is nonlexical with respect to $\mathscr{L}$. If every abstract constant of an ACG is lexical, it is called a lexicalized $A C G$ and the class of lexicalized ACGs is denoted by $\mathbf{G}^{\text {lex }}$.

If the lexicon $\mathscr{L}$ is clear from the context, we suppress the words "with respect to $\mathscr{L}$ ".

Note that there is a gap between the notion of a lexicalized ACG in this thesis and the definition of a lexicalized TAG by Schabes et al. [47,48] While labels of internal nodes in a tree are not regarded as lexical items in Schabes et al.'s setting, labels of both internal nodes and leaf nodes, which are represented with object constants in ACGs, are regarded as lexical items by our definition. Every ACG encoding a TAG is lexicalized in our setting. This gap comes from the high degree of abstractness and generality of ACGs; they generate not only trees, but also other various types of data. Although it might be possible to define the notion of "lexicalized ACGs" so that it comprehends the notion of lexicalized TAGs by Schabes et al., such a definition would be reasonable only for tree ACGs. For the sake of generality, we adopt the above definition for the term "lexicalized ACGs".

Lemma 4.2. Let a type substitution $\sigma:\{o\} \rightarrow \mathcal{T}(\{o\})$ be defined as $\sigma(o)=$ str and a constant c have type str. For any $\alpha \in \mathcal{T}(\{o\})$, there is a closed term $Z_{\mathrm{c}}^{\sigma(\alpha)}$ of type $\sigma(\alpha)$ that contains exactly one occurrence of c .

Proof. Let $Z_{\mathrm{c}}^{\sigma(\alpha)}=Z^{\text {str } \rightarrow \sigma(\alpha)} \mathrm{c}$ for the linear combinator $Z^{\text {str } \rightarrow \sigma(\alpha)}$ defined in Definition 2.12.

Recall that the proofs of Proposition 2.15 and Corollary 2.19 use ACGs whose lexicons are just identities. Therefore, the results on the decidability of the non-emptiness problem and on the generation of a non-semilinear language hold for lexicalized ACGs. By Lemma 4.2, we can replace the ACG in the proof of Proposition 2.15 with a lexicalized string ACG, and similarly we see that there is a lexicalized string ACG in $\mathbf{G}_{\text {string }}^{\text {lex }}(3,2)$ that generates a non-semilinear string language. However, the complexity of the universal membership problem for ACGs is greatly improved in the case of lexicalized ACGs.

Proposition 4.3. The universal membership problem for lexicalized ACGs is NP-complete.

Proof. First we show that the problem is in NP. For a lexicalized ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$, if $\mathscr{L}(M) \rightarrow_{\beta} P \in \Lambda\left(\Sigma_{1}\right)$, the number of occurrences of constants in $M$ does not exceed the number of occurrences of constants in $P$. Let $P$ have $m$ occurrences of constants. $P \in \mathcal{O}(\mathscr{G})$ iff there are abstract constants $\mathrm{c}_{1}, \ldots, \mathrm{c}_{n}$ for some $n \leq m$ and a combinator $X$ of type $\tau_{0}\left(\mathrm{c}_{1}\right) \rightarrow \cdots \rightarrow \tau_{0}\left(\mathrm{c}_{n}\right) \rightarrow s$ such that $\mathscr{L}\left(X \mathrm{c}_{1} \ldots \mathrm{c}_{n}\right) \rightarrow_{\beta} P$. The size of a linear combinator $X$ is bounded by a polynomial function of the size of its type and the number of $\beta$-reduction steps needed to eliminate redexes of a linear term is bounded by its size. This shows that the question " $P \in \mathcal{O}(\mathscr{G})$ ?" is in NP.

The NP-hardness can be derived from the NP-hardness of the implicational fragment of intuitionistic linear logic $\left(\mathbf{I M L L}^{-}\right)$[28]. For a given sequent $\mathcal{S}: A_{1}, \ldots, A_{n} \Rightarrow B$ of $\mathbf{I M L L}^{-}$, we define an ACG $\mathscr{G}^{\mathcal{S}}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ where $\Sigma_{0}=\left\langle\mathscr{A}_{\mathcal{S}} \cup\{s\},\{\mathrm{a}\}, \tau_{0}\right\rangle, \mathscr{A}_{\mathcal{S}}$ is the set of atomic formulas in the sequent $\mathcal{S}, s \notin \mathscr{A}_{\mathcal{S}}, \tau_{0}(\mathrm{a})=\left(A_{1} \rightarrow \cdots \rightarrow A_{n} \rightarrow B\right) \rightarrow s, \Sigma_{1}=\langle\{o\},\{\mathrm{c}\},\{\mathrm{c} \mapsto s t r\}\rangle$, $\mathscr{L}(p)=s t r$ for all $p \in \mathscr{A}_{\mathcal{S}} \cup\{s\}$, and $\mathscr{L}(\mathrm{a})=Z_{c}^{\mathscr{L}\left(\tau_{0}(\mathrm{a})\right)}$, where $Z_{c}^{\mathscr{L}\left(\tau_{0}(\mathrm{a})\right)}$ is defined in Lemma 4.2. Then, $\mathcal{S}$ is provable in $\mathbf{I M L L}^{\circ}$ iff $\mathrm{c} \in \mathcal{O}\left(\mathscr{G}^{\mathcal{S}}\right)$.

Salvati [43] has given some general results on the generative capacity and complexity of lexicalized ACGs. Salvati has stated the above proposition in a more refined form.

Proposition 4.4 (Salvati [43]). The universal membership problem for $\mathrm{G}_{\text {tree }}^{\mathrm{lex}}(2,2)$ is $N P$-complete.

A question that naturally suggests itself at this point is whether or not there is a lexicalized ACG generating an NP-complete language. He has also answered this question. ${ }^{1}$

[^1]Proposition 4.5 (Salvati [43]). There is a lexicalized $A C G$ in $\mathbf{G}_{\text {tree }}^{\mathrm{lex}}(3,1)$ whose language is NP-complete.

Kanazawa's theorem (Theorem 2.21) holds in a weaker form for lexicalized string ACGs. We say that a class $\mathbf{L}$ of string languages is an $A F L$ if $\mathbf{L}$ is closed under union, concatenation, Kleene plus, $\varepsilon$-free homomorphism, inverse homomorphism, and intersection with regular sets.

Theorem 4.6 (Kanazawa [26]). The class of string languages generated by ACGs in $\mathbf{G} \cdot \operatorname{lex}(m, n)$ is a substitution-closed AFL for all $m, n \geq 2$.

### 4.3 Lexicalization of Semilexicalized ACGs

Definition 4.7 (Salvati [43]). An ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ is said to be semilexicalized iff for every abstract constant a $\in \mathscr{C}_{0}$, either

- a is lexical, or
- $\tau_{0}(\mathrm{a})$ is second-order.

Every lexicalized ACG and second-order ACG is semilexicalized by definition.

Salvati's result on semilexicalized ACGs shows that the classes of semilexicalized ACGs and lexicalized ACGs have similar generative capacity and complexity.

Proposition 4.8 (Salvati [43]). Every semilexicalized ACG generates a language in NP. The universal membership problem for semilexicalized ACGs is decidable.

In this section, we show that semilexicalized ACGs admit lexicalization. Our goals are the following two theorems.

Theorem 4.9. For every semilexicalized $A C G \mathscr{G} \in \mathbf{G}(m, n)$ with $m \in\{2,3\}$, there are lexicalized $A C G s \mathscr{G}^{\prime} \in \mathbf{G}(m, n+1)$ and $\mathscr{G}^{\prime \prime} \in \mathbf{G}(m+1, n)$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\mathcal{O}\left(\mathscr{G}^{\prime \prime}\right)=\{R \in \mathcal{O}(\mathscr{G}) \mid R \text { contains a constant }\} .
$$

Theorem 4.10. For every semilexicalized $A C G \mathscr{G} \in \mathbf{G}(m, n)$ with $m \notin$ $\{2,3\}$, there is a lexicalized $A C G \mathscr{G}^{\prime} \in \mathbf{G}(m, n)$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\{R \in \mathcal{O}(\mathscr{G}) \mid R \text { contains a constant }\} .
$$

The above theorem for $m=1$ is trivial. Each ACG in $\mathbf{G}(1, n)$ defines just a finite language. Lexicalization of a first-order ACG is just eliminating all the nonlexical constants from the abstract vocabulary.

Note that the notion of lexicalization in this thesis differs from Schabes et al.'s one $[47,48]$, which is concerned with grammars that generate strings through tree structures. They have defined the notion of lexicalization so that it preserves not only the string language but also the tree language yielding the string language. This definition requires grammars that should be lexicalized to be finitely ambiguous, i.e., there is no string in the language that has infinitely many trees yielding it in the tree language. Our lexicalization method converts any semilexicalized ACG, which may be infinitely ambiguous in the sense that there are infinitely many $\lambda$-terms in the abstract language that are mapped to the same object term modulo $\beta \eta$, into a lexicalized ACG, where the original abstract language may be lost.

### 4.3.1 Basic Idea

This subsection explains our basic strategy for lexicalizing semilexicalized ACGs. ${ }^{2}$

In the sequel, for an ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$, let us denote the set of lexical constants by $\mathscr{C}_{0}^{+}$and the set of nonlexical constants by $\mathscr{C}_{0}^{-}$. We denote the lexical part of the abstract vocabulary by $\Sigma_{0}^{+}=\left\langle\mathscr{A}_{0}, \mathscr{C}_{0}^{+}, \tau_{0}\right\rangle$, and the nonlexical part by $\Sigma_{0}^{-}=\left\langle\mathscr{A}_{0}, \mathscr{C}_{0}^{-}, \tau_{0}\right\rangle$. An ACG is semilexicalized iff $\Sigma_{0}^{-}$is second-order. For $M, N \in \Lambda\left(\Sigma_{0}\right)$, we write $M \stackrel{\mathscr{L}}{\approx} N$ (or $M \approx N$ if $\mathscr{L}$ is clear from the context) iff $\tau_{0}(M)=\tau_{0}(N)$ and $\mathscr{L}(M)=\mathscr{L}(N)$. For two sets $\Gamma$ and $\Delta$ of terms, we write $\Gamma \approx \Delta$ iff for every $M \in \Gamma$, there is $N \in \Delta$ with $M \approx N$ and vice versa.

For eliminating nonlexical constants from a semilexicalized ACG $\mathscr{G}=$ $\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$, we construct a new ACG $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{0}, \mathscr{L}^{\prime}, s^{\prime}\right\rangle$ such that

- for every $\mathrm{A} \in \mathscr{C}_{0}^{\prime}, \mathscr{L}^{\prime}(\mathrm{A})$ contains at least one lexical (w.r.t. $\mathscr{L}$ ) constant $a \in \mathscr{C}_{0}^{+}$.
- $\mathcal{O}\left(\mathscr{G}^{\prime}\right) \stackrel{\mathscr{L}}{\approx} \mathcal{A}(\mathscr{G})$.

Then, $\mathscr{G}^{\prime \prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L} \circ \mathscr{L}^{\prime}, s^{\prime}\right\rangle$ is a lexicalized ACG by the first condition and we see that $\mathcal{O}\left(\mathscr{G}^{\prime \prime}\right)=\mathcal{O}(\mathscr{G})$ by the second condition.

[^2]For defining $\Sigma_{0}^{\prime}$ and $\mathscr{L}^{\prime}$, in fact it is enough to find an integer $k_{\mathscr{G}} \in \mathbb{N}$ such that for every $M \in \mathcal{A}(\mathscr{G})$, there is $N \in \mathcal{A}(\mathscr{G})$ with $N \approx M$ and

$$
\begin{equation*}
\#_{\mathscr{C}_{0}^{-}}(N) \leq k_{\mathscr{G}} \cdot \#_{\mathscr{C}_{0}^{+}}(N) \tag{4.1}
\end{equation*}
$$

Suppose that such an integer $k_{\mathscr{G}}$ exists. Define $\Sigma_{0}^{\prime}$ and $\mathscr{L}^{\prime}: \Sigma_{0}^{\prime} \rightarrow \Sigma_{0}$ by

$$
\begin{aligned}
\mathscr{A}_{0}^{\prime} & =\mathscr{A}_{0}, \\
\mathscr{C}_{0}^{\prime} & =\left\{\llbracket \mathrm{a}, \mathrm{c}_{1}, \ldots, \mathrm{c}_{k} \rrbracket \mid 0 \leq k \leq k \mathscr{G}, \mathrm{a} \in \mathscr{C}_{0}^{+}, \mathrm{c}_{i} \in \mathscr{C}_{0}^{-}\right\} \\
\tau_{0}^{\prime}\left(\llbracket \mathrm{a}, \mathrm{c}_{1}, \ldots, \mathrm{c}_{k} \rrbracket\right) & =\left(\tau_{0}(\mathrm{a}) \rightarrow \tau_{0}\left(\mathrm{c}_{1}\right) \rightarrow \cdots \rightarrow \tau_{0}\left(\mathrm{c}_{k}\right) \rightarrow s\right) \rightarrow s \\
\mathscr{L}^{\prime}(p) & =p \text { for } p \in \mathscr{A}_{0}, \\
\mathscr{L}^{\prime}\left(\llbracket \mathrm{a}, \mathrm{c}_{1}, \ldots, \mathrm{c}_{k} \rrbracket\right) & =\lambda w . w \mathrm{ac}_{1} \ldots \mathrm{c}_{k} .
\end{aligned}
$$

Then, for $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L} \circ \mathscr{L}^{\prime}, s\right\rangle$, we see that $\mathcal{O}(\mathscr{G})=\mathcal{O}\left(\mathscr{G}^{\prime}\right)$. The inclusion $\mathcal{O}\left(\mathscr{G}^{\prime}\right) \subseteq \mathcal{O}(\mathscr{G})$ is trivial, since for every $P \in \mathcal{A}\left(\mathscr{G}^{\prime}\right)$ we have $\left|\mathscr{L}^{\prime}(P)\right|_{\beta \eta} \in$ $\mathcal{A}(\mathscr{G})$. We show the converse. For $M \in \mathcal{A}(\mathscr{G})$, by the hypothesis we can assume that

$$
\#_{\mathscr{C}_{0}^{-}}(M) \leq k_{\mathscr{G}} \cdot \#_{\mathscr{C}_{0}^{+}}(M)
$$

Let $M^{\prime}$ be obtained from $M$ by replacing each occurrence of a constant by a fresh variable so that

$$
M=M^{\prime}\left[\mathrm{a}_{1} / x_{1}, \ldots, \mathrm{a}_{m} / x_{m}, \mathrm{c}_{1} / y_{1}, \ldots, \mathrm{c}_{n} / y_{n}\right]
$$

where each $\mathrm{a}_{i}$ is lexical and $\mathrm{c}_{i}$ is nonlexical. Since $n \leq k_{\mathscr{G}} \cdot m$, we can partition the set $\{1, \ldots, n\}$ into $m$ (possibly empty) subsets $I_{1}, \ldots, I_{m}$ such that $\left|I_{i}\right| \leq k_{\mathscr{G}}$. Let

$$
P=\llbracket \mathrm{a}_{1}, \overrightarrow{\mathrm{c}}_{I_{1}} \rrbracket\left(\lambda x_{1} \vec{y}_{I_{1}} \cdot \llbracket \mathrm{a}_{2}, \overrightarrow{\mathrm{c}}_{I_{2}} \rrbracket\left(\lambda x_{2} \vec{y}_{I_{2}} \ldots . \llbracket \mathrm{a}_{m}, \overrightarrow{\mathrm{c}}_{I_{m}} \rrbracket\left(\lambda x_{m} \vec{y}_{I_{m}} \cdot M^{\prime}\right) \ldots\right)\right)
$$

where $\vec{y}_{I_{i}}$ is a sequence of variables $y_{j}$ with $j \in I_{i}$ and $\vec{c}_{I_{i}}$ is the corresponding sequence of $\mathrm{c}_{j}$ with $j \in I_{i}$. Then we have $P \in \mathcal{A}\left(\mathscr{G}^{\prime}\right), \mathscr{L}^{\prime}(P)=M$ and thus $\mathcal{O}\left(\mathscr{G}^{\prime}\right) \subseteq \mathcal{O}(\mathscr{G})$. Note that the above construction increases the order of the abstract vocabulary by 2 . Although this construction will have to be modified, explaining this basic idea should help the reader to understand our lexicalization method.

How can we find that key integer $k_{\mathscr{G}}$ ? In fact, if the given ACG $\mathscr{G}$ contains no nullary or unary nonlexical constants, then $k_{g}$ is given as

$$
\begin{equation*}
k_{\mathscr{G}}=\max \left\{\left|\tau_{0}(\mathrm{a})\right| \mid \mathrm{a} \in \mathscr{C}_{0}^{+}\right\} . \tag{4.2}
\end{equation*}
$$

Our algorithm first transforms the given semilexicalized ACG into an equivalent one that satisfies the assumption that every nonlexical constant is neither
nullary nor unary, but, we postpone the discussion of this transformation until later. We here explain why the number given by (4.2) is enough under the assumption. First we introduce the notion of a type occurrence in a term, by explicitly displaying the type of every subterm occurring in a term, like we do with variables. For instance, if $M$ has type $\gamma \rightarrow \delta$ and $N$ has type $\gamma$, then we write $\left(M^{\gamma \rightarrow \delta} N^{\gamma}\right)^{\delta}$ instead of $M N$. If an atomic type $p$ occurs in $\gamma$, we have two corresponding occurrences of $p$ in $\left(M^{\gamma \rightarrow \delta} N^{\gamma}\right)^{\delta}$, one in the type $\gamma \rightarrow \delta$ of $M$ and one in the type $\gamma$ of $N$. If an atomic type $q$ occurs in $\delta$, we have two corresponding occurrences of $q$, one in the type $\gamma \rightarrow \delta$ of $M$ and one in the type $\delta$ of $M N$. We then draw links connecting corresponding atomic type occurrences as

$$
\left(M^{\gamma[p] \rightarrow \delta[q]} N^{\gamma[p]}\right)^{\delta[q]}
$$

where $\gamma[p]$ specifies an occurrence of the type $p$ in $\gamma$ and likewise for $\delta[q]$. For an abstraction term $\left(\lambda x^{\gamma} . M^{\delta}\right)^{\gamma \rightarrow \delta}$, since $M$ has exactly one occurrence of the free variable $x^{\gamma}$ by the linearity, we can give links as follows

$$
\left(\lambda x^{\gamma} \cdot M^{\delta[q]}\left[x^{\gamma[p]}\right]\right)^{\gamma[p]} \rightarrow \delta[q]
$$

where $M[x]$ specifies the occurrence of the variable $x$ in $M$. Though the above two examples illustrate links of length just one, links form longer paths in a term. Consequently, for instance, the occurrence of $p$ in the type of $c$ is linked to the one in the type of $y$ in

$$
y^{o \rightarrow(p \rightarrow q) \rightarrow r} a^{o}\left(\lambda x^{p} \cdot \mathrm{c}^{\beta \rightarrow q} x^{p}\right)^{p \rightarrow q} .
$$

Here we suppress the types of some subterms for conciseness. Actually the path illustrated above passes through the occurrence of $p$ in $(y a)^{(p \rightarrow q) \rightarrow r}$. Note that in a linear term $M$, no path branches and every path has exactly two end-points.

For conciseness, if the intended occurrence is clear, by "term" we mean an appropriate occurrence of that term, and "type" we mean an appropriate occurrence of that type.

Suppose that a $\beta$-normal term $M$ is in $\mathcal{A}(\mathscr{G})$. To compare the numbers of occurrences of lexical and nonlexical constants, in other words, to make sure that $\#_{\mathscr{C}_{0}^{-}}(M) \leq k_{\mathscr{G}} \cdot \#_{\mathscr{C}_{0}^{+}}(M)$ for $k_{\mathscr{G}}$ defined as (4.2), we divide $M$ into one lexical part $M_{0}$ and several nonlexical parts $M_{1}, \ldots, M_{m}$ so that

1. $M=\left(\lambda x_{1} \ldots x_{m} \cdot M_{0}\right) M_{1} \ldots M_{m}$,
2. $M_{0} \in \Lambda\left(\Sigma_{0}^{+}\right)$and $M_{0}$ is in long normal form,
3. $M_{i} \in \Lambda\left(\Sigma_{0}^{-}\right)$and $M_{i}$ is in long normal form for $i \geq 1$,
4. for each $i \geq 1, M_{i}$ has the form $M_{i}=\lambda \vec{z}_{i} \cdot M_{i}^{\prime}$, where $\vec{z}_{i}$ consists of variables of atomic types, and $M_{i}^{\prime}$ also has an atomic type and contains no variables other than $\vec{z}_{i}$,
5. $m$ is as small as possible.

The last condition forbids $M_{0}$ to have a subterm of the form $x_{i} \vec{N}_{1}\left(x_{j} \vec{N}_{2}\right)$. If $M_{0}$ has such a subterm, we can decrease the number $m$ of nonlexical parts by replacing the subterm with $x_{i, j} \vec{N}_{1} \vec{N}_{2}$ and combining two nonlexical parts $M_{i}$ and $M_{j}$ as follows:

$$
M_{i, j}=\lambda \vec{z}_{1} \vec{z}_{2} \cdot M_{i} \vec{z}_{1}\left(M_{j} \vec{z}_{2}\right),
$$

where $\left|\vec{z}_{1}\right|=\left|\vec{N}_{1}\right|$ and $\left|\vec{z}_{2}\right|=\left|\vec{N}_{2}\right|$. Of course each $M_{i}$ cannot be $\lambda z . z$. For instance, if

$$
M \equiv \mathrm{~b}\left(\lambda y_{1}^{s} y_{2}^{s} \cdot \mathrm{c}\left(\mathrm{ca}\left(\mathrm{~b}\left(\lambda y_{0}^{s} \cdot y_{0}\right)\right)\right)\left(\mathrm{b}\left(\lambda y_{3}^{s} y_{4}^{s} \cdot \mathrm{c}\left(\mathrm{c} y_{1} y_{2}\right)\left(\mathrm{c} y_{3} y_{4}\right)\right)\right),\right.
$$

where $\mathrm{a}, \mathrm{b} \in \mathscr{C}_{0}^{+}, \mathrm{c} \in \mathscr{C}_{0}^{-}, \tau_{0}(\mathrm{a})=s, \tau_{0}(\mathrm{~b})=(s \rightarrow s) \rightarrow s$, and $\tau_{0}(\mathrm{c})=s^{2} \rightarrow$ $s$, we let

$$
\begin{aligned}
& M_{0} \equiv \mathrm{~b}\left(\lambda y_{1}^{s} y_{2}^{s} \cdot x_{1}^{s^{3} \rightarrow s} \mathrm{a}\left(\mathrm{~b}\left(\lambda y_{0}^{s} \cdot y_{0}\right)\right)\left(\mathrm{b}\left(\lambda y_{3}^{s} y_{4}^{s} \cdot x_{2}^{s^{4} \rightarrow s} y_{1} y_{2} y_{3} y_{4}\right)\right)\right) \\
& M_{1} \equiv \lambda z_{1,1}^{s} z_{1,2}^{s} z_{1,3}^{s} \cdot \mathrm{c}\left(\mathrm{c} z_{1,1} z_{1,2}\right) z_{1,3} \\
& M_{2}
\end{aligned}
$$

Suppose that a nonlexical part $M_{i}$ with $i \geq 1$ has the form $M_{i}=$ $\lambda z_{1}^{p_{1}} \ldots . z_{n}^{p_{n}} . M_{i}^{\prime}$ with $\tau_{0}\left(M_{i}\right)=p_{1} \rightarrow \cdots \rightarrow p_{n} \rightarrow q . \quad M_{i}^{\prime}$ can be regarded as a tree over $\mathscr{C}_{0}^{-} \cup\left\{z_{1}, \ldots, z_{n}\right\}$. Since $M$ contains no nullary or unary nonlexical constants, every constant in $M_{i}^{\prime}$ has at least two arguments. If a tree contains no node of rank 1 , then the number of leaves is larger than the number of internal nodes. In $M_{i}$, every leaf node is a variable and every internal node is a nonlexical constant. Thus, we have

$$
\begin{equation*}
1 \leq \#_{\mathscr{C}_{0}^{-}}\left(M_{i}\right)<n=\left|\tau_{0}\left(M_{i}\right)\right|-1 . \tag{4.3}
\end{equation*}
$$

On the other hand, the type of $x_{i}$ in $M_{0}$ is $\tau_{0}\left(M_{i}\right)=p_{1} \rightarrow \cdots \rightarrow p_{n} \rightarrow q$. Now we are going to show that

$$
\left\{\begin{array}{l}
\text { the path from each } p_{j} \text { in the type of } x_{i}  \tag{4.4}\\
\text { ends in the type of a lexical constant in } M_{0}
\end{array}\right.
$$

Since $M_{0}$ is in long normal form, the occurrence of $x_{i}$ in $M_{0}$ has exactly $n$ arguments $N_{1}, \ldots, N_{n}$. Since the type $p_{j}$ of $N_{j}$ is atomic, $N_{j}$ is not an
abstraction term. By the fifth condition, the head of $N_{j}$ cannot be one of the free variables $x_{1}, \ldots, x_{m}$ of $M_{0}$. If the head of $N_{j}$ is a lexical constant, (4.4) holds. Otherwise, it is a bound variable $w_{0}^{\vec{\alpha}_{0} \rightarrow p_{j}}$, where we find a path

$$
x_{i}^{p_{1} \rightarrow \cdots \rightarrow p_{j} \rightarrow \cdots \rightarrow p_{n} \rightarrow q} N_{1} \ldots N_{j-1}\left(w_{0}^{\vec{\alpha}_{0} \rightarrow p_{j}} \vec{N}\right) N_{j+1} \ldots N_{n} .
$$

The path goes to the type of the abstraction term $L_{0}^{\left(\vec{\alpha}_{0} \rightarrow p_{j}\right) \rightarrow \beta_{0}}$ that binds the variable $w_{0}$,
$L_{0}=\left(\lambda w_{0}^{{\overrightarrow{\alpha_{0}} \rightarrow p_{j}}^{j}} L\left[x_{i}^{p_{1} \rightarrow \cdots \rightarrow p_{j} \rightarrow \ldots \rightarrow p_{n} \rightarrow q} N_{1} \ldots N_{i-1}\left(w_{0}^{\left.\left.\left.{\overrightarrow{\alpha_{0}} \rightarrow p_{j}}^{N}\right) N_{i+1} \ldots N_{n} / z\right]\right)^{\left(\vec{\alpha}_{0} \rightarrow p_{j}\right) \rightarrow \beta_{0}}}\right.\right.\right.$
Since $M_{0}$ has the distinguished type $s, L_{0}$ of type $\left(\vec{\alpha}_{0} \rightarrow p_{j}\right) \rightarrow \beta_{0}$ cannot be $M_{0}$. Since $M_{0}$ is $\beta$-normal, $L_{0}$ cannot have an argument. Thus we can find the smallest application term $K_{0}$ which has $L_{0}$ as a proper subterm. $K_{0}$ has the form

$$
K_{0}=x^{\vec{\gamma} \rightarrow\left(\vec{\beta} \rightarrow\left(\vec{\alpha}_{0} \rightarrow p_{j}\right) \rightarrow \beta_{0}\right) \rightarrow \delta_{0}} \vec{L}_{1}^{\vec{\gamma}}\left(\lambda \vec{v}^{\vec{\beta}} \cdot L_{0}^{\left(\vec{\alpha}_{0} \rightarrow p_{j}\right) \rightarrow \beta_{0}}\right) .
$$

If x is a lexical constant, (4.4) holds. Otherwise, since all free variables $x_{1}, \ldots, x_{m}$ of $M_{0}$ have a second-order type, x must be a bound variable. Let $\mathrm{x}=w_{1}^{\vec{\gamma} \rightarrow\left(\vec{\beta} \rightarrow\left(\vec{\alpha}_{0} \rightarrow p_{j}\right) \rightarrow \beta_{0}\right) \rightarrow \delta_{0}}$ and find the unique abstraction term that binds the occurrence of $w_{1}$. This way we can trace the path from the occurrence of $p_{j}$ in the type of $x_{i}$. If the path passes through the types of variables $x_{i}, w_{0}, w_{1}, \ldots$ in this order, then the type of $w_{l+1}$ is more complicated than that of $x_{l}$, i.e., $\operatorname{ord}\left(\tau_{0}\left(w_{l+1}\right)\right) \geq \operatorname{ord}\left(\tau_{0}\left(w_{l}\right)\right)+2$. Since $M_{0}$ is finite, eventually we see that the path ends in a type occurrence of a lexical constant. We have proved the statement (4.4).

Suppose that the path from $p_{j}$ in the type of $x_{i}$ ends in the type $\alpha\left[p_{j}\right]$ of a lexical constant a. When conversely we start tracing the path from the occurrence $p_{j}$ in $\alpha\left[p_{j}\right]=\tau_{0}(\mathrm{a})$, it cannot end in any other place than the occurrence $p_{j}$ in the type $p_{1} \rightarrow \cdots \rightarrow p_{n} \rightarrow q$ of $x_{i}$. Therefore,

$$
\sum_{1 \leq i \leq m}\left(\left|\tau_{0}\left(x_{i}\right)\right|-1\right) \leq \sum_{\mathbf{a} \in \mathscr{C}_{0}^{+}} \#_{\mathrm{a}}(M)\left|\tau_{0}(\mathrm{a})\right| .
$$

By (4.3) and the definition (4.2) of the key integer $k_{\mathscr{G}}$, we have

$$
\begin{aligned}
& \#_{\mathscr{C}_{0}^{-}}(M)=\sum_{1 \leq i \leq m} \#_{\mathscr{C}_{0}^{-}}\left(M_{i}\right)<\sum_{1 \leq i \leq m}\left(\left|\tau_{0}\left(M_{i}\right)\right|-1\right)=\sum_{1 \leq i \leq m}\left(\left|\tau_{0}\left(x_{i}\right)\right|-1\right) \\
\leq & \sum_{\mathrm{a} \in \mathscr{C}_{0}^{+}} \#_{\mathrm{a}}(M)\left|\tau_{0}(\mathrm{a})\right| \leq \#_{\mathscr{C}_{0}^{+}}(M) \cdot \max \left\{\left|\tau_{0}(\mathrm{a})\right| \mid \mathrm{a} \in \mathscr{C}_{0}^{+}\right\}=k_{\mathscr{G}} \cdot \#_{\mathscr{C}_{0}^{+}}(M) .
\end{aligned}
$$

Remark 4.11. We note some properties of a path here.

- In $\times M_{1} \ldots M_{n}$ where $\times$ is an atomic term, if a path starts from inside $M_{i}$, then it cannot go into any other $M_{j}$ with $j \neq i$.
- If both two end-points of a path are inside the types of some constants, then one point is at a positive occurrence of the atomic type, and the other is at a negative occurrence of the atomic type.


### 4.3.2 First Step

Our actual lexicalization algorithm consists of three steps: firstly we convert $\mathscr{G}$ into an ACG satisfying the conditions in Definition 4.12, secondly we eliminate nullary and unary nonlexical constants, and finally all the nonlexical constants are eliminated. This section is devoted to the first step. The treatment of remaining steps varies depending on the order of the given semilexicalized ACG.

The conditions in Definition 4.12 talk about the nonlexical part of the abstract vocabulary.

Definition 4.12 (Condition I \& II). A semilexicalized ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}\right.$, $\mathscr{L}, s\rangle$ satisfies Condition $I$ if for every second-order closed term $M \in \Lambda\left(\Sigma_{0}^{-}\right)$, there is $N \approx M$ such that
i. if $\tau_{0}(M)$ is nullary, then $N \in \mathscr{C}_{0}^{-}$,
ii. if $\tau_{0}(M)$ is not nullary, then $N$ contains no nullary nonlexical constants.

A semilexicalized ACG $\mathscr{G}$ satisfies Condition II, if for every second-order closed term $M \in \Lambda\left(\Sigma_{0}^{-}\right)$, there is a $\beta$-normal term $N$ such that $N \approx M$ and
i. if $\tau_{0}(M)$ is unary, then $N \in \mathscr{C}_{0}^{-}$,
ii. if $N$ contains a unary constant $\mathrm{c} \in \mathscr{C}_{0}^{-}$, then the argument of c is a bound variable (if any).

Lemma 4.13. Suppose that a semilexicalized $A C G \mathscr{G}$ satisfies Condition I. Then for every $M \in \Lambda\left(\Sigma_{0}\right)$, there is a $\beta$-normal term $N \in \Lambda\left(\Sigma_{0}\right)$ such that $N \approx M$ and if a nullary nonlexical constant c occurs in $N$, then either

- $N=\mathrm{c}$,
- coccurs as an argument of a lexical constant, or
- c occurs as an argument of a variable.

Proof. Let us divide $M$ into one lexical part $M_{0}$ and several nonlexical parts $M_{1}, \ldots, M_{m}$ so that

1. $M=\left(\lambda x_{1} \ldots x_{m} \cdot M_{0}\right) M_{1} \ldots M_{m}$,
2. $M_{0} \in \Lambda\left(\Sigma_{0}^{+}\right)$and $M_{0}$ is in long normal form,
3. $M_{i} \in \Lambda\left(\Sigma_{0}^{-}\right)$and $M_{i}$ is in long normal form for $i \geq 1$,
4. for each $i \geq 1, M_{i}$ has the form $M_{i}=\lambda \vec{z}_{i} \cdot M_{i}^{\prime}$, where $\vec{z}_{i}$ consists of variables of atomic types, and $M_{i}^{\prime}$ also has an atomic type and contains no variables other than $\vec{z}_{i}$,
5. $m$ is as small as possible.

By Condition I, there are $N_{1}, \ldots, N_{m}$ such that $N_{i} \approx M_{i}$ and either $N_{i}$ is a nullary nonlexical constant or $N_{i}$ contains no nullary nonlexical constants. Let $N=\left|M_{0}\left[N_{1} / x_{1}, \ldots, N_{m} / x_{m}\right]\right|_{\beta}$ and suppose that $N_{j}$ is a nullary nonlexical constant c. If $M_{0}=x_{j}$, then $N=N_{j}=$ c. If $M_{0} \neq x_{j}$, there is $M_{0}^{\prime}$ and an atomic term $\times$ such that $M_{0}=M_{0}^{\prime}\left[\times \vec{L}_{1} x_{j} / z\right]$, since $x_{j}$ has an atomic type. Since $M_{0}$ has no subterm of the form $x_{i} \vec{N}_{1}\left(x_{j} \vec{N}_{2}\right), \mathrm{x}$ is either a lexical constant or a variable other than $x_{1}, \ldots, x_{m}$. In $N, N_{j}$ appears as an argument of $x$.

Similarly we have the following lemma.
Lemma 4.14. Suppose that a semilexicalized $A C G \mathscr{G}$ satisfies Condition II. Then for every $M \in \Lambda\left(\Sigma_{0}\right)$, there is a $\beta$-normal term $N \in \Lambda\left(\Sigma_{0}\right)$ such that $N \approx M$ and if a unary nonlexical constant coccurs in a subterm c $N^{\prime}$ of $N$, then the head of $N^{\prime}$ is either a lexical constant or a variable.

Note that Condition II-i ensures that for every atomic type $p$, there is a unary nonlexical constant $\mathrm{e}_{p} \in \mathscr{C}_{0}^{-}$of type $p \rightarrow p$ which is mapped to the identity $\lambda x^{\mathscr{L}(p)} . x$.

First we present the procedure for converting a semilexicalized ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ into an equivalent ACG $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L}^{\prime}, s\right\rangle$ satisfying Condition I in Definition 4.12. We define the new abstract vocabulary $\Sigma_{0}^{\prime}$ and lexicon $\mathscr{L}^{\prime}$ as extensions of $\Sigma_{0}$ and $\mathscr{L}$ by adding new nonlexical constants. This procedure is performed on the nonlexical part $\Sigma_{0}^{-}$of the abstract vocabulary and we ignore the lexical part $\Sigma_{0}^{+}$.

Let $\Gamma_{\mathscr{G}} \subseteq \mathscr{A}_{0} \times\left\{P \in \Lambda\left(\Sigma_{1}\right) \mid P\right.$ is a combinator $\}$ be defined as follows:

$$
\begin{aligned}
\Gamma_{0} & =\varnothing \\
\Gamma_{n+1} & =\Gamma_{n} \cup\left\{\left.\langle q,| \mathscr{L}(\mathrm{c}) P_{1} \ldots P_{m}\right|_{\beta \eta}\right\rangle \mid \\
& \left.\quad \mathrm{c} \in \mathscr{C}_{0}^{-}, \tau_{0}(\mathrm{c})=p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow q,\left\langle p_{i}, P_{i}\right\rangle \in \Gamma_{n}\right\}, \\
\Gamma_{\mathscr{G}} & =\bigcup_{n \geq 1} \Gamma_{n} .
\end{aligned}
$$

Recall that for a fixed type, only finitely many linear combinators of that type exist (modulo $\beta \eta$ ). Thus, $\Gamma_{\mathscr{G}}$ is finite, because $\tau_{1}(P)=\mathscr{L}(p)$ for every element $\langle p, P\rangle \in \Gamma_{\mathscr{G}}$. Obviously each $\Gamma_{n}$ can be computed effectively. If $\Gamma_{n}=\Gamma_{n+1}$, then $\Gamma_{n}=\Gamma_{n+k}$ for all $k \geq 0$ and thus $\Gamma_{\mathscr{G}}=\Gamma_{n}$. Therefore, $\Gamma_{\mathscr{G}}$ can be computed effectively. Moreover, immediately by the construction of $\Gamma_{\mathscr{G}}$, for every variable-free term $M \in \Lambda\left(\Sigma_{0}^{-}\right)$of an atomic type, we have $\left.\left.\left\langle\tau_{0}(M),\right| \mathscr{L}(M)\right|_{\beta \eta}\right\rangle \in \Gamma_{\mathscr{G}}$.

To satisfy Condition I-i, for each $\langle p, P\rangle \in \Gamma_{\mathscr{G}}$, let us put the following lexical entries ${ }^{3}$ into $\mathscr{C}_{0}^{\prime}$

$$
\langle\llbracket p, P \rrbracket, p, P\rangle .
$$

Next, for Condition I-ii, we add the following lexical entries

$$
\left.\langle\llbracket \vec{p} \rightarrow q,| \lambda \vec{x} .\left.\mathscr{L}(\mathrm{c}) P_{1} \ldots P_{m}\right|_{\beta \eta} \rrbracket, \vec{p} \rightarrow q, \lambda \vec{x} \cdot \mathscr{L}(\mathrm{c}) P_{1} \ldots P_{m}\right\rangle
$$

where

- c $\in \mathscr{C}_{0}^{-}$,
- $\tau_{0}(\mathrm{c})=p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow q$,
- either $\left\langle p_{i}, P_{i}\right\rangle \in \Gamma_{\mathscr{G}}$ or $P_{i}=x_{i}$,
- $\vec{p}$ and $\vec{x}$ are the subsequences of $p_{1} \ldots p_{m}$ and $x_{1} \ldots x_{m}$, respectively, such that $p_{i}$ is in $\vec{p}$ iff $x_{i}$ is in $\vec{x}$ iff $P_{i}=x_{i}$.

We obtain the following lemma.
Lemma 4.15 (Condition I). Every semilexicalized $A C G \mathscr{G} \in \mathbf{G}(m, n)$ has an equivalent semilexicalized $A C G \mathscr{G}^{\prime} \in \mathbf{G}(m, n)$ that satisfies Condition I in Definition 4.12.

[^3]Proof. Let $\mathscr{G}^{\prime}$ be the ACG obtained from an ACG $\mathscr{G}$ by the above procedure. Since $\mathscr{G}^{\prime}$ is an extension of $\mathscr{G}$, clearly $\mathcal{O}(\mathscr{G}) \subseteq \mathcal{O}\left(\mathscr{G}^{\prime}\right)$ holds. Moreover since the role of a new constant can be played by finitely many nonlexical constants of $\mathscr{G}$, the converse $\mathcal{O}\left(\mathscr{G}^{\prime}\right) \subseteq \mathcal{O}(\mathscr{G})$ holds. It is clear that for every $p \in \mathscr{A}_{0}$ and $P \in \Lambda\left(\Sigma_{1}\right)$ the following three are equivalent:

- there is a variable-free term $L \in \Lambda\left(\Sigma_{0}^{-}\right)$of type $p$ with $|\mathscr{L}(L)|_{\beta \eta} \equiv P$,
- $\langle p, P\rangle \in \Gamma_{\mathscr{G}}$,
- there is a nullary nonlexical constant of type $p$ in $\mathscr{C}_{0}^{\prime-}$ that is mapped to $P$.

We show that $\mathscr{G}^{\prime}$ satisfies Condition I. Since for every $M \in \Lambda\left(\Sigma_{0}^{\prime-}\right)$, there is $L \in \Lambda\left(\Sigma_{0}^{-}\right)$such that $\tau_{0}(L)=\tau_{0}^{\prime}(M)$ and $\mathscr{L}(L)=\mathscr{L}^{\prime}(M)$, it is enough to show that for every closed second-order term $L \in \Lambda\left(\Sigma_{0}^{-}\right)$, there is $N \in \Lambda\left(\Sigma_{0}^{\prime-}\right)$ such that

- $\tau_{0}(L)=\tau_{0}^{\prime}(N)$ and $\mathscr{L}(L)=\mathscr{L}^{\prime}(N)$,
i. if $\tau_{0}(L)$ is nullary, then $N \in \mathscr{C}_{0}^{\prime-}$,
ii. if $\tau_{0}(L)$ is not nullary, then $N$ contains no nullary nonlexical constants.

Condition I-i is satisfied by the above equivalence relation. We show that Condition I-ii is satisfied by induction on $\#_{\mathscr{C}_{0}^{-}}(L)$. Suppose that a closed second-order term $L \in \Lambda\left(\Sigma_{0}^{-}\right)$has the form

$$
L=\lambda \vec{y} \cdot c L_{1} \ldots L_{m}
$$

with $\vec{y} \neq \varepsilon$ for a nonlexical constant c of type $p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow q$. Let $\vec{y}_{i}$ be the subsequence of $\vec{y}$ whose elements appear in $L_{i}$ and $I=\left\{i \mid \vec{y}_{i} \neq \varepsilon\right\}$. Note that $I \neq \varnothing$ by $\vec{y} \neq \varepsilon$. By the induction hypothesis and Condition I-i, for each $i \in\{1, \ldots, m\}$ we have $N_{i} \in \Lambda\left(\Sigma_{0}^{\prime-}\right)$ such that

- $\tau_{0}^{\prime}\left(N_{i}\right)=\tau_{0}\left(\lambda \vec{y}_{i} . L_{i}\right)$ and $\mathscr{L}^{\prime}\left(N_{i}\right)=\mathscr{L}\left(\lambda \vec{y}_{i} . L_{i}\right)$
- $N_{i} \in \mathscr{C}_{0}^{\prime-}$ (and thus $\left.\left.\left.\left\langle p_{i},\right| \mathscr{L}^{\prime}\left(N_{i}\right)\right|_{\beta_{\eta}}\right\rangle \in \Gamma_{\mathscr{G}}\right)$ if $i \notin I$,
- $N_{i}$ contains no nullary constants if $i \in I$,

By the definition, we have the constant $\llbracket \vec{p}_{I} \rightarrow q, P \rrbracket$ where

- $P \equiv\left|\lambda \vec{x}_{I} \cdot \mathscr{L}(\mathrm{c}) P_{1} \ldots P_{m}\right|_{\beta \eta}$,
- $P_{i}=\mathscr{L}^{\prime}\left(N_{i}\right)$ for $i \notin I$,
- $P_{i}=x_{i}$ for $i \in I$,
- $\vec{p}_{I}$ and $\vec{x}_{I}$ are the subsequences of $p_{1} \ldots p_{m}$ and $x_{1} \ldots x_{m}$ respectively such that $p_{i}$ is in $\vec{p}_{I}$ iff $x_{i}$ is in $\vec{x}_{I}$ iff $i \in I$.

Thus, for

$$
N=\lambda \vec{y} \cdot \llbracket \vec{p} \rightarrow q, P \rrbracket\left\langle N_{i} \vec{y}_{i}\right\rangle_{i \in I} \in \Lambda\left(\Sigma_{0}^{\prime-}\right)
$$

we have $\tau_{0}^{\prime}(N)=\tau_{0}(L)$ and $\mathscr{L}^{\prime}(N)=\mathscr{L}(L)$. Since $I \neq \varnothing, \llbracket \vec{p}_{I} \rightarrow q, P \rrbracket$ is not nullary. $N$ contains no nullary constants.

The procedure for converting a semilexicalized ACG $\mathscr{G}$ into an equivalent ACG satisfying Condition II in Definition 4.12 is similar to the procedure for Condition I. We define an extension $\mathscr{G}^{\prime}$ of $\mathscr{G}^{\prime}$ assuming that the ACG $\mathscr{G}$ satisfies Condition I. Let $\Delta_{\mathscr{G}} \subseteq\left(\mathscr{A}_{0} \rightarrow \mathscr{A}_{0}\right) \times\left\{P \in \Lambda\left(\Sigma_{1}\right) \mid P\right.$ is a combinator $\}$ be defined as follows:

$$
\begin{aligned}
\Delta_{0} & =\left\{\left\langle p \rightarrow p, \lambda x^{\mathscr{L}(p)} . x\right\rangle \mid p \in \mathscr{A}_{0}\right\} \\
\Delta_{n+1} & =\Delta_{n} \cup\left\{\langle p \rightarrow r,| \lambda x .\left.\mathscr{L}(\mathrm{c})(P x)\right|_{\beta \eta}\right\rangle \mid \mathrm{c} \in \mathscr{C}_{0}^{-}, \tau_{0}(\mathrm{c})=q \rightarrow r, \\
& \left.\langle p \rightarrow q, P\rangle \in \Delta_{n} \text { for some } q \in \mathscr{A}_{0}\right\} \\
\Delta_{\mathscr{G}} & =\bigcup_{n \geq 0} \Delta_{n}
\end{aligned}
$$

By a similar reason to the reason for the computability of $\Gamma_{\mathscr{G}}, \Delta_{\mathscr{G}}$ can be computed effectively.

To satisfy Condition II-i, for each $\langle p \rightarrow q, P\rangle \in \Delta_{\mathscr{G}}$, we add the lexical entries

$$
\langle\llbracket p \rightarrow q, P \rrbracket, p \rightarrow q, P\rangle
$$

To satisfy Condition II-ii, furthermore we add the following constants

$$
\llbracket p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow r,\left|\lambda x_{1} \ldots x_{m} \cdot P_{0}\left(\mathscr{L}(\mathrm{c}) x_{1} \ldots x_{m}\right)\right|_{\beta \eta} \rrbracket
$$

where

- c $\in \mathscr{C}_{0}^{-}$,
- $\tau_{0}(\mathrm{c})=p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow q$,
- $\left\langle q \rightarrow r, P_{0}\right\rangle \in \Delta_{\mathscr{G}}$.

Let $\tau_{0}^{\prime}(\llbracket \alpha, P \rrbracket)=\alpha$ and $\mathscr{L}^{\prime}(\llbracket \alpha, P \rrbracket)=P$. This way, we obtain the following lemma.

Lemma 4.16 (Condition II). Every semilexicalized $A C G \mathscr{G} \in \mathbf{G}(m, n)$ has an equivalent semilexicalized $A C G \mathscr{G}^{\prime} \in \mathbf{G}(m, n)$ that satisfies Conditions I and II in Definition 4.12.

Proof. Let $\mathscr{G}^{\prime}$ be the ACG obtained by the above procedure from an ACG $\mathscr{G}$, which satisfies Condition I. It is clear that $\mathcal{O}(\mathscr{G})=\mathcal{O}\left(\mathscr{G}^{\prime}\right)$ and for every $p, q \in \mathscr{A}_{0}$ and $P \in \Lambda\left(\Sigma_{1}\right)$ the following three are equivalent:

- there is a closed term $M \in \Lambda\left(\Sigma_{0}^{-}\right)$of type $p \rightarrow q$ with $\mathscr{L}(M) \equiv|P|_{\beta \eta}$,
- $\langle p \rightarrow q, P\rangle \in \Delta_{\mathscr{G}}$,
- there is a unary nonlexical constant of type $p \rightarrow q$ in $\mathscr{C}_{0}^{\prime-}$ that is mapped to $P$.

This equivalence relation entails that $\mathscr{G}^{\prime}$ satisfies Condition II-i. To prove that $\mathscr{G}^{\prime}$ satisfies Condition II, it is enough to show that for every closed term $L \in \Lambda\left(\Sigma_{0}^{-}\right)$of type $\vec{p} \rightarrow r$ with $|\vec{p}| \geq 2$, there is $N \in \Lambda\left(\Sigma_{0}^{\prime-}\right)$ such that

- $\tau_{0}(L)=\tau_{0}^{\prime}(N)$ and $\mathscr{L}(L)=\mathscr{L}^{\prime}(N)$,
- if $N$ contains a unary constant $\mathrm{c} \in \mathscr{C}_{0}^{-}$, then the argument of c is a bound variable.

We show that Condition II-ii is satisfied by induction on $\#_{\mathscr{C}_{0}^{-}}(L)$. Let a second-order closed term $L \in \Lambda\left(\Sigma_{0}^{-}\right)$have the form $L=\lambda \vec{y}$.d $\vec{L}$ where $|\vec{y}| \geq 2$ and $\tau_{0}(\mathrm{~d} \vec{L}) \in \mathscr{A}_{0}$. Since $\mathscr{G}$ satisfies Condition I, we can assume that $L$ contains no nullary constants. If $|\vec{L}| \geq 2$, the lemma holds immediately by the induction hypothesis. Suppose that $|\vec{L}|=1 . L$ has the form

$$
L=\vec{y} \cdot \mathrm{~d}_{1}\left(\ldots\left(\mathrm{~d}_{n}\left(\mathrm{c} L_{1} \ldots L_{m}\right)\right) \ldots\right)
$$

where $\mathrm{d}_{1}=\mathrm{d}$, all $\mathrm{d}_{i}$ are unary, and c is $m$-ary for some $m \geq 2$. By $|\vec{y}| \geq 2$, such a constant c occurs in $L$. Let $\tau_{0}(\mathrm{c})=p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow q$ and

$$
P_{0}=\mathscr{L}\left(\lambda z^{q} \cdot \mathrm{~d}_{1}\left(\ldots\left(\mathrm{~d}_{n} z\right) \ldots\right)\right)
$$

Since $\left\langle q \rightarrow r, P_{0}\right\rangle \in \Delta_{\mathscr{G}}$ for $q \rightarrow r=\tau_{0}\left(\lambda z \cdot \mathrm{~d}_{1}\left(\ldots\left(\mathrm{~d}_{n} z\right)\right), \mathscr{C}_{0}^{\prime}\right.$ has the constant

$$
\mathrm{e}=\llbracket p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow r,\left|\lambda x_{1} \ldots x_{m} \cdot P_{0}\left(\mathscr{L}(\mathrm{c}) x_{1} \ldots x_{m}\right)\right|_{\beta \eta} \rrbracket .
$$

Let $N_{i} \in \Lambda\left(\Sigma_{0}^{\prime-}\right)$ be obtained by applying the induction hypothesis to $\lambda \vec{y}_{i} . L_{i}$ where $\vec{y}_{i}$ is the subsequence of $\vec{y}$ whose elements appear in $L_{i}$. Then, the following $N$ satisfies the condition:

$$
N=\lambda \vec{y} \cdot \mathrm{e}\left(N_{1} \vec{y}_{1}\right) \ldots\left(N_{m} \vec{y}_{m}\right) .
$$

### 4.3.3 Second-Order Case

Though Yoshinaka and Kanazawa [61] have already stated that every secondorder ACG has an equivalent lexicalized second-order ACG modulo combinators, the proof of that paper is in error. This subsection presents a correct lexicalization method for second-order ACGs. Another correction is shown by Kanazawa and Yoshinaka [27].

## Elimination of Nullary and Unary Nonlexical Constants

Suppose that a given second-order ACG satisfies Conditions I and II of Definition 4.12. We eliminate nullary (resp. unary) nonlexical constants in a way similar to the procedure to make an second-order ACG satisfy Condition I-ii (resp. Condition II-ii). Let us add the following lexical entries to $\mathscr{G}$ :

$$
\left\langle\llbracket \lambda \vec{x} \cdot \mathrm{a} P_{1} \ldots P_{m} \rrbracket, \vec{p} \rightarrow q, \mathscr{L}\left(\lambda \vec{x} \cdot \mathrm{a} P_{1} \ldots P_{m}\right)\right\rangle
$$

where

- $a \in \mathscr{C}_{0}^{+}$,
- $\tau_{0}(\mathrm{a})=p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow q$ and $\tau_{0}\left(P_{i}\right)=p_{i}$,
- either $P_{i} \in \mathscr{C}_{0}^{-}$or $P_{i}=x_{i}$,
- $p_{i}$ is in $\vec{p}$ iff $x_{i}$ is in $\vec{x}$ iff $P_{i}=x_{i}$.
and
$\left\langle\llbracket \lambda x_{1} \ldots x_{m} . \mathrm{c}\left(\mathrm{b} x_{1} \ldots x_{m}\right) \rrbracket, p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow r, \mathscr{L}\left(\lambda x_{1} \ldots x_{m} . \mathrm{c}\left(\mathrm{b} x_{1} \ldots x_{m}\right)\right)\right\rangle$
where
- $\mathrm{b} \in \mathscr{C}_{0}^{+}$,
- $\tau_{0}(\mathbf{b})=p_{1} \rightarrow \cdots \rightarrow p_{m} \rightarrow q$,
- c $\in \mathscr{C}_{0}^{-}$of type $q \rightarrow r$.

This way, we obtain the following lemma.
Lemma 4.17. Every $A C G \mathscr{G}$ belonging to $\mathbf{G}(2, n)$, has an equivalent (modulo combinators) $A C G \mathscr{G}^{\prime} \in \mathbf{G}(2, n)$ which contains no nullary or unary nonlexical constants.

Proof. Let $\mathscr{G}^{\prime}$ be the ACG obtained from a second-order ACG $\mathscr{G}$ by the above procedure. It is enough to show by induction on $M \in \Lambda\left(\Sigma_{0}\right)-\Lambda\left(\Sigma_{0}^{-}\right)$that if a variable-free term $M$ has an atomic type, then there is $N \in \Lambda\left(\Sigma_{0}\right)$ such that $N \approx M$ and $N$ contains no nullary or unary constants. The proof for Lemma 4.16 can be applied to this claim with few modification. Therefore, when we eliminate all the nullary and unary nonlexical constants from $\mathscr{G}^{\prime}$, the object language does not shrink other than combinators.

## Elimination of Nonlexical Constants

If a second-order ACG $\mathscr{G} \in \mathbf{G}(2, n)$ has no nullary or unary nonlexical constants, then every $M \in \mathcal{A}(\mathscr{G})$ contains more occurrences of lexical constants than those of nonlexical constants. Therefore, as we have described in Section 4.3.1, the equation (4.1) holds for $k_{\mathscr{G}}=1$. Let $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L} \circ \mathscr{L}^{\prime}, s\right\rangle \in$ $\mathrm{G}^{\operatorname{lex}}(4, n)$ where

| $\mathrm{x} \in \mathscr{C}_{0}^{\prime}$ | $\tau_{0}^{\prime}(\mathrm{x})$ | $\mathscr{L}^{\prime}(\mathrm{x})$ | for |
| :---: | :---: | :---: | :---: |
| a | $\tau_{0}(\mathrm{a})$ | a | $\mathrm{a} \in \mathscr{C}_{0}^{+}$ |
| $\mathrm{A}_{\mathrm{ac}}$ | $\left(\tau_{0}(\mathrm{c}) \rightarrow \tau_{0}(\mathrm{a}) \rightarrow s\right) \rightarrow s$ | $\lambda x . x \mathrm{ca}$ | $\mathrm{a} \in \mathscr{C}_{0}^{+}, \mathrm{c} \in \mathscr{C}_{0}^{-}$ |

However, this construction is not satisfactory because the order of $\mathscr{G}^{\prime}$ is four. A naive alternative strategy would be to define $\tau_{0}^{\prime}$ and $\mathscr{L}^{\prime}$ so that

- For each variable-free term $M \in \Lambda\left(\Sigma_{0}\right)$ of an atomic type, there is $P \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\mathscr{L}^{\prime}(P)=M$.

This idea, however, does not work. Suppose that $M=\mathrm{c} M_{1} \ldots M_{k}$ for a nonlexical constant $\mathrm{c} \in \mathscr{C}_{0}^{-}$and we have $P_{1}, \ldots, P_{k} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ with $\mathscr{L}^{\prime}\left(P_{i}\right)=$ $M_{i}$ by the induction hypothesis. To construct $M$ from $\mathscr{L}^{\prime}\left(P_{1}\right), \ldots, \mathscr{L}^{\prime}\left(P_{k}\right)$, exactly one extra nonlexical constant c is required, but no lexical constant is needed. It is impossible if $\mathscr{G}^{\prime}$ is lexicalized. We should use a lexical constant together whenever we use a nonlexical constant. Rather, consider the following statement:
$\star$ For each variable-free term $M \in \Lambda\left(\Sigma_{0}\right)$ of an atomic type, there are $P^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ and $\mathrm{a} \in \mathscr{C}_{0}^{+}$such that $\tau_{0}(\mathrm{a}) \in \mathscr{A}_{0}$ and $\mathscr{L}^{\prime}\left(P^{\prime}\right) \mathrm{a}=M$.

Suppose that $M=\mathrm{c} M_{1} \ldots M_{k}$ for $\mathrm{c} \in \mathscr{C}_{0}^{-}$and we have $\mathscr{L}^{\prime}\left(P_{i}^{\prime}\right) \mathrm{a}_{i}=M_{i}$. Note that $k \geq 2$, since a nonlexical constant is neither nullary nor unary. Then,

$$
\begin{aligned}
M= & \mathrm{c}\left(\mathscr{L}^{\prime}\left(P_{1}^{\prime}\right) \mathrm{a}_{1}\right) \ldots\left(\mathscr{L}^{\prime}\left(P_{k}^{\prime}\right) \mathrm{a}_{k}\right) \\
= & \left(\lambda x_{1} \ldots x_{k} z . \mathrm{c} x_{1} \ldots x_{k-2}\left(x_{k-1} \mathrm{a}_{k-1}\right)\left(x_{k} z\right)\right) \\
& \quad\left(\mathscr{L}^{\prime}\left(P_{1}^{\prime}\right) \mathrm{a}\right) \ldots\left(\mathscr{L}^{\prime}\left(P_{k-2}^{\prime}\right) \mathrm{a}_{k-2}\right) \mathscr{L}^{\prime}\left(P_{k-1}^{\prime}\right) \mathscr{L}^{\prime}\left(P_{k}^{\prime}\right) \mathrm{a}_{k} .
\end{aligned}
$$

If we have a constant $A \in \mathscr{C}_{0}^{\prime}$ such that

$$
\mathscr{L}^{\prime}(\mathrm{A})=\lambda x_{1} \ldots x_{k} z \cdot \mathrm{c} x_{1} \ldots x_{k-2}\left(x_{k-1} \mathrm{a}_{k-1}\right)\left(x_{k} z\right)
$$

then for $P^{\prime}=\mathrm{A}_{\mathrm{ac}}\left(P_{1}^{\prime} \mathrm{a}_{1}\right) \ldots\left(P_{k-2}^{\prime} \mathrm{a}_{k-2}\right) P_{k-1}^{\prime} P_{k}^{\prime}$, we have $M=\mathscr{L}^{\prime}\left(P^{\prime}\right) \mathrm{a}_{k-1}$. That is, we can use the statement $(\star)$ as an induction hypothesis.
Proposition 4.18. Let $\mathscr{G}$ be a second-order ACG that contains no nullary or unary nonlexical constants. For the following third-order ACG $\mathscr{G}^{\prime}=$ $\left\langle\Sigma_{0}^{\prime}, \Sigma_{0}, \mathscr{L}^{\prime}, s\right\rangle$ with $\mathscr{L}^{\prime}(p)=p$, we have $\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\mathcal{A}(\mathscr{G})$ :

| $\mathrm{x} \in \mathscr{C}_{0}^{\prime}$ | $\tau_{0}^{\prime}(\mathrm{x})$ | $\mathscr{L}^{\prime}(\mathrm{x})$ |
| :---: | :---: | :---: |
| a | $\tau_{0}(\mathrm{a})$ | a |
| $\mathrm{D}_{\mathrm{a} c q^{\prime}}$ | $\vec{t} \rightarrow\left(p^{\prime} \rightarrow p\right) \rightarrow\left(q^{\prime} \rightarrow q\right) \rightarrow q \rightarrow r$ | $\lambda \vec{w}^{\vec{t}} x^{p^{\prime} \rightarrow p} y^{q^{\prime} \rightarrow q} z^{q^{\prime}} . \mathrm{c} \vec{w}(x \mathrm{a})(y z)$ |

where $\mathrm{a} \in \mathscr{C}_{0}^{+}, \mathrm{c} \in \mathscr{C}_{0}^{-}, \tau_{0}(\mathrm{a})=p^{\prime}, \tau_{0}(\mathrm{c})=\vec{t} \rightarrow p \rightarrow q \rightarrow r$, and $q^{\prime} \in \mathscr{A}_{0}$.
Proof. It is enough to prove the above statement $(\star)$.
Basis. If $M=\mathrm{a} \in \mathscr{C}_{0}^{+}$, let $P^{\prime}=\lambda x^{\tau_{0}(\mathrm{a})} . x$.
Step. If $M=\mathrm{c} M_{1} \ldots M_{k}$ for $\mathrm{c} \in \mathscr{C}_{0}^{-}$, we can get $P^{\prime}$ that satisfies $(\star)$ as discussed above.

Suppose that $M=\mathrm{b} M_{1} \ldots M_{k}$ for $\mathrm{b} \in \mathscr{C}_{0}^{+}$and $k \geq 1$. By the induction hypothesis, we have $P_{1}^{\prime}, \ldots, P_{k}^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ and $\mathrm{a}_{1}, \ldots, \mathrm{a}_{k} \in \mathscr{C}_{0}^{+}$such that $M_{i}=$ $\mathscr{L}^{\prime}\left(P_{i}^{\prime}\right) \mathrm{a}_{i}$ for each $i \in\{1, \ldots, k\}$. Let

$$
P^{\prime}=\lambda z \cdot \mathrm{~b}\left(P_{1}^{\prime} \mathrm{a}_{1}\right) \ldots\left(P_{k-1}^{\prime} \mathrm{a}_{k-1}\right)\left(P_{k}^{\prime} z\right) .
$$

Corollary 4.19. For every second-order $A C G \mathscr{G} \in \mathbf{G}(2, n)$, there is a thirdorder $A C G \mathscr{G}^{\prime} \in \mathbf{G}^{\operatorname{lex}}(3, n)$ such that $\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\{P \in \mathcal{O}(\mathscr{G}) \mid P$ contains a constant $\}$.

This solution is not yet satisfactory, because the lexicalized ACG is thirdorder.

If every $P \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\mathscr{L}^{\prime}(P)=M$ for $M \in \mathcal{A}(\mathscr{G})$ can be constructed with application only, we can reduce the order of the abstract vocabulary to two. Since the usage of variables and $\lambda$-abstraction in the proof of Proposition 4.18 is tightly limit, we can modify $\mathscr{G}^{\prime}$ into $\mathscr{G}^{\prime \prime}$ so that $P^{\prime}$ in the statement $(\star)$ can be constructed only by application. Let $\mathscr{G}^{\prime \prime}=\left\langle\Sigma_{0}^{\prime \prime}, \Sigma_{0}, \mathscr{L}^{\prime \prime}, s\right\rangle$ consist of the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}^{\prime \prime}$ | $\tau_{0}^{\prime \prime}(\mathrm{x})$ | $\mathscr{L}^{\prime \prime}(\mathrm{x})$ |
| :---: | :---: | :---: |
| $\mathrm{e}_{p}$ | $p \rightarrow p$ | $\lambda x^{p} \cdot x$ |
| a | $\tau_{0}(\mathrm{a})$ | a |
| $\mathrm{D}_{\mathrm{a} c q^{\prime}}$ | $\vec{t} \rightarrow\left(p^{\prime} \rightarrow p\right) \rightarrow\left(q^{\prime} \rightarrow q\right) \rightarrow q^{\prime} \rightarrow r$ | $\lambda \overrightarrow{w^{t}} x^{p^{\prime} \rightarrow p} y^{q^{\prime} \rightarrow q} z^{q^{\prime}} \cdot \mathrm{c} \vec{w}(x \mathrm{a})(y z)$ |
| $\mathrm{b}_{q^{\prime}}$ | $\vec{t} \rightarrow\left(q^{\prime} \rightarrow q\right) \rightarrow q^{\prime} \rightarrow r$ | $\lambda \vec{w}^{\vec{t}} y^{q^{\prime} \rightarrow q} z^{q^{\prime}} \cdot \mathrm{b} \vec{w}(y z)$ |

where $\mathrm{a} \in \mathscr{C}_{0}^{+}, \tau_{0}(\mathrm{a}) \in \mathscr{A}_{0}, \mathrm{c} \in \mathscr{C}_{0}^{-}, \tau_{0}(\mathrm{c})=\vec{t} \rightarrow p \rightarrow q \rightarrow r, \mathrm{~b} \in \mathscr{C}_{0}^{+}$, $\tau_{0}(\mathrm{~b})=\vec{t} \rightarrow q \rightarrow r$ ．
Then，replacing each subtype $(p \rightarrow q)$ with a new abstract atomic type $[p \rightarrow q]^{4}$ ，we get a second－order ACG．Although $\mathscr{L}^{\prime \prime}\left(\mathrm{e}_{p}\right)$ contains no lexi－ cal constants with respect to $\mathscr{L}$ and thus $\mathrm{e}_{p}$ is nonlexical with respect to $\mathscr{L} \circ \mathscr{L}^{\prime \prime}$ ，we can eliminate this nonlexical constant $\mathrm{e}_{p}$ in the same way as Lemma 4．17．This way，we obtain the following definition，where a further minor modification is added for the compactness of the new grammar．

Definition 4．20．Let a second－order ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle \in \mathbf{G}(2, n)$ contain no nullary or unary nonlexical constants．We define a second－order ACG $\mathscr{G}^{l}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L} \circ \mathscr{L}^{\prime}, s\right\rangle \in \mathbf{G}(2, n+1)\left(\mathscr{L}^{\prime}: \Sigma_{0}^{\prime} \rightarrow \Sigma_{0}\right)$ by

$$
\begin{aligned}
\mathscr{A}_{0}^{\prime} & =\mathscr{A}_{0} \cup\left\{[p \rightarrow q] \mid p, q \in \mathscr{A}_{0}\right\} \\
\mathscr{L}^{\prime}(p) & =p \text { for } p \in \mathscr{A}_{0}, \text { and } \mathscr{L}^{\prime}([p \rightarrow q])=p \rightarrow q \text { for } p, q \in \mathscr{A}_{0},
\end{aligned}
$$

| $\mathrm{x} \in \mathscr{C}_{0}^{\prime}$ | $\tau_{0}^{\prime}(\mathrm{x})$ | $\mathscr{L}^{\prime}(\mathrm{x})$ |
| :---: | :---: | :---: |
| $\llbracket \mathrm{a} \rrbracket$ | $p$ | a |
| $\llbracket \mathrm{a} \rrbracket \rrbracket$ | $[p \rightarrow r] \rightarrow r$ | $\lambda x^{p \rightarrow r} . x \mathrm{a}$ |

for $\mathrm{a} \in \mathscr{C}_{0}^{+}, \tau_{0}(\mathrm{a})=p \in \mathscr{A}_{0}, r \in \mathscr{A}_{0}$ ，

| 【b】 | $\vec{t} \rightarrow[q \rightarrow r]$ | $\lambda \overrightarrow{\vec{w}^{t}} z^{q} . \mathrm{b} \vec{w} z$ |
| :---: | :---: | :---: |
| ［ $\mathrm{b}_{p}$ 】 | $\vec{t} \rightarrow[p \rightarrow q] \rightarrow[p \rightarrow r]$ | $\lambda \vec{w}^{\vec{t}} y^{p \rightarrow q} z^{p} \cdot \mathrm{~b} \vec{w}(y z)$ |

for $\mathrm{b} \in \mathscr{C}_{0}^{+}, \tau_{0}(\mathrm{~b})=\vec{t} \rightarrow q \rightarrow r, p \in \mathscr{A}_{0}$,

| $\mathrm{A}_{\mathrm{ac}}$ | $\vec{t} \rightarrow[q \rightarrow r]$ | $\lambda \overrightarrow{w^{t}} z^{q} . \mathrm{c} \vec{w} \mathrm{a} z$ |
| :---: | :---: | :---: |
| for $\mathrm{a} \in \mathscr{C}_{0}^{+}, \tau_{0}(\mathrm{a})=p, \mathrm{c} \in \mathscr{C}_{0}^{-}, \tau_{0}(\mathrm{c})=\vec{t} \rightarrow p \rightarrow q \rightarrow r$ |  |  |


| $\mathrm{B}_{\mathrm{acc} q^{\prime}}$ | $\vec{t} \rightarrow\left[q^{\prime} \rightarrow q\right] \rightarrow\left[q^{\prime} \rightarrow r\right]$ | $\lambda \vec{w}^{\vec{t}} y^{q^{\prime} \rightarrow q} z^{q^{\prime}} . \mathrm{c} \overrightarrow{\mathrm{w} a}(y z)$ |
| :---: | :---: | :---: |


| $\mathrm{C}_{\mathrm{ac}}$ | $\vec{t} \rightarrow\left[p^{\prime} \rightarrow p\right] \rightarrow[q \rightarrow r]$ | $\lambda \overrightarrow{w^{t}} x^{p^{\prime} \rightarrow p} z^{q} . \mathrm{c} \vec{w}(x \mathrm{a}) z$ |
| :---: | :---: | :---: |

for $\mathrm{a} \in \mathscr{C}_{0}^{+}, \tau_{0}(\mathrm{a})=p^{\prime}, \mathrm{c} \in \mathscr{C}_{0}^{-}, \tau_{0}(\mathrm{c})=\vec{t} \rightarrow p \rightarrow q \rightarrow r$,

| $\mathrm{Dacq}_{\text {a }}$ | $\vec{t} \rightarrow\left[p^{\prime} \rightarrow p\right] \rightarrow\left[q^{\prime} \rightarrow q\right] \rightarrow\left[q^{\prime} \rightarrow r\right]$ | $\lambda \overrightarrow{w^{t}} x^{p^{\prime} \rightarrow p} y^{q^{\prime} \rightarrow q} z^{q^{\prime}} \cdot \mathrm{c} \vec{w}(x \mathrm{a})(y z)$ |
| :---: | :---: | :---: |
|  | for $\mathbf{a} \in \mathscr{C}_{0}^{+}, \tau_{0}(\mathbf{a})=p^{\prime}, \mathrm{c} \in \mathscr{C}_{0}^{-}, \tau_{0}$ | $\overrightarrow{\text { a }}$ |

Lemma 4．21． $\mathcal{O}\left(\mathscr{G}^{l}\right)=\mathcal{O}(\mathscr{G})$ ．

[^4]Proof. The inclusion $\mathcal{O}\left(\mathscr{G}^{l}\right) \subseteq \mathcal{O}(\mathscr{G})$ is trivial. To see the converse $\mathcal{O}(\mathscr{G}) \subseteq$ $\mathcal{O}\left(\mathscr{G}^{l}\right)$, we show that the following claim by induction on $M$.

For every variable-free term $M \in \Lambda\left(\Sigma_{0}\right)$ of an atomic type $r$, there is $P \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ of type $r$ such that $\mathscr{L}^{\prime}(P)=M$. Moreover, if $M \notin \mathscr{C}_{0}$, there are $P^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ and $\mathrm{a} \in \mathscr{C}_{0}^{+}$such that $\tau_{0}(\mathrm{a})=p, \tau_{0}^{\prime}\left(P^{\prime}\right)=[p \rightarrow r]$, and $\mathscr{L}^{\prime}\left(P^{\prime}\right) \mathrm{a}=M$.

Basis. $M$ is a constant. If $M=\mathrm{a} \in \mathscr{C}_{0}$, then $\mathrm{a} \in \mathscr{C}_{0}^{+}$, since $\mathscr{G}$ has no nullary nonlexical constants. Let $P=\llbracket a \rrbracket \in \mathscr{C}_{0}^{\prime}$.

Step. Case 1. The head of $M$ is a lexical constant $\mathrm{b} \in \mathscr{C}_{0}^{+}$. Let $M=$ $\mathrm{b} M_{1} \ldots M_{k} N, \tau_{0}(\mathrm{~b})=t_{1} \rightarrow \cdots \rightarrow t_{k} \rightarrow q \rightarrow r, \tau_{0}\left(M_{i}\right)=t_{i}$, and $\tau_{0}(N)=q$. By the induction hypothesis, there are $P_{1}, \ldots, P_{k} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}\left(P_{i}\right)=$ $t_{i}$, and $\mathscr{L}^{\prime}\left(P_{i}\right)=M_{i}$.

Case 1.1. $N=\mathrm{a} \in \mathscr{C}_{0}^{+}$. By the definition, there are constants $\llbracket \mathrm{b} \rrbracket, \llbracket \mathrm{a}_{\mathrm{r}} \rrbracket \in$ $\mathscr{C}_{0}^{\prime}$ such that

$$
\begin{aligned}
\tau_{0}^{\prime}(\llbracket \mathrm{b} \rrbracket) & =t_{1} \rightarrow \cdots \rightarrow t_{k} \rightarrow[q \rightarrow r], & \mathscr{L}^{\prime}(\llbracket \mathrm{b} \rrbracket) & =\lambda w_{1}^{t_{1}} \ldots w_{k}^{t} \\
\tau_{0}^{\prime}\left(\llbracket \mathbf{a}_{r} \rrbracket\right) & =[q \rightarrow r] \rightarrow r, & \mathscr{L}^{\prime}\left(\llbracket \mathbf{a}_{r} \rrbracket\right) & =\lambda x^{q \rightarrow r} . x \mathrm{a} .
\end{aligned}
$$

Let $P^{\prime}=\llbracket \mathrm{b} \rrbracket P_{1} \ldots P_{k}$ and $P=\llbracket \mathrm{a}_{r} \rrbracket P^{\prime}$. We have $\tau_{0}\left(P^{\prime}\right)=[q \rightarrow r], \tau_{0}(P)=r$, and

$$
\mathscr{L}^{\prime}(P)=\mathscr{L}^{\prime}\left(P^{\prime}\right) \mathrm{a}=\left(\lambda z^{q} \cdot \mathrm{~b} M_{1} \ldots M_{k} z\right) \mathrm{a}=\mathrm{b} M_{1} \ldots M_{k} N=M .
$$

Case 1.2. $N \notin \mathscr{C}_{0}$. By the induction hypothesis, there are $Q^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ and $\mathrm{a} \in \mathscr{C}_{0}^{+}$such that $\tau_{0}^{\prime}\left(Q^{\prime}\right)=[p \rightarrow q], \tau_{0}(\mathrm{a})=p$, and $\mathscr{L}^{\prime}\left(Q^{\prime}\right) \mathrm{a}=N$. There is a constant $\llbracket \mathrm{b}_{p} \rrbracket \in \mathscr{C}_{0}^{\prime}$ such that

$$
\begin{aligned}
\tau_{0}^{\prime}\left(\mathrm{b}_{p}\right) & =t_{1} \rightarrow \cdots \rightarrow t_{k} \rightarrow[p \rightarrow q] \rightarrow[p \rightarrow r] \\
\mathscr{L}^{\prime}\left(\mathrm{b}_{p}\right) & =\lambda w_{1}^{t_{1}} \ldots w_{k}^{t_{k}} y^{p \rightarrow q} z^{p} \cdot \mathrm{~b} w_{1} \ldots w_{k}(y z)
\end{aligned}
$$

Let $P^{\prime}=\llbracket \mathrm{b}_{p} \rrbracket P_{1} \ldots P_{k} Q^{\prime}$ and $P=\llbracket \mathrm{a}_{r} \rrbracket P^{\prime}$. We have $\tau_{0}^{\prime}\left(P^{\prime}\right)=[p \rightarrow r]$, $\tau_{0}^{\prime}(P)=r$, and

$$
\mathscr{L}^{\prime}(P)=\mathscr{L}^{\prime}\left(P^{\prime}\right) \mathrm{a}=\mathrm{b} M_{1} \ldots M_{k}\left(\mathscr{L}^{\prime}\left(Q^{\prime}\right) \mathrm{a}\right)=\mathrm{b} M_{1} \ldots M_{k} N=M
$$

Case 2. The head of $M$ is a nonlexical constant $\mathrm{c} \in \mathscr{C}_{0}^{-}$. Let $M=$ $\mathrm{c} M_{1} \ldots M_{k} N_{1} N_{2}$ and $\tau_{0}(\mathrm{c})=t_{1} \rightarrow \cdots \rightarrow t_{k} \rightarrow p \rightarrow q \rightarrow r$. By the induction hypothesis, there are $P_{1}, \ldots, P_{k}$ such that $\tau_{0}^{\prime}\left(P_{i}\right)=t_{i}$ and $\mathscr{L}^{\prime}\left(P_{i}\right)=M_{i}$.

Case 2.1. $N_{1}=\mathrm{a} \in \mathscr{C}_{0}^{+}$and $N_{2}=\mathrm{b} \in \mathscr{C}_{0}^{+}$. There is a constant $\mathrm{A}_{\mathrm{ac}} \in \mathscr{C}_{0}^{\prime}$ such that

$$
\begin{aligned}
\tau_{0}^{\prime}\left(\mathrm{A}_{\mathrm{ac}}\right) & =t_{1} \rightarrow \cdots \rightarrow t_{k} \rightarrow[q \rightarrow r] \\
\mathscr{L}^{\prime}\left(\mathrm{A}_{\mathrm{ac}}\right) & =\lambda w_{1}^{t_{1}} \ldots w_{k}^{t_{k}} z^{q} \cdot \mathrm{c} w_{1} \ldots w_{k} \mathrm{a} z .
\end{aligned}
$$

Let $P^{\prime}=\mathrm{A}_{\mathrm{ac}} P_{1} \ldots P_{k}$ and $P=\llbracket \mathrm{b}_{r} \rrbracket P^{\prime}$. We have $\tau_{0}^{\prime}\left(P^{\prime}\right)=[q \rightarrow r], \tau_{0}^{\prime}(P)=r$, and

$$
\mathscr{L}^{\prime}(P)=\mathscr{L}^{\prime}\left(P^{\prime}\right) \mathrm{b}=\mathrm{c} M_{1} \ldots M_{k} \mathrm{ab}=M .
$$

Case 2.2. $N_{1}=\mathrm{a} \in \mathscr{C}_{0}^{+}$and $N_{2} \notin \mathscr{C}_{0}^{+}$. By the induction hypothesis, we have $Q_{2}^{\prime}$ of type $\left[q^{\prime} \rightarrow q\right]$ such that $\mathscr{L}^{\prime}\left(Q_{2}^{\prime}\right) \mathrm{b}=N_{2}$ for some $\mathrm{b} \in \mathscr{C}_{0}^{+}$of type $q^{\prime}$. There is a constant $\mathrm{B}_{\mathrm{a} c q^{\prime}} \in \mathscr{C}_{0}^{\prime}$ such that

$$
\begin{aligned}
\tau_{0}^{\prime}\left(\mathrm{B}_{\mathrm{a} c q^{\prime}}\right) & =t_{1} \rightarrow \cdots \rightarrow t_{k} \rightarrow\left[q^{\prime} \rightarrow q\right] \rightarrow\left[q^{\prime} \rightarrow r\right] \\
\mathscr{L}^{\prime}\left(\mathrm{B}_{\mathrm{a} q^{\prime}}\right) & =\lambda w_{1}^{t_{1}} \ldots w_{k}^{t_{k}} y^{q^{\prime} \rightarrow q} z^{q^{\prime}} . \mathrm{c} w_{1} \ldots w_{k} \mathrm{a}(y z) .
\end{aligned}
$$

Let $P^{\prime}=\mathrm{B}_{\mathrm{ac} q^{\prime}} P_{1} \ldots P_{k} Q_{2}^{\prime}$ and $P=\llbracket \mathrm{b}_{r} \rrbracket P^{\prime}$. we have $\tau_{0}^{\prime}\left(P^{\prime}\right)=\left[q^{\prime} \rightarrow r\right]$, $\tau_{0}^{\prime}(P)=r$ and

$$
\mathscr{L}^{\prime}(P)=\mathscr{L}^{\prime}\left(P^{\prime}\right) \mathrm{b}=\mathrm{c} M_{1} \ldots M_{k} \mathrm{a}\left(\mathscr{L}^{\prime}\left(Q_{2}^{\prime}\right) \mathrm{b}\right)=\mathrm{c} M_{1} \ldots M_{k} N_{1} N_{2}=M .
$$

Case 2.3. $N_{1} \notin \mathscr{C}_{0}^{+}$and $N_{2}=\mathrm{b} \in \mathscr{C}_{0}^{+}$. By the induction hypothesis, we have $Q_{1}^{\prime}$ of type $\left[p^{\prime} \rightarrow p\right]$ such that $\mathscr{L}^{\prime}\left(Q_{1}^{\prime}\right)$ a $=N_{1}$ for some a $\in \mathscr{C}_{0}^{+}$of type $p^{\prime}$. There is a constant $\mathrm{C}_{\mathrm{ac}} \in \mathscr{C}_{0}^{\prime}$ such that

$$
\begin{aligned}
\tau_{0}^{\prime}\left(\mathrm{C}_{\mathrm{ac}}\right) & =t_{1} \rightarrow \cdots \rightarrow t_{k} \rightarrow\left[p^{\prime} \rightarrow p\right] \rightarrow[q \rightarrow r], \\
\mathscr{L}^{\prime}\left(\mathrm{C}_{\mathrm{ac}}\right) & =\lambda w_{1}^{t_{1}} \ldots w_{k}^{t_{k}} x^{p^{\prime} \rightarrow p} z^{q} \cdot \mathrm{c} w_{1} \ldots w_{k}(x \mathrm{a}) z .
\end{aligned}
$$

For $P^{\prime}=\mathrm{C}_{\mathrm{ac}} P_{1} \ldots P_{k} Q_{1}^{\prime}$ and $P=\llbracket \mathrm{b}_{r} \rrbracket P^{\prime}$, we have $\tau_{0}^{\prime}\left(P^{\prime}\right)=[q \rightarrow r], \tau_{0}^{\prime}(P)=r$ and

$$
\mathscr{L}^{\prime}(P)=\mathscr{L}^{\prime}\left(P^{\prime}\right) \mathrm{b}=\mathrm{c} M_{1} \ldots M_{k}\left(\mathscr{L}^{\prime}\left(Q_{1}^{\prime}\right) \mathrm{a}\right) \mathrm{b}=\mathrm{c} M_{1} \ldots M_{k} N_{1} N_{2}=M
$$

Case 2.4. $N_{1} \notin \mathscr{C}_{0}^{+}$and $N_{2} \notin \mathscr{C}_{0}^{+}$. By the induction hypothesis, we have $Q_{1}^{\prime}, Q_{2}^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ and $\mathrm{a}, \mathrm{b} \in \mathscr{C}_{0}^{+}$such that

$$
\begin{array}{lll}
\tau_{0}^{\prime}\left(Q_{1}^{\prime}\right)=\left[p^{\prime} \rightarrow p\right], & \mathscr{L}^{\prime}\left(Q_{1}^{\prime}\right) \mathrm{a}=N_{1}, & \tau_{0}(\mathrm{a})=p^{\prime}, \\
\tau_{0}^{\prime}\left(Q_{2}^{\prime}\right)=\left[q^{\prime} \rightarrow q\right], & \mathscr{L}^{\prime}\left(Q_{2}^{\prime}\right) \mathrm{b}=N_{2}, & \tau_{0}(\mathrm{~b})=q^{\prime} .
\end{array}
$$

By the definition of $\mathscr{G}^{l}$, there is a constant $\mathrm{D}_{\mathrm{ac} q^{\prime}} \in \mathscr{C}_{0}^{\prime}$ such that

$$
\begin{aligned}
\tau_{0}^{\prime}\left(\mathrm{D}_{\mathrm{ac} q^{\prime}}\right) & =t_{1} \rightarrow \cdots \rightarrow t_{k} \rightarrow\left[p^{\prime} \rightarrow p\right] \rightarrow\left[q^{\prime} \rightarrow q\right] \rightarrow\left[q^{\prime} \rightarrow r\right], \\
\mathscr{L}^{\prime}\left(\mathrm{D}_{\mathrm{ac} q^{\prime}}\right) & =\lambda w_{1}^{t_{1}} \ldots w_{k}^{t_{k}} x^{p^{\prime} \rightarrow p} y^{q^{\prime} \rightarrow q} z^{q^{\prime}} . \mathrm{c}_{1} \ldots w_{k}(x \mathrm{a})(y z) .
\end{aligned}
$$

Let $P^{\prime}=\mathrm{D}_{\mathrm{ac} q^{\prime}} P_{1} \ldots P_{k} Q_{1}^{\prime} Q_{2}^{\prime}$ and $P=\llbracket \mathrm{b}_{r} \rrbracket P^{\prime}$. We have $\tau_{0}^{\prime}\left(P^{\prime}\right)=\left[q^{\prime} \rightarrow r\right]$, $\tau_{0}^{\prime}(P)=r$ and

$$
\mathscr{L}^{\prime}(P)=\mathscr{L}^{\prime}\left(P^{\prime}\right) \mathrm{b}=\mathrm{c} M_{1} \ldots M_{k}\left(\mathscr{L}^{\prime}\left(Q_{1}^{\prime}\right) \mathrm{a}\right)\left(\mathscr{L}^{\prime}\left(Q_{2}^{\prime}\right) \mathrm{b}\right)=M .
$$

Theorem 4.22. For every second-order $A C G \mathscr{G} \in \mathbf{G}(2, n)$, there is a lexicalized second-order $A C G \mathscr{G}^{\prime} \in \mathbf{G}(2, n+1)$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\{R \in \mathcal{O}(\mathscr{G}) \mid R \text { contains a constant }\}
$$

### 4.3.4 Fourth or Higher-Order Case

As we have mentioned at the beginning of Section 4.3.2, we start this section with elimination of nullary and unary nonlexical constants.

## Elimination of Nullary and Unary Nonlexical Constants

Lemma 4.23 (Elimination of Nullary Nonlexical Constants). For every semilexicalized $A C G \mathscr{G} \in \mathbf{G}(m, n)$, we can find a semilexicalized $A C G$ $\mathscr{G}^{\prime} \in \mathbf{G}(\max \{3, m\}, n)$ such that $\mathscr{G}^{\prime}$ contains no nullary nonlexical constants and

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\{R \in \mathcal{O}(\mathscr{G}) \mid R \text { contains a constant }\} .
$$

Proof. By Lemma 4.16, we can assume that a given ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ satisfies Conditions I and II in Definition 4.12. We let an ACG $\mathscr{G}^{\prime}=$ $\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L}^{\prime}, s\right\rangle$ have the following lexical entries, where $\mathscr{A}_{0}^{\prime}=\mathscr{A}_{0}$ and $\mathscr{L}^{\prime}(p)=$ $\mathscr{L}(p)$ for $p \in \mathscr{A}_{0}$.

- $\left\langle\mathrm{c}, \tau_{0}(\mathrm{c}), \mathscr{L}(\mathrm{c})\right\rangle$ for all $\mathrm{c} \in \mathscr{C}_{0}^{-}$unless $\tau_{0}(\mathrm{c}) \in \mathscr{A}_{0}$,
- $\left\langle\llbracket M \rrbracket, \tau_{0}(M), \mathscr{L}(M)\right\rangle$ for each $M$ of the form

$$
M=\lambda x_{1} \ldots x_{n} \cdot \mathrm{a}\left(x_{1} \mathrm{c}_{1,1} \ldots \mathrm{c}_{1, m_{1}}\right) \ldots\left(x_{n} \mathrm{c}_{n, 1} \ldots \mathrm{c}_{n, m_{n}}\right)
$$

where a $\in \mathscr{C}_{0}^{+}, \tau_{0}(\mathrm{a})=\gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n} \rightarrow q, \mathrm{c}_{i, j} \in \mathscr{C}_{0}^{-}, \tau_{0}\left(\mathrm{c}_{i, j}\right)=p_{i, j}$, $\tau_{0}^{\prime}\left(x_{i}\right)=p_{i, 1} \rightarrow \cdots \rightarrow p_{i, m_{i}} \rightarrow \gamma_{i}$ and $0 \leq m_{i} \leq\left|\gamma_{i}\right|$.

Since the order of the type of the bound variable $x_{i}$ in the above $M$ is

$$
\operatorname{ord}\left(p_{i, 1} \rightarrow \cdots \rightarrow p_{i, m_{i}} \rightarrow \gamma_{i}\right)= \begin{cases}\max \left\{2, \operatorname{ord}\left(\gamma_{i}\right)\right\} & \text { if } m_{i} \geq 1 \\ \operatorname{ord}\left(\gamma_{i}\right) & \text { if } m_{i}=0\end{cases}
$$

we have $\operatorname{ord}\left(\tau_{0}^{\prime}(\llbracket M \rrbracket)\right) \leq \max \left\{3, \operatorname{ord}\left(\tau_{0}(\mathrm{a})\right)\right\}$. Thus, $\mathscr{G}^{\prime} \in \mathbf{G}(\max \{3, m\}, n)$ if $\mathscr{G} \in \mathbf{G}(m, n)$. Moreover, $\mathscr{G}^{\prime}$ also satisfies Condition II.
$\left[\mathcal{O}\left(\mathscr{G}^{\prime}\right) \subseteq\{R \in \mathcal{O}(\mathscr{G}) \mid R\right.$ contains a constant $\left.\}\right]$
The inclusion $\mathcal{O}\left(\mathscr{G}^{\prime}\right) \subseteq \mathcal{O}(\mathscr{G})$ is obvious. We show that $\mathcal{O}\left(\mathscr{G}^{\prime}\right)$ contains no combinators. Since every nonlexical constant $\mathrm{c} \in \mathscr{C}_{0}^{\prime-}$ has non-nullary second-order type, if $P \in \Lambda\left(\Sigma_{0}^{\prime-}\right)$ is closed, it cannot have an atomic type, in particular, the distinguished type. Thus if $P \in \mathcal{A}\left(\mathscr{G}^{\prime}\right)$, it contains a lexical constant.
$\left[\{R \in \mathcal{O}(\mathscr{G}) \mid R\right.$ contains a constant $\left.\} \subseteq \mathcal{O}\left(\mathscr{G}^{\prime}\right)\right]$
We say that an occurrence of a nullary nonlexical constant is linked to an occurrence of a lexical constant if the path which starts from the type of the
occurrence of the nullary nonlexical constant ends in a subtype of the type of the occurrence of the lexical constant. A term $M \in \Lambda\left(\Sigma_{0}\right)$ is said to satisfy the linking condition if for every occurrence of a nullary nonlexical constant, there is an occurrence of a lexical constant to which the occurrence of the nullary nonlexical constant is linked.

In fact, every $M \in \mathcal{A}(\mathscr{G})-\Lambda\left(\Sigma_{0}^{-}\right)$has an equivalent term $M^{\prime} \approx M$ that satisfies the linking condition. Since $\mathscr{G}$ satisfies Condition I in Definition 4.12, we can assume that if a nullary constant c occurs in $M$, the occurrence is as an argument of a lexical constant or a bound variable, by Lemma 4.13. If the functor is a lexical constant a, then c is linked to the lexical constant a. If the functor is a bound variable, then as we have described in Section 4.3.1, we can trace the path until it ends in a subtype of the type of a lexical constant. Thus, here we assume that $M$ satisfies the linking condition.

By induction on the size of $M$, we show that for every long normal term $M \in \Lambda\left(\Sigma_{0}\right)$ that satisfies the linking condition, there is $P \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ with $\tau_{0}^{\prime}(P)=\tau_{0}(M)$ and $\mathscr{L}^{\prime}(P)=\mathscr{L}(M)$.

Case 1. Suppose that $M$ is of the form $M=\lambda x . M^{\prime}$. Since $M$ satisfies the linking condition, also $M^{\prime}$ satisfies the linking condition. By the induction hypothesis, we have $P^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ with $\tau_{0}^{\prime}\left(P^{\prime}\right)=\tau_{0}\left(M^{\prime}\right)$ and $\mathscr{L}^{\prime}\left(P^{\prime}\right)=\mathscr{L}\left(M^{\prime}\right)$. For $P=\lambda x \cdot P^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$, we have $\tau_{0}^{\prime}(P)=\tau_{0}(M)$ and $\mathscr{L}^{\prime}(P)=\mathscr{L}(M)$.

Case 2. If $M$ is of the form $M=\mathrm{c}$ for a nullary nonlexical constant c , then $M$ cannot satisfy the linking condition.

Case 3. If $M$ is of the form $M=x M_{1} \ldots M_{n}$ for some variable or nonnullary nonlexical constant x, then each $M_{i}$ satisfies the linking condition (see Remark 4.11). By the induction hypothesis, there is $P_{i} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ with $\tau_{0}^{\prime}\left(P_{i}\right)=\tau_{0}\left(M_{i}\right)$ and $\mathscr{L}^{\prime}\left(P_{i}\right)=\mathscr{L}\left(M_{i}\right)$. For $P=x P_{1} \ldots P_{n}$, we have $\tau_{0}^{\prime}(P)=$ $\tau_{0}(M)$ and $\mathscr{L}^{\prime}(P)=\mathscr{L}(M)$.

Case 4. The remaining case is that $M$ is of the form

$$
M=\mathrm{a} M_{1} \ldots M_{n}
$$

for some lexical constant a $\in \mathscr{C}_{0}^{+}$of type $\gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n} \rightarrow q$. For each $i \in$ $\{1, \ldots, n\}$, let $M_{i}^{\prime}$ be obtained from $M_{i}$ by replacing with fresh variables $\vec{z}_{i}$ all the occurrences of nullary nonlexical constants linked to the head occurrence of a, so that $M_{i}^{\prime}$ satisfies the linking condition. By the induction hypothesis, there is $P_{i}^{\prime}$ such that $\tau_{0}^{\prime}\left(P_{i}^{\prime}\right)=\tau_{0}\left(M_{i}^{\prime}\right)$ and $\mathscr{L}^{\prime}\left(P_{i}^{\prime}\right)=\mathscr{L}\left(M_{i}^{\prime}\right)$. Let $\overrightarrow{\mathrm{c}}_{i}$ be such that

$$
M_{i}=M_{i}^{\prime}\left[\vec{c}_{i} / \vec{z}_{i}\right] .
$$

By the property of paths, $\left|\vec{c}_{i}\right| \leq\left|\gamma_{i}\right|$. By the definition of $\mathscr{G}^{\prime}$, we have $\llbracket N \rrbracket \in \mathscr{C}_{0}^{\prime}$ for

$$
N=\lambda x_{1} \ldots x_{n} \cdot \mathbf{a}\left(x_{1} \overrightarrow{\mathrm{c}}_{1}\right) \ldots\left(x_{n} \vec{c}_{n}\right) .
$$

Thus, for $P=\llbracket N \rrbracket\left(\lambda \vec{z}_{1} \cdot P_{1}^{\prime}\right) \ldots\left(\lambda \vec{z}_{n} \cdot P_{n}^{\prime}\right)$, we have $\tau_{0}^{\prime}(P)=\tau_{0}(M)$ and

$$
\mathscr{L}^{\prime}(P)=\mathscr{L}\left(\mathrm{a}\left(M_{1}^{\prime}\left[\overrightarrow{\mathrm{c}}_{1} / \vec{z}_{1}\right]\right) \ldots\left(M_{m}^{\prime}\left[\vec{c}_{n} / \vec{z}_{n}\right]\right)\right)=\mathscr{L}(M)
$$

Lemma 4.24 (Elimination of Unary Nonlexical Constants). For every semilexicalized $A C G \mathscr{G} \in \mathbf{G}(m, n)$, we can find a semilexicalized $A C G$ $\mathscr{G}^{\prime} \in \mathbf{G}(\max \{4, m\}, n)$ such that $\mathscr{G}^{\prime}$ contains no nullary or unary nonlexical constants and

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\{R \in \mathcal{O}(\mathscr{G}) \mid R \text { contains a constant }\}
$$

Proof. By Lemmas 4.16 and 4.23 , we assume that $\mathscr{G}$ contains no nullary nonlexical constants and satisfies Condition II in Definition 4.12. $\mathcal{O}(\mathscr{G})$ contains no combinators.

We let an ACG $\mathscr{G}^{\prime}$ have the following lexical entries:

- $\left\langle\mathrm{c}, \tau_{0}(\mathrm{c}), \mathscr{L}(\mathrm{c})\right\rangle$ for all $\mathrm{c} \in \mathscr{C}_{0}^{-}$unless $\tau_{0}(\mathrm{c})$ is unary,
- $\left\langle\llbracket M \rrbracket, \tau_{0}(M), \mathscr{L}(M)\right\rangle$ for each $M$ of the form

$$
M=\lambda x_{1}^{\alpha_{1,1} \rightarrow \cdots \rightarrow \alpha_{1, m_{1}} \rightarrow \gamma_{1}} \ldots x_{n}^{\alpha_{n, 1} \rightarrow \cdots \rightarrow \alpha_{n, m_{n}} \rightarrow \gamma_{n}} \cdot \mathrm{c}_{0}\left(\mathrm{a}\left(x_{1} \overrightarrow{\mathrm{c}}_{1}\right) \ldots\left(x_{n} \overrightarrow{\mathrm{c}}_{n}\right)\right)
$$

where $\mathrm{a} \in \mathscr{C}_{0}^{+}, \tau_{0}(\mathrm{a})=\gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n} \rightarrow q, \overrightarrow{\mathrm{c}}_{i}=\mathrm{c}_{i, 1} \ldots \mathrm{c}_{i, m_{i}}, m_{i} \leq\left|\gamma_{i}\right|$, and each $\mathbf{c}_{i, j}$ is a nonlexical constant of unary type $\alpha_{i, j}$.

Since the order of the type of the bound variables $x_{i}$ in $M$ is

$$
\begin{aligned}
& \operatorname{ord}\left(\alpha_{i, 1} \rightarrow \cdots \rightarrow \alpha_{i, m_{i}} \rightarrow \gamma_{i}\right) \\
= & \max \left\{\operatorname{ord}\left(\alpha_{i, j}\right)+1, \operatorname{ord}\left(\gamma_{i}\right) \mid 1 \leq j \leq m_{i}\right\} \leq \max \left\{3, \operatorname{ord}\left(\gamma_{i}\right)\right\},
\end{aligned}
$$

we have $\operatorname{ord}\left(\tau_{0}^{\prime}(\llbracket M \rrbracket)\right) \leq \max \left\{4, \operatorname{ord}\left(\tau_{0}(\mathrm{a})\right)\right\}$. Thus, $\mathscr{G}^{\prime} \in \mathbf{G}(\max \{4, m\}, n)$ if $\mathscr{G} \in \mathbf{G}(m, n)$.

The inclusion $\mathcal{O}\left(\mathscr{G}^{\prime}\right) \subseteq \mathcal{O}(\mathscr{G})$ is trivial. We show the converse relation $\mathcal{O}(\mathscr{G}) \subseteq \mathcal{O}\left(\mathscr{G}^{\prime}\right)$.

We say that an occurrence of a unary nonlexical constant of type $p \rightarrow q$ is linked to an occurrence of a lexical constant if the path which starts from the negative occurrence of the subtype $p$ in the type $p \rightarrow q$ of the occurrence of the unary nonlexical constant ends in $p$ in the type of the occurrence of the lexical constant. A term $M \in \Lambda\left(\Sigma_{0}\right)$ is said to satisfy the linking condition if for every occurrence of a unary nonlexical constant, there is an occurrence of a lexical constant to which the occurrence of the unary nonlexical constant is linked.

In fact, every $M \in \mathcal{A}(\mathscr{G})-\Lambda\left(\Sigma_{0}^{-}\right)$has an equivalent term $M^{\prime} \approx M$ that satisfies the linking condition. Since $\mathscr{G}$ satisfies Condition II in Definition 4.12, we can assume that if a unary constant c occurs in $M$, the head of the argument of c is either a lexical constant or a bound variable by Lemma 4.14. If it is a lexical constant a, then c is linked to the lexical constant a. If it is a bound variable, then as we have described in Section 4.3.1, we can trace the path until it ends in a subtype of the type of a lexical constant. Thus, here we assume that $M$ satisfies the linking condition.

By induction on the size of $M$, we show that for every long normal term $M \in \Lambda\left(\Sigma_{0}\right)$ that satisfies the linking condition, there is $P \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ with $\tau_{0}^{\prime}(P)=\tau_{0}(M)$ and $\mathscr{L}^{\prime}(P)=\mathscr{L}(M)$.

Case 1. $M=\lambda x . M^{\prime}$ for some $x$. By the induction hypothesis we have $P^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}\left(P^{\prime}\right)=\tau_{0}\left(M^{\prime}\right)$ and $\mathscr{L}^{\prime}\left(P^{\prime}\right)=\mathscr{L}\left(M^{\prime}\right)$. Thus, for $P=\lambda x . P^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$, we have $\tau_{0}^{\prime}(P)=\tau_{0}(M)$ and $\mathscr{L}^{\prime}(P)=\mathscr{L}(M)$.

Case 2. $M=\times M_{1} \ldots M_{n}$ where x is a variable or a non-unary nonlexical constant. By the induction hypothesis we have $P_{i} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}\left(P_{i}\right)=$ $\tau_{0}\left(M_{i}\right)$ and $\mathscr{L}^{\prime}\left(P_{i}\right)=\mathscr{L}\left(M_{i}\right)$ for $1 \leq i \leq n$. Thus, for $P=x P_{1} \ldots P_{n} \in$ $\Lambda\left(\Sigma_{0}^{\prime}\right)$, we have $\tau_{0}^{\prime}(P)=\tau_{0}(M)$ and $\mathscr{L}^{\prime}(P)=\mathscr{L}(M)$.

Case 3. Suppose that the head of $M$ is a unary nonlexical constant $c_{0} \in \mathscr{C}_{0}^{-}$of type $p \rightarrow q$. Since $M$ satisfies the linking condition, $M$ is of the form

$$
M=\mathrm{c}_{0}\left(\mathrm{a} M_{1} \ldots M_{n}\right)
$$

for some lexical constant a of type $\gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n} \rightarrow p$. Let $M_{i}^{\prime}$ be obtained from $M_{i}$ by replacing with fresh variables $\vec{z}_{i}$ all the occurrences of unary nonlexical constants linked to the head occurrence of a, so that $M_{i}^{\prime}$ satisfies the linking condition. By the induction hypothesis, there is $P_{i}^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}\left(P_{i}^{\prime}\right)=\tau_{0}\left(M_{i}^{\prime}\right)$ and $\mathscr{L}^{\prime}\left(P_{i}^{\prime}\right)=\mathscr{L}\left(M_{i}^{\prime}\right)$. Let $\overrightarrow{\mathrm{c}}_{i}$ be such that

$$
M_{i}=M_{i}^{\prime}\left[\overrightarrow{\mathrm{c}}_{i} / \vec{z}_{i}\right] .
$$

By the property of paths, $\left|\vec{c}_{i}\right| \leq\left|\gamma_{i}\right|$. By the definition of $\mathscr{G}^{\prime}$, we have $\llbracket N \rrbracket \in \mathscr{C}_{0}^{\prime}$ where

$$
N=\lambda w_{1} \ldots w_{n} \cdot \mathrm{c}_{0}\left(\mathrm{a}\left(w_{1} \overrightarrow{\mathrm{c}}_{1}\right) \ldots\left(w_{n} \overrightarrow{\mathrm{c}}_{n}\right)\right) .
$$

Thus, for $P=\llbracket N \rrbracket\left(\lambda \vec{z}_{1} \cdot P_{1}^{\prime}\right) \ldots\left(\lambda \vec{z}_{n} \cdot P_{n}^{\prime}\right)$, we have

$$
\begin{aligned}
\mathscr{L}^{\prime}(P) & =\mathscr{L}\left(\mathrm{c}_{0}\left(\mathrm{a}\left(M_{1}^{\prime}\left[\overrightarrow{\mathrm{c}}_{1} / \vec{z}_{1}\right]\right) \ldots\left(M_{n}^{\prime}\left[\overrightarrow{\mathrm{c}}_{n} / \vec{z}_{n}\right]\right)\right)\right) \\
& =\mathscr{L}\left(\mathrm{c}_{0}\left(\mathrm{a} M_{1} \ldots M_{n}\right)\right)=\mathscr{L}(M) .
\end{aligned}
$$

Case 4. Suppose that the head of $M$ is a lexical constant a of type $\gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n} \rightarrow p . \quad M$ has form $M=\mathrm{a} M_{1} \ldots M_{n}$. Since $\mathscr{G}$ satisfies

Condition II-i, there is a nonlexical constant $\mathrm{e}_{p} \in \mathscr{C}_{0}$ of type $p \rightarrow p$ with $\mathscr{L}\left(\mathrm{e}_{p}\right)=\lambda x . x$, so $M \approx \mathrm{e}_{p} M$. This case is reduced to Case 3 , where we use the induction hypothesis with respect to each $M_{i}$ but not to a $M_{1} \ldots M_{n}$.

## Elimination of Nonlexical Constants

Lemma 4.24 implies that every semilexicalized ACG has an equivalent lexicalized ACG as we have discussed in Section 4.3.1. The lexicalization method in Section 4.3.1 constructs a new ACG whose object vocabulary is the original abstract vocabulary and whose lexical entries have the form

$$
\left\langle\llbracket \mathrm{a}, \overrightarrow{\mathrm{c}} \rrbracket,\left(\tau_{0}(\mathrm{a}) \rightarrow \tau_{0}(\overrightarrow{\mathrm{c}}) \rightarrow s\right) \rightarrow s, \lambda w \cdot w \mathrm{a} \overrightarrow{\mathrm{c}}\right\rangle,
$$

where the bound variable $w$ "grasps" and "controls" original abstract constants and then puts them at any places in a term we want. That trick is powerful, but causes the increase of the order of the grammar by two. To avoid this, we refrain from controlling lexical abstract constants, but continue to grasp nonlexical ones, by a bound variable in new lexical entries. Since the orders of nonlexical constants are at most two, this new strategy does not increase the order of the grammar if the order of the given semilexicalized ACG is four or higher.
Definition 4.25. Let a semilexicalized ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle \in \mathbf{G}(m, n)$ contain no nullary or unary nonlexical constants. Let $\Gamma \subseteq \Lambda\left(\Sigma_{0}\right)$ be the set of closed terms of the form

$$
\begin{gathered}
\lambda w_{1} \ldots w_{n} \cdot \mathrm{a}\left(w_{1} \overrightarrow{\mathrm{c}}_{1}\right) \ldots\left(w_{n} \overrightarrow{\mathrm{c}}_{n}\right) \quad \text { or } \\
\left.\lambda w_{0} w_{1} \ldots w_{n} \cdot w_{0} \mathrm{c}_{0}\left(\mathrm{a}\left(w_{1} \overrightarrow{\mathrm{c}}_{1}\right) \ldots\left(w_{n} \overrightarrow{\mathrm{c}}_{n}\right)\right)\right)
\end{gathered}
$$

where $\mathrm{a} \in \mathscr{C}_{0}^{+}, \mathrm{c}_{0}$ and constants in $\overrightarrow{\mathrm{c}}_{i}$ are all nonlexical, $\tau_{0}(\mathrm{a})=\gamma_{1} \rightarrow \cdots \rightarrow$ $\gamma_{n} \rightarrow p, \tau_{0}\left(w_{0}\right)=\tau_{0}\left(\mathrm{c}_{0}\right) \rightarrow p \rightarrow q$ for some $q \in \mathscr{A}_{0}, \tau_{0}\left(w_{i}\right)=\tau_{0}\left(\vec{c}_{i}\right) \rightarrow \gamma_{i}$ for $i \geq 1$, and $0 \leq\left|\vec{c}_{i}\right| \leq\left|\tau_{0}\left(\gamma_{i}\right)\right|$. Note that $\Gamma$ is a finite set.

The lexicalized form of $\mathscr{G}$ is an ACG $\mathscr{G}^{l} \in \mathbf{G}(\max \{4, m\}, n)$ that has the form

$$
\mathscr{G}^{l}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L} \circ \mathscr{L}^{\prime}, s\right\rangle
$$

where

$$
\begin{aligned}
\mathscr{A}_{0}^{\prime} & =\mathscr{A}_{0}, \\
\mathscr{C}_{0}^{\prime} & =\{\llbracket M \rrbracket \mid M \in \Gamma\}, \\
\tau_{0}^{\prime}(\llbracket M \rrbracket) & =\tau_{0}(M), \\
\mathscr{L}^{\prime}(p) & =p \text { for } p \in \mathscr{A}_{0}, \\
\mathscr{L}^{\prime}(\llbracket M \rrbracket) & =M \text { for } \llbracket M \rrbracket \in \mathscr{C}_{0}^{\prime} .
\end{aligned}
$$

Example 4.26. Let $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ have a lexical constant a $\in \mathscr{C}_{0}^{+}$of type $(s \rightarrow s) \rightarrow s$ and a nonlexical constant $\mathrm{c} \in \mathscr{C}_{0}^{-}$of type $s^{2} \rightarrow s$. Let us consider the terms $M$ and $c M M$ in $\mathcal{A}(\mathscr{G})$, where

$$
M=\mathrm{a}\left(\lambda x_{1}^{s} \cdot \mathrm{a}\left(\lambda x_{2}^{s} \cdot \mathrm{a}\left(\lambda x_{3}^{s} \cdot \mathrm{a}\left(\lambda x_{4}^{s} \cdot \mathrm{c}\left(\mathrm{c} x_{1} x_{2}\right)\left(\mathrm{c} x_{3} x_{4}\right)\right)\right)\right)\right)
$$

Let $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{0}, \mathscr{L}^{\prime}, s\right\rangle$ consist of the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}^{\prime}$ | $\tau_{0}^{\prime}(\mathrm{x})$ | $\mathscr{L}^{\prime}(\mathbf{x})$ |
| :---: | :---: | :---: |
| a | $(s \rightarrow s) \rightarrow s$ | a |
| A | $\gamma \rightarrow s$ | $\lambda w^{\gamma} \cdot \mathrm{a}(w \mathrm{c})$ |
| B | $\gamma \rightarrow \gamma \rightarrow s$ | $\lambda w_{0}^{\gamma} w^{\gamma} \cdot w_{0} \mathrm{ca}(w \mathrm{c})$ |

where $\gamma=\left(s^{2} \rightarrow s\right) \rightarrow s \rightarrow s$. For

$$
\begin{aligned}
N= & \mathrm{A}\left(\lambda y_{1}^{s^{2} \rightarrow s} x_{1}^{s} \cdot \mathrm{~A}\left(\lambda y_{2}^{s^{2} \rightarrow s} x_{2}^{s} \cdot \mathrm{~A}\left(\lambda y_{3}^{s^{2} \rightarrow s} x_{3}^{s} \cdot \mathrm{a}\left(\lambda x_{4}^{s} \cdot y_{1}\left(y_{2} x_{1} x_{2}\right)\left(y_{3} x_{3} x_{4}\right)\right)\right)\right)\right), \\
N^{\prime}= & \mathrm{B}\left(\lambda y_{0}^{s^{2} \rightarrow s} z^{s} \cdot y_{0} z N\right) \\
& \quad\left(\lambda y_{1}^{s^{2} \rightarrow s} x_{1}^{s} \cdot \mathrm{~A}\left(\lambda y_{2}^{s^{2} \rightarrow s} x_{2}^{s} \cdot \mathrm{~A}\left(\lambda y_{3}^{s^{2} \rightarrow s} x_{3}^{s} \cdot \mathrm{a}\left(\lambda x_{4}^{s} \cdot y_{1}\left(y_{2} x_{1} x_{2}\right)\left(y_{3} x_{3} x_{4}\right)\right)\right)\right)\right),
\end{aligned}
$$

we have $\mathscr{L}^{\prime}(N)=M$ and $\mathscr{L}^{\prime}\left(N^{\prime}\right)=\mathrm{c} M M$.
Definition 4.27. Let $\mathscr{G}$ be a semilexicalized ACG that has no nullary or unary nonlexical constants. A $\beta$-expansion $\left(\lambda x_{1} \ldots x_{m} . M_{0}\right) M_{1} \ldots M_{m}$ of a $\beta$-normal term $M \in \Lambda\left(\Sigma_{0}\right)$ is standard if

- $M_{0} \in \Lambda\left(\Sigma_{0}^{+}\right)$,
- $M_{i} \in \Lambda\left(\Sigma_{0}^{-}\right)$for $i \in\{1, \ldots, m\}$,
- every bound variable of $M_{i}$ has an atomic type for $i \in\{1, \ldots, m\}$,
- for every free variable $y$ of $M_{i}$ for $i \geq 1$, there is a nonlexical constant c with $\tau_{0}(y)=\tau_{0}(\mathrm{c})$.
- there is no subterm of the form $x_{i} \vec{N}_{1}\left(x_{j} \vec{N}_{2}\right) \vec{N}_{3}$ of $M_{0}$ for $i, j \in\{1, \ldots, m\}$.

Let \#( $\left.M_{0}, x_{i}\right)$ denote the number of negative occurrences of atomic types in the type $\tau_{0}\left(x_{i}\right)$ of the occurrence of $x_{i}$ which are linked to a subtype of the type of some occurrence of a lexical constant in $M_{0}$. A standard $\beta$-expansion is good iff $\#_{\mathscr{C}_{0}^{-}}\left(M_{i}\right) \leq \#\left(M_{0}, x_{i}\right)$ holds for every $M_{i}$.

Note that every $M_{i}\left(x_{i}\right)$ has a second-order type for $i \geq 1$.
Lemma 4.28. For every $M \in \mathcal{A}(\mathscr{G}), M$ has a good standard $\beta$-expansion.

Proof. First we show that $M$ has a standard $\beta$-expansion. Suppose that $M \in \mathcal{A}(\mathscr{G})$ has $m$ occurrences of nonlexical constants $\mathrm{c}_{1}, \ldots, \mathrm{c}_{m}$. Let $M_{0} \in$ $\Lambda\left(\Sigma_{0}^{+}\right)$be such that $M=\left(\lambda x_{1} \ldots x_{m} \cdot M_{0}\right) \mathrm{c}_{1} \ldots \mathrm{c}_{m}$ and $M_{i}=\mathrm{c}_{i}$ for $i \in$ $\{1, \ldots, m\}$. If $\left(\lambda x_{1} \ldots x_{m} \cdot M_{0}\right) M_{1} \ldots M_{m}$ is a good $\beta$-expansion of $M$, the proof is done. Otherwise, $M_{0}$ has the form $M_{0}=N_{0}\left[x_{i} \vec{N}_{1}\left(x_{j} \vec{N}_{2}\right) / w\right]$ for some $i, j \in\{1, \ldots, m\}$. Assume that $i=m-1$ and $j=m$. Let $M_{0}^{\prime}=$ $N_{0}\left[x_{i} \vec{N}_{1} \vec{N}_{2} / w\right]$ and $M_{i}^{\prime}=\lambda \vec{z}_{1} \vec{z}_{2} \cdot M_{i} \vec{z}_{1}\left(M_{j} \vec{z}_{2}\right)$, where $\left|\vec{z}_{1}\right|=\left|\vec{N}_{1}\right|$ and $\left|\overrightarrow{z_{2}}\right|=$ $\left|\vec{N}_{2}\right|$. If $\left(\lambda x_{1} \ldots x_{m-1} \cdot M_{0}^{\prime}\right) M_{1} \ldots M_{m-2} M_{m-1}^{\prime}$ is a good $\beta$-expansion of $M$, the proof is done. Otherwise, by repeating this procedure, eventually we get a good $\beta$-expansion of $M$.

Let $\left(\lambda \vec{x} \cdot M_{0}\right) \vec{M}$ be a standard $\beta$-expansion of $M$ and $\tau_{0}\left(x_{i}\right)=p_{i, 1} \rightarrow \cdots \rightarrow$ $p_{i, n_{i}} \rightarrow p_{i}$. The occurrence of $p_{i, j}$ must be linked to some subtype of the type of an occurrence of a lexical constant, as we have described in Section 4.3.1. Thus \# $\left(M_{0}, x_{i}\right)=n_{i}$. On the other hand, the long normal form of $M_{i}$ has the form $\lambda z_{i, 1} \ldots z_{i, n_{1}} \cdot M_{i}^{\prime}$, where $M_{i}^{\prime}$ has an atomic type. $M_{i}^{\prime}$ can be regarded as a tree on $\mathscr{C}_{0}^{-} \cup\left\{z_{i, 1}, \ldots, z_{i, n_{i}}\right\}$ whose leaves are exactly $z_{i, 1}, \ldots, z_{i, n_{i}}$. Recall that if a tree contains no node of rank 1 , then the number of leaves is larger than the number of internal nodes. Since $M_{i}^{\prime}$ contains no nullary or unary constant, we have $\#_{\mathscr{C}_{0}^{-}}\left(M_{i}\right)<n_{i}$. Thus the standard $\beta$-expansion is good.

Lemma 4.29. If $M$ has a good $\beta$-expansion and contains a nonlexical constant, then $M$ contains some lexical constant.

Proof. Let $\left(\lambda x_{1} \ldots x_{m} \cdot M_{0}\right) M_{1} \ldots M_{m}$ be a standard $\beta$-expansion of $M$. Since $M$ contains a nonlexical constant, there is $M_{k}$ with $\#_{\mathscr{C}_{0}^{-}}\left(M_{k}\right) \geq 1$. By $\#_{\mathscr{C}_{0}^{-}}\left(M_{k}\right) \leq \#\left(M_{0}, x_{k}\right), M_{0}$ contains a lexical constant.

Lemma 4.30. For a semilexicalized $A C G \mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle \in \mathbf{G}(m, n)$ which contains no nullary or unary nonlexical constants, let $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{0}\right.$, $\left.\mathscr{L}^{\prime}, s\right\rangle \in \mathbf{G}(m, 1)$ where $\Sigma_{0}^{\prime}$ and $\mathscr{L}^{\prime}$ are defined in Definition 4.25. We have $\mathcal{A}(\mathscr{G})=\mathcal{O}\left(\mathscr{G}^{\prime}\right)$.

Proof. The inclusion $\mathcal{O}\left(\mathscr{G}^{\prime}\right) \subseteq \mathcal{A}(\mathscr{G})$ is trivial.
To show the converse, by induction on the size of $M$, we show that if a $\beta$-normal term $M \in \Lambda\left(\Sigma_{0}\right)$ has a good $\beta$-expansion, then there is $P \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}(P)=\tau_{0}(M)$ and $\mathscr{L}^{\prime}(P)=M$.

Let $\left(\lambda \vec{x} . M_{0}\right) \vec{M}$ be a good $\beta$-expansion of $M$. There are four cases, depending on the form of $M_{0}$. Without loss of generality, we can assume that every element of $\vec{M}$ contains a constant. If $\left(\lambda \vec{x}_{1} x_{0} \vec{x}_{2} \cdot M_{0}\right) \vec{M}_{1} L \vec{M}_{2}$ is a good $\beta$-expansion of $M$ where $L$ is constant-free, then $\left(\lambda \vec{x}_{1} \vec{x}_{2} \cdot M_{0}\left[L / x_{0}\right]\right) \vec{M}_{1} \vec{M}_{2}$ is also a good $\beta$-expansion.

Case 1. $M_{0}=\lambda z . M_{0}^{\prime}$. Clearly $\left(\lambda \vec{x} \cdot M_{0}^{\prime}\right) \vec{M}$ is a good $\beta$-expansion of its $\beta$-normal form $M^{\prime}=\left|\left(\lambda \vec{x} \cdot M_{0}^{\prime}\right) \vec{M}\right|_{\beta}$. Since $M^{\prime}$ is strictly smaller than $M=$ $\lambda z . M^{\prime}$, by the induction hypothesis, we get $P^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ with $\tau_{0}^{\prime}\left(P^{\prime}\right)=\tau_{0}\left(M^{\prime}\right)$ and $\mathscr{L}^{\prime}\left(P^{\prime}\right)=M^{\prime}$. For $P=\lambda z . P^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$, we have $\tau_{0}^{\prime}(P)=\tau_{0}(M)$ and $\mathscr{L}^{\prime}(P)=M$.

Case 2. $M_{0}=w N_{1,0} \ldots N_{n, 0}$ for some variable $w$ not in $\vec{x}$. Let $\vec{x}_{i}$ be the subsequence of $\vec{x}$ such that $\vec{x}_{i}$ consists of the variables appearing in $N_{i, 0}$, and $\vec{M}_{i}$ the corresponding subsequence of $\vec{M}$. Let $N_{i}=N_{i, 0}\left[\vec{M}_{i} / \vec{x}_{i}\right]$. That is,

$$
\begin{aligned}
M & =M_{0}[\vec{M} / \vec{x}]=w N_{1,0} \ldots N_{n, 0}[\vec{M} / \vec{x}] \\
& =w\left(N_{1,0}\left[\vec{M}_{1} / \vec{x}_{1}\right]\right) \ldots\left(N_{n, 0}\left[\vec{M}_{n} / \vec{x}_{n}\right]\right)=w N_{1} \ldots N_{n} .
\end{aligned}
$$

Clearly $\left(\lambda \vec{x}_{i} . N_{i, 0}\right) \vec{M}_{i}$ is a good $\beta$-expansion of $N_{i}$ (see Remark 4.11). By the induction hypothesis, we have $P_{i} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ with $\tau_{0}^{\prime}\left(P_{i}\right)=\tau_{0}\left(N_{i}\right)$ and $\mathscr{L}^{\prime}\left(P_{i}\right)=N_{i}$. For $P=w P_{1} \ldots P_{n} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$, we have $\tau_{0}^{\prime}(P)=\tau_{0}(M)$ and $\mathscr{L}^{\prime}(P)=M$.

Case 3. $M_{0}=\mathrm{a} N_{1,0} \ldots N_{n, 0}$ for some lexical constant $\mathrm{a} \in \mathscr{C}_{0}^{+}$of type $\gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n} \rightarrow p$. Let $x_{i, 1} \ldots x_{i, m_{i}}$ be the subsequence of $\vec{x}$ such that $x_{i, j}$ appears in $N_{i, 0}$, and $M_{i, 1} \ldots M_{i, m_{i}}$ the corresponding subsequence of $\vec{M}$. Let

$$
\begin{equation*}
N_{i}=N_{i, 0}\left[M_{i, j} / x_{i, j}\right]_{1 \leq j \leq m_{i}} . \tag{4.6}
\end{equation*}
$$

That is,

$$
\begin{aligned}
\left(\lambda \vec{x} \cdot M_{0}\right) \vec{M} & =\left(\lambda\left\langle x_{i, j}\right\rangle_{1 \leq i \leq n}^{1 \leq m_{i}} . M_{0}\right)\left\langle M_{i, j}\right\rangle_{1 \leq j \leq m_{i}}^{1 \leq i \leq n} \\
& =M_{0}\left[M_{i, j} / x_{i, j}\right]_{1 \leq i \leq j \leq m_{i}}^{1 \leq j} \\
& =\mathrm{a} N_{1,0} \ldots N_{n, 0}\left[M_{i, j} / x_{i, j}\right]_{1 \leq j \leq m_{i}}^{1 \leq i \leq n} \\
& =\mathrm{a}\left(N_{1,0}\left[M_{1, j} / x_{1, j}\right]_{1 \leq j \leq m_{1}}\right) \ldots\left(N_{n, 0}\left[M_{n, j} / x_{n, j}\right]_{1 \leq j \leq m_{n}}\right) \\
& =\mathrm{a} N_{1} \ldots N_{n}=M .
\end{aligned}
$$

For each $i, j$ with $1 \leq i \leq n$ and $1 \leq j \leq m_{i}$, Let

$$
\begin{equation*}
d_{i, j}=\#\left(M_{0}, x_{i, j}\right)-\#\left(N_{i, 0}, x_{i, j}\right) \tag{4.7}
\end{equation*}
$$

that is, $d_{i, j}$ denotes the number of paths from negative occurrences of atomic types in the type of the variable $x_{i, j}$ ending in the type of the head occurrence of a in $M_{0}$. Hence $\sum_{1 \leq j \leq m_{i}} d_{i, j} \leq\left|\gamma_{i}\right|$ holds. Let

$$
\begin{equation*}
k_{i j}=\min \left\{d_{i, j}, \#_{\mathscr{C}_{0}^{-}}\left(M_{i, j}\right)\right\} . \tag{4.8}
\end{equation*}
$$

We have

$$
\begin{equation*}
\sum_{1 \leq j \leq m_{i}} k_{i j} \leq \sum_{1 \leq j \leq m_{i}} d_{i, j} \leq\left|\gamma_{i}\right| . \tag{4.9}
\end{equation*}
$$

Let $M_{i, j}^{\prime}$ be obtained from $M_{i, j}$ by replacing first $k_{i j}$ occurrences of nonlexical constants $\mathrm{c}_{i, j, 1}, \ldots, \mathrm{c}_{i, j, k_{i j}}$ with fresh variables $y_{i, j, 1}, \ldots, y_{i, j, k_{i j}}$, i.e.,

$$
M_{i, j}=M_{i, j}^{\prime}\left[\mathrm{c}_{i, j, 1} / y_{i, j, 1}, \ldots, \mathrm{c}_{i, j, k_{i j}} / y_{i, j, k_{i j}}\right] .
$$

Let

$$
N_{i}^{\prime} \equiv N_{i, 0}\left[M_{i, j}^{\prime} / x_{i, j}\right]_{1 \leq j \leq m_{i}},
$$

i.e., by (4.6),

$$
\begin{equation*}
N_{i} \equiv N_{i}^{\prime}\left[\vec{c}_{i} / \vec{y}_{i}\right], \tag{4.10}
\end{equation*}
$$

where $\vec{y}_{i}=\left\langle y_{i, j, h}\right\rangle_{1 \leq h \leq k_{i j}}^{\substack{1 \leq j \leq m_{i}}}$ and $\overrightarrow{\mathrm{c}}_{i}=\left\langle\mathrm{c}_{i, j, h}\right\rangle_{1 \leq h \leq k_{i j}}^{1 \leq j \leq m_{i}}$. Since $\left(\lambda \vec{x} . M_{0}\right) \vec{M}$ is a good $\beta$-expansion of $M, \#_{\mathscr{C}_{0}}\left(M_{i, j}\right)-\#\left(M_{0}, x_{i, j}\right) \leq 0$ holds. Hence

$$
\begin{align*}
\#_{\mathscr{C}_{0}^{-}}\left(M_{i, j}^{\prime}\right) & =\#_{\mathscr{C}_{0}^{-}}\left(M_{i, j}\right)-k_{i j} \\
& =\max \left\{0, \#_{\mathscr{C}_{0}^{-}}\left(M_{i, j}\right)-d_{i, j}\right\}  \tag{4.8}\\
& =\max \left\{0, \# \mathscr{\mathscr { C }}_{0}^{-}\right.  \tag{4.7}\\
& \left.\left(M_{i, j}\right)-\left(\#\left(M_{0}, x_{i, j}\right)-\#\left(N_{i, 0}, x_{i, j}\right)\right)\right\} \\
& \leq \max \left\{0, \#\left(N_{i, 0}, x_{i, j}\right)\right\} \\
& =\#\left(N_{i, 0}, x_{i, j}\right) .
\end{align*}
$$

Therefore, $\left(\lambda x_{i, 1} \ldots x_{i, m_{i}} . N_{i, 0}\right) M_{i, 1}^{\prime} \ldots M_{i, m_{i}}^{\prime}$ is a good $\beta$-expansion of $N_{i}^{\prime}$. By the induction hypothesis, we have $P_{i}^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that

$$
\begin{equation*}
\mathscr{L}^{\prime}\left(P_{i}^{\prime}\right)=N_{i}^{\prime} . \tag{4.11}
\end{equation*}
$$

Since $\left|\vec{c}_{i}\right|=\sum_{1 \leq j \leq m_{i}} k_{i j} \leq\left|\gamma_{i}\right|$ by (4.9), we have a constant $A \in \mathscr{C}_{0}^{\prime}$ such that

$$
\mathscr{L}^{\prime}(\mathbf{A}) \equiv \lambda w_{1} \ldots w_{n} \cdot \mathrm{a}\left(w_{1} \overrightarrow{\mathrm{c}}_{1}\right) \ldots\left(w_{n} \overrightarrow{\mathrm{c}}_{n}\right)
$$

For $P \equiv \mathrm{~A}\left(\lambda \vec{y}_{1} \cdot P_{1}^{\prime}\right) \ldots\left(\lambda \vec{y}_{n} \cdot P_{n}^{\prime}\right) \in \Lambda\left(\Sigma_{0}^{\prime}\right)$, we have

$$
\begin{align*}
\mathscr{L}^{\prime}(P) & =\mathscr{L}^{\prime}(\mathrm{A})\left(\lambda \vec{y}_{1} \cdot N_{1}^{\prime}\right) \ldots\left(\lambda \vec{y}_{n} \cdot N_{n}^{\prime}\right)  \tag{4.11}\\
& =\mathrm{a}\left(N_{1}^{\prime}\left[\overrightarrow{\mathrm{c}}_{1} / \vec{y}_{1}\right]\right) \ldots\left(N_{n}^{\prime}\left[\overrightarrow{\mathrm{c}}_{n} / \vec{y}_{n}\right]\right) \\
& =\mathrm{a} N_{1} \ldots N_{n}  \tag{4.10}\\
& =M .
\end{align*}
$$

Case 4. For a good $\beta$-expansion $\left(\lambda \vec{x} . M_{0}\right) \vec{M}$ of $M, M_{0}$ is of the form $x_{i} \vec{K}$ for some $x_{i}$ in $\vec{x}$. Without loss of generality, we can assume that it is $x_{1}$. Let

$$
\begin{aligned}
M_{0} & =x_{1} K_{1} \ldots K_{m} \\
M & =M_{0}[\vec{M} / \vec{x}] \\
& =x_{1} K_{1} \ldots K_{m}\left[M_{1} / x_{1}, \vec{M}_{1} / \vec{x}_{1}, \ldots, \vec{M}_{m} / \vec{x}_{m}\right] \\
& =M_{1} K_{1}\left[\vec{M}_{1} / \vec{x}_{1}\right] \ldots K_{m}\left[\vec{M}_{m} / \vec{x}_{m}\right]
\end{aligned}
$$

where $\vec{x}_{i}$ is the subsequence of $\vec{x}$ consisting of the variables in $K_{i}$, and $\vec{M}_{i}$ is the corresponding subsequence of $\vec{M}$. Since $x_{1}$ has a second-order type, each $K_{i}$ has an atomic type.

Recall that we assume that $M_{1}$ contains at least one nonlexical constant. Since

$$
\#\left(M_{0}, x_{1}\right) \geq \#_{\mathscr{C}_{0}^{-}}\left(M_{1}\right) \geq 1
$$

there is $K_{k}$ for $k \in\{1, \ldots, m\}$ of the form

$$
K_{k}=\mathrm{a} N_{1,0} \ldots N_{n, 0}
$$

for some $\mathrm{a} \in \mathscr{C}_{0}^{+}$of type $\gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n} \rightarrow p$ and $N_{i, 0} \in \Lambda\left(\Sigma_{0}^{+}\right)$of type $\gamma_{i}$ for $i \in\{1, \ldots, n\}$. Let $M_{1}^{\prime}$ be obtained from $M_{1}$ by replacing an occurrence of a constant $\mathrm{c}_{0} \in \mathscr{C}_{0}^{-}$with $y_{0}$, and $M_{0}^{\prime}$ obtained from $M_{0}$ by replacing $K_{k}$ with a fresh variable $z_{k}$, i.e.,

$$
\begin{aligned}
& M_{1}=M_{1}^{\prime}\left[c_{0} / y_{0}\right] \\
& M_{0}^{\prime}=x_{1} K_{1} \ldots K_{k-1} z_{k} K_{k+1} \ldots K_{m}
\end{aligned}
$$

and define $M^{\prime}$ as

$$
M^{\prime}=M_{0}^{\prime}\left[M_{1}^{\prime} / x_{1}, \vec{M}_{1} / \vec{x}_{1}, \ldots, \vec{M}_{k-1} / \vec{x}_{k-1}, \vec{M}_{k+1} / \vec{x}_{k+1}, \ldots, \vec{M}_{m} / \vec{x}_{m}\right]
$$

Then,

$$
\begin{align*}
M & =M_{0}\left[M_{1} / x_{1}, \vec{M}_{1} / \vec{x}_{1}, \ldots, \vec{M}_{m} / \vec{x}_{m}\right] \\
& =M_{0}^{\prime}\left[K_{k} / z_{k}\right]\left[M_{1}^{\prime}\left[c_{0} / y_{0}\right] / x_{1}, \vec{M}_{1} / \vec{x}_{1}, \ldots, \vec{M}_{m} / \vec{x}_{m}\right] \\
& =M^{\prime}\left[c_{0} / y_{0}, K_{k}\left[\vec{M}_{k} / \vec{x}_{k}\right] / z_{k}\right] . \tag{4.12}
\end{align*}
$$

For $i \neq k$, we have $\#\left(M_{0}^{\prime}, x_{i, j}\right)=\#\left(M_{0}, x_{i, j}\right) \geq \#_{\mathscr{C}_{0}^{-}}\left(M_{i, j}\right)$ for every $x_{i, j}$ in $\vec{x}_{i}$ and the corresponding term $M_{i, j}$ in $\vec{M}_{i}$, and

$$
\#_{\mathscr{C}_{0}^{-}}\left(M_{1}^{\prime}\right)=\#_{\mathscr{C}_{0}^{-}}\left(M_{1}\right)-1 \leq \#\left(M_{0}, x_{1}\right)-1=\#\left(M_{0}^{\prime}, x_{1}\right) .
$$

Therefore

$$
\left(\lambda x_{1} \vec{x}_{1} \ldots \vec{x}_{k-1} \vec{x}_{k+1} \ldots \vec{x}_{m} \cdot M_{0}^{\prime}\right) M_{1}^{\prime} \vec{M}_{1} \ldots \vec{M}_{k-1} \vec{M}_{k+1} \ldots \vec{M}_{m}
$$

is a good $\beta$-expansion of $M^{\prime}$. By the induction hypothesis we have $P^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that

$$
\mathscr{L}^{\prime}\left(P^{\prime}\right)=M^{\prime}
$$

On the other hand, by $\#\left(M_{0}, x_{i, j}\right)=\#\left(K_{i}, x_{i, j}\right)$ for each $x_{i, j}$ in $\vec{x}_{i}$ for all $i \in\{1, \ldots, m\},\left(\lambda \vec{x}_{i} \cdot K_{i}\right) \vec{M}_{i}$ is a good $\beta$-expansion of $K_{i}\left[\vec{M}_{i} / \vec{x}_{i}\right]$. Therefore, we can apply the same discussion ${ }^{5}$ in Case 3 to

$$
K_{k}\left[\vec{M}_{k} / \vec{x}_{k}\right]=\mathrm{a} N_{1,0} \ldots N_{n, 0}\left[\vec{M}_{k} / \vec{x}_{k}\right]
$$

and then we find $P_{i}^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ and a sequence $\vec{c}_{i}$ of nonlexical constants such that $\left|\vec{c}_{i}\right| \leq\left|\gamma_{i}\right|$ and

$$
\mathscr{L}^{\prime}\left(P_{i}^{\prime}\right)\left[\vec{c}_{i} / \vec{y}_{i}\right]=N_{i, 0}\left[\vec{M}_{k} / \vec{x}_{k}\right] .
$$

By the definition of $\Sigma_{0}^{\prime}$, we have a constant B which is mapped to

$$
\mathscr{L}^{\prime}(\mathrm{B})=\lambda w_{0} \ldots w_{n} \cdot w_{0} \mathrm{c}_{0}\left(\mathrm{a}\left(w_{1} \overrightarrow{\mathrm{c}}_{1}\right) \ldots\left(w_{n} \overrightarrow{\mathrm{c}}_{n}\right)\right)
$$

Thus, for $P=\mathrm{B}\left(\lambda y_{0} z_{k} \cdot P^{\prime}\right)\left(\lambda \vec{y}_{1} \cdot P_{1}^{\prime}\right) \ldots\left(\lambda \vec{y}_{n} \cdot P_{n}^{\prime}\right) \in \Lambda\left(\Sigma_{0}^{\prime}\right)$, we have

$$
\begin{align*}
\mathscr{L}^{\prime}(P) & =\left(\lambda y_{0} z_{k} \cdot \mathscr{L}^{\prime}\left(P^{\prime}\right)\right) \mathrm{c}_{0}\left(\mathrm{a}\left(\mathscr{L}^{\prime}\left(P_{1}^{\prime}\right)\left[\vec{c}_{1} / \vec{y}_{1}\right]\right) \ldots\left(\mathscr{L}^{\prime}\left(P_{n}^{\prime}\right)\left[\vec{c}_{n} / \vec{y}_{n}\right]\right)\right) \\
& =M^{\prime}\left[c_{0} / y_{0}, \mathrm{a} N_{i, 0} \ldots N_{n, 0}\left[\vec{M}_{k} / \vec{x}_{k}\right] / z_{k}\right] \\
& =M^{\prime}\left[c_{0} / y_{0}, K_{k}\left[\vec{M}_{k} / \vec{x}_{k}\right] / z_{k}\right] \\
& =M . \tag{4.12}
\end{align*}
$$

Theorem 4.31. For every semilexicalized $A C G \mathscr{G} \in \mathbf{G}(m, n)$, there is a lexicalized $A C G \mathscr{G}^{\prime} \in \mathbf{G}(\max \{4, m\}, n)$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\{R \in \mathcal{O}(\mathscr{G}) \mid R \text { contains a constant }\}
$$

Proof. By Lemmas 4.28 and 4.30.

[^5]
### 4.3.5 Third-Order Case

In the previous subsection, for the fourth or higher-order case, we have constructed a lexicalized ACG $\mathscr{G}^{l}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L} \circ \mathscr{L}^{\prime}, s\right\rangle$, where the ACG $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{0}, \mathscr{L}^{\prime}, s\right\rangle$ "controls" nonlexical (w.r.t. $\mathscr{L}$ ) constants $\mathrm{c} \in \mathscr{C}_{0}^{-}$by bound variables in its lexical entries, and then we can put nonlexical constants in almost arbitrary places. Since variables that can be functors of second-order nonlexical constants are at least third-order, then the order of abstract vocabulary becomes more than three.

Recall Example 4.26. For the third-order semilexicalized ACG $\mathscr{G} \in$ $\mathbf{G}(3, n)$, let

$$
\begin{equation*}
M=\mathrm{a}_{1}\left(\lambda x_{1}^{s} \cdot \mathrm{a}_{2}\left(\lambda x_{2}^{s} \cdot \mathrm{a}_{3}\left(\lambda x_{3}^{s} \cdot \mathrm{a}_{4}\left(\lambda x_{4}^{s} \cdot \mathrm{c}_{1}\left(\mathrm{c}_{2} x_{1} x_{2}\right)\left(\mathrm{c}_{3} x_{3} x_{4}\right)\right)\right)\right)\right) \in \mathcal{A}(\mathscr{G}) \tag{4.13}
\end{equation*}
$$

where $\mathrm{a}_{i} \in \mathscr{C}_{0}^{+}$has type $(s \rightarrow s) \rightarrow s$ and $\mathrm{c}_{i} \in \mathscr{C}_{0}^{-}$has type $s^{2} \rightarrow s$. Here we number each occurrence of a constant to facilitate discussion. In Example 4.26, we first replace occurrences of $\mathrm{c}_{i}$ with new bound variables $y_{i}$, and then replace occurrences of $\mathrm{a}_{i} N_{i}$ with $\mathrm{A}_{i}\left(\lambda y_{i} \cdot N_{i}\right)$, where $\mathrm{A}_{i} \in \mathscr{C}_{0}^{\prime}$ is a new abstract constant mapped to $\lambda w \cdot \mathrm{a}_{i}\left(w \mathrm{c}_{i}\right)$. The resultant term is

$$
\begin{equation*}
\mathrm{A}_{1}\left(\lambda y_{1}^{s^{2} \rightarrow s} x_{1}^{s} \cdot \mathrm{~A}_{2}\left(\lambda y_{2}^{s^{2} \rightarrow s} x_{2}^{s} \cdot \mathrm{~A}_{3}\left(\lambda y_{3}^{s^{2} \rightarrow s} x_{3}^{s} \cdot \mathrm{a}_{4}\left(\lambda x_{4}^{s} \cdot y_{1}\left(y_{2} x_{1} x_{2}\right)\left(y_{3} x_{3} x_{4}\right)\right)\right)\right)\right) . \tag{4.14}
\end{equation*}
$$

This way we associate occurrences of nonlexical constants with occurrences of lexical constants. To keep the order of the abstract vocabulary in three, however, that strategy must be abandoned. We are no longer able to use second-order variables in the abstract vocabulary.

Recall the idea that we have used for lexicalization of second-order ACGs, where the third-order lexicalized ACG $\mathscr{G}^{\prime}$ in Proposition 4.18 is transformed into the second-order lexicalized ACG $\mathscr{G}^{l}$ in Definition 4.20. We have introduced new abstract atomic types in $\mathscr{A}_{0}^{\prime}$ mapped to second-order types on $\mathcal{T}\left(\mathscr{A}_{0}\right)$ by the new lexicon $\mathscr{L}^{\prime}$. That trick works well because the usage of free variables and $\lambda$-abstraction in an element of $\mathcal{A}\left(\mathscr{G}^{\prime}\right)$ is very much limited so that we can embed terms involving $\lambda$-abstraction into finitely many lexical entries of $\mathscr{G}^{\prime \prime}$ shown in (4.5) on page 64 .

We want to apply that technique to the third-order case. As we see in (4.14), the lexicalization method for the fourth or higher-order case constructs the form of the nonlexical part $c_{1}\left(c_{2} x_{1} x_{2}\right)\left(c_{3} x_{3} x_{4}\right)$ by variables $y_{i}$ of second-order type, and then substitutes nonlexical constants for those variables. To exclude variables of second-order types from terms in the abstract language in the third-order case, we must embed those variables into the lexical entries so that new lexical entries can build any nonlexical parts in a term in the abstract language.

Let us associate occurrences of nonlexical constants with ones of lexical constants in $M$ as follows:


According to this association, we let an ACG $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{0}, \mathscr{L}^{\prime},[s]\right\rangle \in \mathbf{G}(3,2)$ consist of the following lexical entries. ${ }^{6}$

| $\mathrm{x} \in \mathscr{C}_{0}^{\prime}$ | $\tau_{0}^{\prime}(\mathrm{x})$ | $\mathscr{L}^{\prime}(\mathrm{x})$ |
| :---: | :---: | :---: |
| $\mathrm{a}_{4}$ | $[s \rightarrow s] \rightarrow[s]$ | $\mathrm{a}_{4}$ |
| $\mathrm{~A}_{3}$ | $[s \rightarrow s] \rightarrow([s \rightarrow s] \rightarrow[s]) \rightarrow[s]$ | $\lambda y w \cdot \mathrm{a}_{3}\left(\lambda x_{3} \cdot w\left(\lambda x_{4} \cdot y\left(\mathrm{c}_{3} x_{3} x_{4}\right)\right)\right)$ |
| $\mathrm{A}_{2}$ | $[s \rightarrow s] \rightarrow([s \rightarrow s] \rightarrow[s]) \rightarrow[s]$ | $\lambda y w \cdot \mathrm{a}_{2}\left(\lambda x_{2} \cdot w\left(\lambda v \cdot \mathrm{c}_{1}\left(y x_{2}\right) v\right)\right)$ |
| $\mathrm{A}_{1}$ | $([s \rightarrow s] \rightarrow[s]) \rightarrow[s]$ | $\lambda w \cdot \mathrm{a}_{1}\left(\lambda x_{1} \cdot w\left(\lambda v \cdot \mathrm{c}_{2} x_{1} v\right)\right)$ |

where $[s \rightarrow s]$ and $[s]$ are new abstract atomic types mapped to $s \rightarrow s$ and $s$, respectively. It is a routine to check that $\mathscr{L}^{\prime}(N)=M$ for

$$
N=\mathrm{A}_{1}\left(\lambda y_{1}^{[s \rightarrow s]} \cdot \mathrm{A}_{2} y_{1}\left(\lambda y_{2}^{[s \rightarrow s]} \cdot \mathrm{A}_{3} y_{2} \mathrm{a}_{4}\right)\right)
$$

## Elimination of Unary Nonlexical Constants

We start lexicalization of third-order semilexicalized ACGs with a modification of Lemma 4.24. While Lemmas 4.16 and 4.23 (elimination of nullary nonlexical constants) are good enough for third-order ACGs, Lemma 4.24 is not.

Lemma 4.32 (Elimination of Unary Nonlexical Constants). For every semilexicalized $A C G \mathscr{G} \in \mathbf{G}(3, n)$, we can find a semilexicalized $A C G \mathscr{G}^{\prime} \in$ $\mathbf{G}(3, n)$ such that $\mathscr{G}^{\prime}$ contains no nullary or unary nonlexical constants and

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\{R \in \mathcal{O}(\mathscr{G}) \mid R \text { contains a constant }\}
$$

Proof. We assume that $\mathscr{G}$ contains no nullary nonlexical constants and satisfies Condition II in Definition 4.12. $\mathcal{O}(\mathscr{G})$ contains no combinators.

We let an ACG $\mathscr{G}^{\prime}$ consist of the following lexical entries:

[^6]- $\left\langle\mathrm{c}, \tau_{0}(\mathrm{c}), \mathscr{L}(\mathrm{c})\right\rangle$ for all $\mathrm{c} \in \mathscr{C}_{0}^{-}$unless $\tau_{0}(\mathrm{c})$ is unary,
- $\left\langle\llbracket M \rrbracket, \tau_{0}(M), \mathscr{L}(M)\right\rangle$ for each $M$ of the form

$$
M=\lambda x_{1} \ldots x_{n} \cdot \mathrm{c}_{0}\left(\mathrm{a}\left(\lambda \vec{z}_{1} \cdot x_{1} M_{i, 1} \ldots M_{1, m_{1}}\right) \ldots\left(\lambda \vec{z}_{n} \cdot x_{n} M_{n, 1} \ldots M_{n, m_{n}}\right)\right)
$$

where $\vec{z}_{i}=z_{i, 1}^{p_{i, 1}} \ldots z_{i, m_{i}}^{p_{i, m_{i}}}, \mathrm{a} \in \mathscr{C}_{0}^{+}$,

$$
\begin{gathered}
\tau_{0}(\mathrm{a})=\left(p_{1,1} \rightarrow \cdots \rightarrow p_{1, m_{1}} \rightarrow p_{1}\right) \rightarrow \cdots \rightarrow\left(p_{n, 1} \rightarrow \cdots \rightarrow p_{n, m_{n}} \rightarrow p_{n}\right) \rightarrow p, \\
M_{i, j}=\left\{\begin{array}{ll}
\mathrm{c}_{i, j} z_{i, j} & \text { if } j \in I_{i} \\
z_{i, j} & \text { if } j \notin I_{i}
\end{array} \quad \text { for some } I_{i} \subseteq\left\{1, \ldots, m_{i}\right\},\right. \\
\mathrm{c}_{0}, \mathrm{c}_{i, j} \in \mathscr{C}_{0}^{-}, \quad \tau_{0}\left(\mathrm{c}_{0}\right)=p \rightarrow q, \quad \tau_{0}\left(\mathrm{c}_{i, j}\right)=p_{i, j} \rightarrow q_{i, j}, \\
\tau_{0}\left(x_{i}\right)=q_{i, 1} \rightarrow \cdots \rightarrow q_{i, m_{i}} \rightarrow p_{i} \text { where } p_{i, j}=q_{i, j} \text { if } j \notin I_{i} .
\end{gathered}
$$

Since the order of the type of the bound variables $x_{i}$ in the above $M$ is at most 2 , we have $\operatorname{ord}\left(\tau_{0}^{\prime}(\llbracket M \rrbracket)\right) \leq \max \left\{3, \operatorname{ord}\left(\tau_{0}(\mathrm{a})\right) \mid \mathrm{a} \in \mathscr{C}_{0}\right\}$. Thus, $\mathscr{G}^{\prime} \in \mathbf{G}(3, n)$ if $\mathscr{G} \in \mathbf{G}(3, n)$.

The inclusion $\mathcal{O}\left(\mathscr{G}^{\prime}\right) \subseteq \mathcal{O}(\mathscr{G})$ is trivial. We show the converse relation $\mathcal{O}(\mathscr{G}) \subseteq \mathcal{O}\left(\mathscr{G}^{\prime}\right)$.

We say that an occurrence of a unary nonlexical constant of type $p \rightarrow q$ is linked to an occurrence of a lexical constant if the path which starts from the negative occurrence of the subtype $p$ in the type $p \rightarrow q$ of the occurrence of the unary nonlexical constant ends in $p$ in the type of the occurrence of the lexical constant. We say that a term $M \in \Lambda\left(\Sigma_{0}\right)$ satisfies the linking condition if for every occurrence of a unary nonlexical constant, there is an occurrence of a lexical constant to which the occurrence of the unary nonlexical constant is linked. In fact, we can assume that every $M \in \mathcal{A}(\mathscr{G})$ satisfies the linking condition, as explained in the proof of Lemma 4.24. In the third-order case, when a unary nonlexical constant c is linked to a lexical constant a in some term in long normal form, their occurrences can be written either

- $\mathrm{c}(\mathrm{a} \vec{M})$ (direct link) or
- $\operatorname{a} \vec{M}(\lambda \vec{y} z . N[c z / x])$,
since the order of a is at most three. There cannot be two or more bound variables on which a path passes such as

$$
\mathrm{a}(\lambda y \cdot y(\lambda z . \mathrm{c} z)) .
$$

If the above term is well-typed, the order of a must be more than three.

Now, we show by induction on the size of $M$ that for every term $M \in$ $\Lambda\left(\Sigma_{0}\right)$ satisfying the linking condition, there is $P \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}(P)=$ $\tau_{0}(M)$ and $\mathscr{L}^{\prime}(P)=\mathscr{L}(M)$. We assume that $M$ is in long normal form.

If $M$ contains no constants, let $P=M \in \Lambda\left(\Sigma_{0}^{\prime}\right)$. Otherwise, it is enough to treat cases where $M$ is of the form

$$
M=\mathrm{b} M_{1} \ldots M_{n}
$$

where $b$ is either a lexical constant or a unary nonlexical constant. Other cases are trivial as in the proof of Lemma 4.23.

Case 1. Suppose that the head of $M$ is a unary nonlexical constant $\mathrm{c}_{0} \in \mathscr{C}_{0}^{-}$. By the linking condition, $M$ is of the form

$$
M=\mathrm{c}_{0}\left(\mathrm{a} M_{1} \ldots M_{n}\right)
$$

for some $a \in \mathscr{C}_{0}^{+}$of type

$$
\tau_{0}(\mathrm{a})=\left(p_{1,1} \rightarrow \cdots \rightarrow p_{1, m_{1}} \rightarrow p_{1}\right) \rightarrow \cdots \rightarrow\left(p_{n, 1} \rightarrow \cdots \rightarrow p_{n, m_{n}} \rightarrow p_{n}\right) \rightarrow p .
$$

We can assume that $M_{i}$ has the form $M_{i}=\lambda \vec{z}_{i} \cdot M_{i}^{\prime}$ with $\vec{z}_{i}=z_{i, 1}^{p_{i, 1}} \ldots z_{i, m_{i}}^{p_{i, m_{i}}}$. Let

$$
\begin{array}{r}
I_{i}=\left\{j \mid \text { some occurrence of a unary nonlexical constant } \mathrm{c}_{i, j}\right. \\
\text { is linked to a via } \left.p_{i, j}\right\} .
\end{array}
$$

Then, we can find $M_{i}^{\prime \prime}$ and $M_{i, j}$ such that

$$
\begin{aligned}
M_{i}^{\prime} & =M_{i}^{\prime \prime}\left[M_{i, j} / y_{i, j}\right]_{1 \leq j \leq m_{i}} \\
M_{i, j} & = \begin{cases}\mathrm{c}_{i, j} z_{i, j} & \text { if } j \in I_{i}, \\
z_{i, j} & \text { if } j \in\left\{1, \ldots, m_{i}\right\}-I_{i} .\end{cases}
\end{aligned}
$$

Clearly each $M_{i}^{\prime \prime}$ satisfies the linking condition. By the induction hypothesis, there is $P_{i}^{\prime}$ such that $\tau_{0}^{\prime}\left(P_{i}^{\prime}\right)=\tau_{0}\left(M_{i}^{\prime \prime}\right)$ and $\mathscr{L}^{\prime}\left(P_{i}^{\prime}\right)=\mathscr{L}\left(M_{i}^{\prime \prime}\right)$. By the definition of $\mathscr{G}^{\prime}$, there is a constant $\llbracket N \rrbracket \in \mathscr{C}_{0}^{\prime}$ for

$$
N=\lambda x_{1} \ldots x_{n} \cdot c_{0}\left(\mathrm{a}\left(\lambda \vec{z}_{1} \cdot x_{1} M_{1,1} \ldots M_{1, m_{1}}\right) \ldots\left(\lambda \vec{z}_{n} \cdot x_{n} M_{n, 1} \ldots M_{n, m_{n}}\right)\right) .
$$

Thus, for

$$
P=\llbracket N \rrbracket\left(\lambda y_{1,1} \ldots y_{1, m_{1}} \cdot P_{1}^{\prime}\right) \ldots\left(\lambda y_{n, 1} \ldots y_{n, m_{n}} \cdot P_{n}^{\prime}\right)
$$

we have

$$
\begin{aligned}
\mathscr{L}^{\prime}(P) & =\mathscr{L}(N)\left(\lambda y_{1,1} \ldots y_{1, m_{1}} \cdot \mathscr{L}\left(M_{1}^{\prime \prime}\right)\right) \ldots\left(\lambda y_{n, 1} \ldots y_{n, m_{n}} \cdot \mathscr{L}\left(M_{n}^{\prime \prime}\right)\right) \\
& =\mathscr{L}\left(\mathrm{c}_{0}\left(\mathrm{a}\left(\lambda \vec{z}_{1} \cdot M_{1}^{\prime \prime} M_{1,1} \ldots M_{1, m_{1}}\right) \ldots\left(\lambda \vec{z}_{n} \cdot M_{n}^{\prime \prime} M_{n, 1} \ldots M_{n, m_{n}}\right)\right)\right) \\
& =\mathscr{L}\left(\mathrm{c}_{0}\left(\mathrm{a}\left(\lambda \vec{z}_{1} \cdot M_{1}^{\prime}\right) \ldots\left(\lambda \vec{z}_{n} \cdot M_{n}^{\prime}\right)\right)\right) \\
& =\mathscr{L}(M) .
\end{aligned}
$$

Case 2. Suppose that the head of $M$ is a lexical constant $\mathrm{a} \in \mathscr{C}_{0}^{+}$and $M$ is of the form $M=\mathrm{a} M_{1} \ldots M_{n}$. Since $M$ is long normal form, $M$ has an atomic type $p$. Since $\mathscr{G}^{\prime}$ satisfies Condition II, there is $\mathrm{e}_{p} \in \mathscr{C}_{0}$ such that $\tau_{0}\left(\mathrm{e}_{p}\right)=p \rightarrow p$ and $\mathscr{L}\left(\mathrm{e}_{p}\right)=\lambda x . x$. By $\mathrm{e}_{p} M \approx M$, this case reduces to Case 1, where the induction hypothesis is applied to terms strictly smaller than $M$.

## Technical Definition and Lemma

As a preparation for elimination of nonlexical constants, we need some technical definition and lemma. These are necessary for building nonlexical parts in a term in the original abstract language. The reader can assume that the following second-order signature $\Sigma$ is the nonlexical part $\Sigma_{0}^{-}$of the abstract vocabulary of the given semilexicalized ACG.

Hereafter we denote the set of nullary free variables of $M$ by $\operatorname{Fv}^{0}(M)$. We say that a second-order type is multi-ary if it is $k$-ary for $k \geq 2$. Suppose that a second-order signature $\Sigma=\langle\mathscr{A}, \mathscr{C}, \tau\rangle$ is such that every constant has a multi-ary type. Let $k_{\Sigma}=\max \left\{k \mid \tau_{0}(\mathrm{c})=p_{1} \rightarrow \cdots \rightarrow p_{k} \rightarrow q, \mathrm{c} \in \mathscr{C}\right\}$. We define a subset $\Gamma(\Sigma)$ of $\Lambda(\Sigma)$ to consist of $\lambda$-terms $M$ such that

- $M$ has an atomic type,
- $M$ is built from variables of second-order types whose arities are strictly less than $k_{\Sigma}$ and constants of $\Sigma$,
- there is no subterm of the form $\left.y_{1}\left(y_{2} M^{\prime}\right)\right)$ for any two unary variables $y_{1}$ and $y_{2}$.

Formally $\Gamma(\Sigma)$ is defined by

$$
\begin{aligned}
\Gamma_{0}^{-}(\Sigma)= & \left\{z^{p} \mid p \in \mathscr{A}\right\}, \\
\Gamma_{n}(\Sigma)= & \Gamma_{n}^{-}(\Sigma) \cup\left\{y^{p \rightarrow q} M \mid M \in \Gamma_{n}^{-}(\Sigma) \text { and } \tau(M)=p\right\}, \\
\Gamma_{n+1}^{-}(\Sigma)= & \Gamma_{n}^{-}(\Sigma) \cup\left\{\mathrm{c} N_{1} \ldots N_{k} \mid N_{1}, \ldots, N_{k} \in \Gamma_{n}(\Sigma),\right. \\
& \left.\cup \tau_{0}(\mathrm{c})=\tau\left(N_{1}\right) \rightarrow \cdots \rightarrow \tau\left(N_{k}\right) \rightarrow q\right\} \\
& \cup\left\{x^{\tau\left(N_{1}\right) \rightarrow \cdots \rightarrow \tau\left(N_{k}\right) \rightarrow q} N_{1} \ldots N_{k} \mid N_{1}, \ldots, N_{k} \in \Gamma_{n}(\Sigma), 2 \leq k<k_{\Sigma}\right\}, \\
\Gamma(\Sigma)= & \bigcup_{n \geq 0} \Gamma_{n}(\Sigma) .
\end{aligned}
$$

We denote by $\Gamma^{\prime \prime}(\Sigma)$ the set of $\lambda$-terms obtained from elements of $\Gamma(\Sigma)$ by binding all the nullary variables in them. $\Gamma_{1}^{\prime \prime}(\Sigma)$ is the finite subset of $\Gamma^{\prime \prime}(\Sigma)$ whose elements contain at most one occurrence of a constant or multi-ary variable. Each element of $\Gamma^{\prime}(\Sigma)$ is obtained from an element of $\Gamma(\Sigma)$ by binding some nullary variables that have the same parent when we ignore unary variables, but the term rooted by that parent contains at least one nullary free variable. Note that if $M \in \Gamma^{\prime}(\Sigma)$, then the arity of $M$ is less than $k_{\Sigma}$.


Formally each of the above sets are defined as follows:
$\Gamma^{\prime \prime}(\Sigma)=\{\lambda \vec{z} \cdot N \mid N \in \Gamma(\Sigma), \vec{z}$ is a sequence of all the nullary variables in $N\}$, $\Gamma_{1}^{\prime \prime}(\Sigma)=\left\{\lambda \vec{z} \cdot M \in \Gamma^{\prime \prime}(\Sigma) \mid M \in \Gamma_{1}(\Sigma)\right\}$, $\Gamma^{\prime}(\Sigma)=\left\{\lambda z_{1} \ldots z_{k} \cdot M \mid M \in \Gamma(\Sigma), M=M^{\prime}\left[\times \vec{M}_{0} N_{1} \vec{M}_{1} \ldots N_{k} \vec{M}_{k} / z\right]\right.$,
$k \geq 1, \mathrm{x} \in \mathscr{C} \cup \mathscr{X}, \vec{M}_{0}, \ldots, \vec{M}_{k} \in \Gamma(\Sigma), N_{1}, \ldots, N_{k} \in \Gamma_{0}(\Sigma)$, $\left.\operatorname{Fv}^{0}\left(N_{i}\right)=\left\{z_{i}\right\}, \vec{M}_{0} \ldots \vec{M}_{k} \neq \varepsilon\right\}$.

Lemma 4.33. For a fixed $\lambda$-term $M \in \Gamma(\Sigma)$, let us partition the set $\operatorname{Fv}^{0}(M)$ of nullary free variables of $M$ into two disjoint subsets $Y$ and $Z$ such that $Y \neq \varnothing$. We can find $M^{\prime} \in \Gamma(\Sigma)$ and $M_{1}, \ldots, M_{m} \in \Gamma^{\prime}(\Sigma)$ such that

- $M=M^{\prime}\left[M_{i} / x_{i}\right]_{1 \leq i \leq m}$,
- $\operatorname{Fv}^{0}\left(M^{\prime}\right)=Y$ and $\bigcup_{1 \leq i \leq m} \operatorname{Fv}^{0}\left(M_{i}\right)=Z$.


Proof. Induction on $M$. If $M \in \Gamma_{0}(\Sigma)$, then $M=y z$ or $M=z$. Since $Y$ is not empty, $Y=\{z\}$ and $Z=\varnothing$. Letting $M^{\prime}=M$, we see that the lemma holds.

Otherwise, $M$ is of the form

$$
M=\left\{\begin{array}{l}
\mathrm{x} N_{1} \ldots N_{n} \\
y\left(\times N_{1} \ldots N_{n}\right)
\end{array} \quad\right. \text { or }
$$

where $2 \leq n \leq k_{\Sigma}$, x is a constant or a variable, $N_{i} \in \Gamma(\Sigma)$, and $y$ is a unary variable. Let $N_{0}$ be such that

$$
M=N_{0} N_{1} \ldots N_{n}
$$

( $N_{0}=\mathrm{x}$ or $N_{0}=\lambda z_{1} \ldots z_{n} \cdot y\left(\mathrm{x} z_{1} \ldots z_{n}\right)$ depending on the form of $\left.M\right)$. Let us partition $\{1, \ldots, n\}$ into two subsets $I$ and $J$ so that

$$
\begin{aligned}
& I=\left\{i \mid 1 \leq i \leq n, Y \cap \operatorname{Fv}^{0}\left(N_{i}\right) \neq \varnothing\right\} \\
& J=\left\{i \mid 1 \leq i \leq n, \operatorname{Fv}^{0}\left(N_{i}\right) \subseteq Z\right\} .
\end{aligned}
$$

For each $i \in I$, by applying the induction hypothesis to $N_{i}$, we get $N_{i}^{\prime} \in \Gamma(\Sigma)$ and $N_{i, 1}, \ldots, N_{i, m_{i}} \in \Gamma^{\prime}(\Sigma)$ such that

- $N_{i}=N_{i}^{\prime}\left[N_{i, j} / x_{i, j}\right]_{1 \leq j \leq m_{i}}$,
- $Y \cap \operatorname{Fv}^{0}\left(N_{i}\right)=\operatorname{Fv}^{0}\left(N_{i}^{\prime}\right)$ and $Z \cap \operatorname{Fv}^{0}\left(N_{i}\right)=\bigcup_{1 \leq j \leq m_{i}} \operatorname{Fv}^{0}\left(N_{i, j}\right)$.

We see $I \neq \varnothing$ by $Y \neq \varnothing$.
Case 1. $|I|=1$. Let $k$ be the unique integer such that $I=\{k\}$. Depending on the form of $N_{k}^{\prime} \in \Gamma(\Sigma)$, we have two subcases.

Case 1.1. The head of $N_{k}^{\prime}$ is $x_{k, j}$ for some $j \in\left\{1, \ldots, m_{k}\right\} . N_{k, j} \in \Gamma^{\prime}(\Sigma)$ can be represented as

$$
N_{k, j}=\lambda \vec{z}_{k, j} \cdot N_{k, j}^{\prime}
$$

for some $N_{k, j}^{\prime} \in \Gamma(\Sigma)$. Let

$$
\begin{aligned}
& M^{\prime}=N_{k}^{\prime} \\
& M_{i}= \begin{cases}N_{k, i} & \text { if } i \neq j, \\
\lambda z_{k, j} \cdot N_{0} N_{1} \ldots N_{k-1} N_{k, j}^{\prime} N_{k+1} \ldots N_{n} & \text { if } i=j\end{cases}
\end{aligned}
$$

for $1 \leq i \leq m_{k}$. Then, we have

- $M^{\prime} \in \Gamma(\Sigma)$ and $M_{i} \in \Gamma^{\prime}(\Sigma)$ for $1 \leq i \leq m_{k}$,
- $M=M^{\prime}\left[M_{i} / x_{k, i}\right]_{1 \leq i \leq m_{k}}$,
- $\operatorname{Fv}^{0}\left(M^{\prime}\right)=Y$ and $\bigcup_{1 \leq i \leq m_{k}} \operatorname{Fv}^{0}\left(M_{i}\right)=Z$.

Case 1.2. The head of $N_{k}^{\prime}$ is not $x_{k, j}$ for any $j \in\left\{1, \ldots, m_{k}\right\}$. Let
$N_{k}^{\prime \prime}= \begin{cases}N_{k}^{\prime \prime} & \text { if } N_{k}^{\prime} \text { is of the form } N_{k}^{\prime}=y^{\prime} N_{k}^{\prime \prime} \text { for some unary variable } y^{\prime} \\ N_{k}^{\prime} & \text { if the head of } N_{k}^{\prime} \text { is not a unary variable }\end{cases}$
$K= \begin{cases}y^{\prime} z & \text { if } N_{k}^{\prime} \text { is of the form } N_{k}^{\prime}=y^{\prime} N_{k}^{\prime \prime} \text { for some unary variable } y^{\prime} \\ z & \text { if the head of } N_{k}^{\prime} \text { is not a unary variable. }\end{cases}$
Then, $N_{k}^{\prime}=K\left[N_{k}^{\prime \prime} / z\right]$ and the head of $N_{k}^{\prime \prime}$ is not a unary variable. Let

$$
\begin{aligned}
& M^{\prime}=y_{0} N_{k}^{\prime \prime}, \\
& M_{0}=\lambda z \cdot N_{0} N_{1} \ldots N_{k-1} K N_{k+1} \ldots N_{n},
\end{aligned}
$$

where $y_{0}$ is a fresh unary variable of type $\tau\left(N_{k}^{\prime \prime}\right) \rightarrow \tau(M)$. We have

- $M^{\prime} \in \Gamma(\Sigma)$ and $M_{0} \in \Gamma^{\prime}(\Sigma)$,
- $M=M^{\prime}\left[M_{0} / y_{0}\right]\left[N_{k, i} / x_{k, i}\right]_{1 \leq i \leq m_{k}}$,
- $\operatorname{Fv}^{0}\left(M^{\prime}\right)=Y$ and $\operatorname{Fv}^{0}\left(M_{0}\right) \cup \bigcup_{1 \leq i \leq m_{k}} \operatorname{Fv}^{0}\left(N_{k, i}\right)=Z$.

Case 2. $1<|I|<n$. Let

$$
\begin{aligned}
& M^{\prime}=x_{0} \vec{N}_{I}^{\prime} \\
& M_{0}=\lambda \vec{z}_{I} \cdot N_{0} N_{1}^{\prime \prime} \ldots N_{n}^{\prime \prime} \\
& N_{i}^{\prime \prime}= \begin{cases}z_{i} & \text { if } i \in I \\
N_{i} & \text { if } i \notin I\end{cases}
\end{aligned}
$$

where $\vec{N}_{I}^{\prime}=\left\langle N_{i}^{\prime} \mid i \in I\right\rangle, \vec{z}_{I}=\left\langle z_{i} \mid i \in I\right\rangle$, and $\tau\left(x_{0}\right)=\tau\left(M_{0}\right)=\tau\left(\vec{N}_{I}^{\prime}\right) \rightarrow$ $\tau(M)$.

Note that since $1<|I|<n \leq k_{\Sigma}$, $x_{0}$ has neither a unary nor $k_{\Sigma}$-ary type. Therefore,

- $M^{\prime} \in \Gamma(\Sigma)$ and $M_{0} \in \Gamma^{\prime}(\Sigma)$,
- $M=M^{\prime}\left[M_{0} / x_{0}\right]\left[N_{i, j} / x_{i, j}\right]_{1 \leq j \leq m_{i}}^{i \in I}$,
- $\operatorname{Fv}^{0}\left(M^{\prime}\right)=Y$ and $\operatorname{Fv}^{0}\left(M_{0}\right) \cup \bigcup_{\substack{1 \leq j \leq m_{i}}} \operatorname{Fv}^{0}\left(N_{i, j}\right)=Z$.

Case 3. $|I|=n$. Let $M^{\prime}=N_{0} N_{1}^{\prime} \ldots N_{n}^{\prime} \in \Gamma(\Sigma)$. We have

- $M=M^{\prime}\left[N_{i, j} / x_{i, j}\right]_{\substack{1 \leq i \leq n \\ 1 \leq j \leq m_{i}}}^{\substack{ \\\text {, }}}$
- $\operatorname{Fv}^{0}\left(M^{\prime}\right)=Y$ and $\bigcup_{\substack{1 \leq i \leq n \\ 1 \leq j \leq m_{i}}} \mathrm{Fv}^{0}\left(N_{i, j}\right)=Z$.


## Elimination of Nonlexical Constants

Definition 4.34. Let $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle \in \mathbf{G}(3, n)$ be a third-order semilexicalized ACG that has no nullary or unary nonlexical constants. Let

$$
\mathscr{A}_{0}^{\prime}=\left\{[\gamma] \mid \gamma \text { is a second-order type with }|\gamma| \leq k_{\Sigma_{0}^{-}}\right\}
$$

$$
\text { where } k_{\Sigma_{0}^{-}}=\max \left\{\left|\tau_{0}(\mathrm{c})\right|-1 \mid \mathrm{c} \in \mathscr{C}_{0}^{-}\right\} .
$$

Let $[\vec{\alpha}] \rightarrow \beta$ denote $\left[\alpha_{1}\right] \rightarrow \cdots \rightarrow\left[\alpha_{m}\right] \rightarrow \beta$ if $\vec{\alpha}$ represents the sequence $\alpha_{1} \ldots \alpha_{m}$ of types in $\mathcal{T}\left(\mathscr{A}_{0}\right)$ and $\left[\alpha_{i}\right] \in \mathscr{A}_{0}^{\prime}$. Let $\mathscr{C}_{0}^{\prime}$ be the set of constants $\mathrm{A}_{K, M}$ for all $K$ and $M$ such that

- $K M$ is well-typed,
- $K \in \Gamma_{1}^{\prime \prime}\left(\Sigma_{0}^{-}\right)$,
- $M=\mathrm{a}\left(\lambda \vec{v}_{1} \cdot w_{1} \vec{M}_{1} \vec{M}_{1}^{\prime}\right) \ldots\left(\lambda \vec{v}_{m} \cdot w_{m} \vec{M}_{m} \vec{M}_{m}^{\prime}\right)$,
$-\mathrm{a} \in \mathscr{C}_{0}^{+}$of type $\left(\vec{p}_{1} \rightarrow p_{1}\right) \rightarrow \cdots \rightarrow\left(\vec{p}_{m} \rightarrow p_{m}\right) \rightarrow p$,
- $\vec{M}_{i}$ consists of elements of $\Gamma\left(\Sigma_{0}^{-}\right)$,
- $\vec{M}_{i}^{\prime}$ consists of elements of $\Gamma^{\prime}\left(\Sigma_{0}^{-}\right)$,
$-\vec{v}_{i}$ is a sequence of the nullary free variables in $\vec{M}_{i}$ and $\vec{M}_{i}^{\prime}$,
$-\tau_{0}\left(\vec{v}_{i}\right)=\vec{p}_{i}$,

$$
-\tau_{0}\left(w_{i}\right)=\tau_{0}\left(\vec{M}_{i}\right) \rightarrow \tau_{0}\left(\vec{M}_{i}^{\prime}\right) \rightarrow p_{i} .
$$

Let $\left\{y_{1}, \ldots, y_{l}\right\}=\operatorname{Fv}(K M)-\left\{w_{1}, \ldots, w_{m}\right\}$. We define $\tau_{0}^{\prime}$ and $\mathscr{L}^{\prime}$ as

- $\tau_{0}^{\prime}\left(\mathrm{A}_{K, M}\right)=\left[\tau_{0}\left(y_{1}\right)\right] \rightarrow \cdots \rightarrow\left[\tau_{0}\left(y_{l}\right)\right] \rightarrow \gamma_{1} \rightarrow \cdots \rightarrow \gamma_{m} \rightarrow\left[\tau_{0}(K M)\right]$, where $\gamma_{i}=\left[\tau_{0}\left(\vec{M}_{i}\right)\right] \rightarrow\left[\tau_{0}\left(\vec{M}_{i}^{\prime}\right)\right] \rightarrow\left[p_{i}\right]$,
- $\mathscr{L}^{\prime}([\gamma])=\gamma$ for all $[\gamma] \in \mathscr{A}_{0}^{\prime}$,
- $\mathscr{L}^{\prime}\left(\mathrm{A}_{K, M}\right)=\lambda y_{1} \ldots y_{l} w_{1} \ldots w_{m} . K M$.

An ACG $\mathscr{G}^{l}$ called the lexicalized form of $\mathscr{G}$ is defined as

$$
\mathscr{G}^{l}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L} \circ \mathscr{L}^{\prime},[s]\right\rangle \in \mathbf{G}(3, n+1)
$$

where $\Sigma_{0}^{\prime}=\left\langle\mathscr{A}_{0}^{\prime}, \mathscr{C}_{0}^{\prime}, \tau_{0}^{\prime}\right\rangle$.
We need to check the well-definedness of the above definition. First we check that $\tau_{0}^{\prime}\left(\mathrm{A}_{K, M}\right) \in \mathcal{T}\left(\mathscr{A}_{0}^{\prime}\right)$. Since every element of $\Gamma\left(\Sigma_{0}^{-}\right)$has an atomic type, and every element of $\Gamma^{\prime}\left(\Sigma_{0}^{-}\right)$has less than $k_{\Sigma_{0}^{--a r y}}$ type, we see $\gamma_{i} \in \mathcal{T}\left(\mathscr{A}_{0}^{\prime}\right)$. By the definition of $\Gamma\left(\Sigma_{0}^{-}\right)$and $\Gamma^{\prime}\left(\Sigma_{0}^{-}\right)$, the arities of variables $y_{1}, \ldots, y_{l}$ are less than $k_{\Sigma_{0}^{-}}$. Therefore, $\left[\tau_{0}\left(y_{j}\right)\right] \in \mathscr{A}_{0}^{\prime}$ and $\tau_{0}^{\prime}\left(\mathrm{A}_{K, M}\right) \in \mathcal{T}\left(\mathscr{A}_{0}^{\prime}\right)$.

Second we check that $\mathscr{C}_{0}^{\prime}$ is finite. Since $\Gamma_{1}^{\prime \prime}\left(\Sigma_{0}^{-}\right)$is a finite set, finitely many possible forms of $K$ exist. In $M$, the length of $\vec{v}_{i}$ for $1 \leq i \leq n$ is determined by the type of a. Since each element of $\vec{M}_{i}, \vec{M}_{i}^{\prime}$ has at least one element of $\vec{v}_{i}$ and each variable in $\vec{v}_{i}$ appears exactly once in some element of $\vec{M}_{i}, \vec{M}_{i}^{\prime}$, thus we see $\left|\vec{M}_{i} \vec{M}_{i}^{\prime}\right| \leq\left|\vec{v}_{i}\right|$. For each $M_{i, j}$ in $\vec{M}_{i}, \mathrm{Fv}^{0}\left(M_{i, j}\right)$ consists of some variables in $\vec{v}_{i}$ only. If $M_{i, j}$ contains no unary variables, then the size of $M_{i, j}$ is bounded by $\left|\mathrm{Fv}^{0}\left(M_{i, j}\right)\right| \leq\left|\vec{v}_{i}\right|$, otherwise, the size of $M_{i, j}$ is bounded by ${ }_{2}\left|\mathrm{Fv}^{0}\left(M_{i, j}\right)\right|$ by the restriction on occurrences of unary variables in elements of $\Gamma\left(\Sigma_{0}^{-}\right)$. Similarly, the size of $M_{i, j}^{\prime}$ in $\vec{M}_{i}^{\prime}$ is bounded by $2\left(\left|\mathrm{FV}^{0}\left(M_{i, j}^{\prime}\right)\right|+k_{\Sigma_{0}^{-}}\right) \leq$ $2\left(\left|\vec{v}_{i}\right|+k_{\Sigma_{0}^{-}}\right)$. Since the type of $w_{i}$ is determined uniquely by the types of a and elements of $\vec{M}_{i}, \vec{M}_{i}^{\prime}$, finitely many possible forms of $M$ exist.

By the construction, $\Sigma_{0}^{\prime}$ is third-order, and ord $\left(\mathscr{L} \circ \mathscr{L}^{\prime}\right) \leq \operatorname{ord}(\mathscr{L})+$ $\operatorname{ord}\left(\mathscr{L}^{\prime}\right)-1=\operatorname{ord}(\mathscr{L})+1$. Thus $\mathscr{G}^{l} \in \mathbf{G}(3, n+1)$ if $\mathscr{G} \in \mathbf{G}(3, n)$.

Definition 4.35. Let $M$ be an abstract $\lambda$-term of an atomic type of semilexicalized third-order ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$. We say that a $\beta$-expansion $\left(\lambda x_{1} \ldots x_{m} \cdot M_{0}\right) M_{1} \ldots M_{m}$ of a $\beta$-normal term $M \in \Lambda\left(\Sigma_{0}\right)$ is good if

- $M_{0} \in \Lambda\left(\Sigma_{0}^{+}\right)$,
- every variable in $M_{0}$ other than $x_{i}$ has an atomic type,
- $M_{i} \in \Gamma^{\prime \prime}\left(\Sigma_{0}^{-}\right)$for $i \geq 1$,
- [linking condition] every path from a negative occurrence of an atomic type in the type of $x_{i}$ ends in the type of an occurrence of a lexical constant in $M_{0}$.

If $M \in \mathcal{A}(\mathscr{G})$, then every variable in $M$ has an atomic type, since $\mathscr{G}$ is third-order. It is clear that $M$ has a good $\beta$-expansion.

Lemma 4.36. $\mathcal{O}(\mathscr{G})=\mathcal{O}\left(\mathscr{G}^{l}\right)$.
Proof. The inclusion $\mathcal{O}\left(\mathscr{G}^{l}\right) \subseteq \mathcal{O}(\mathscr{G})$ is easily seen. If $P \in \mathcal{A}\left(\mathscr{G}^{l}\right)$, then $\mathscr{L}^{\prime}(P) \in \mathcal{A}(\mathscr{G})$ and thus $\mathscr{L} \circ \mathscr{L}^{\prime}(P) \in \mathcal{O}(\mathscr{G})$. We show the converse $\mathcal{O}(\mathscr{G}) \subseteq \mathcal{O}\left(\mathscr{G}^{l}\right)$. Recall that every $M \in \mathcal{A}(\mathscr{G})$ has a good $\beta$-expansion. By induction on $M$, we show that if a $\beta$-normal term $M \in \Lambda\left(\Sigma_{0}\right)$ of an atomic type has a good $\beta$-expansion, then we have $P \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}(P)=\left[\tau_{0}(M)\right], \mathscr{L}^{\prime}(P)=M$, and moreover, for each free variable $y$ in $M$, $P$ has the corresponding variable of type $\left[\tau_{0}(y)\right] \in \mathscr{A}_{0}^{\prime}$.

Let $\left(\lambda \vec{x} . M_{0}\right) \vec{M}$ be a good $\beta$-expansion of $M . M_{0}$ has an atomic type. We have four cases depending on the head of $M_{0}$.

Case 1. The head of $M_{0}$ is a variable $z$ not in $\vec{x}$. By the definition, every variable appearing in $M_{0}$ other than elements of $\vec{x}$ has an atomic type. That is, $M_{0}=z^{p}$ for some $p \in \mathscr{A}_{0}$ and thus both $\vec{x}$ and $\vec{M}$ are empty. Let $P=z^{[p]}$.

Case 2. If the head of $M_{0}$ is a lexical constant a $\in \mathscr{C}_{0}^{+}$of type $\gamma_{1} \rightarrow$ $\cdots \rightarrow \gamma_{n} \rightarrow p$, then

$$
\begin{align*}
M & =M_{0}\left[M_{i, j} / x_{i, j}\right]_{1 \leq j \leq m_{i}}^{1 \leq i \leq n} \\
& =\mathrm{a}\left(\lambda \vec{v}_{1} . N_{1,0}\right) \ldots\left(\lambda \vec{v}_{n} . N_{n, 0}\right)\left[M_{i, j} / x_{i, j}\right]_{1 \leq j \leq m_{i}}^{1 \leq \leq \leq n} \\
& =\mathrm{a}\left(\lambda \vec{v}_{1} \cdot N_{1,0}\left[M_{1, j} / x_{1, j}\right]_{1 \leq j \leq m_{1}}\right) \ldots\left(\lambda \vec{v}_{n} . N_{n, 0}\left[M_{n, j} / x_{n, j}\right]_{1 \leq j \leq m_{n}}\right) \\
& =\mathrm{a} N_{1} \ldots N_{n} \tag{4.15}
\end{align*}
$$

where $\tau_{0}\left(N_{i, 0}\right) \in \mathscr{A}_{0}$ and $\left(\lambda \vec{x} \cdot M_{0}\right) \vec{M}=\left(\lambda\left\langle x_{i, j}\right\rangle_{1 \leq j \leq m_{i}}^{1 \leq i \leq n} . M_{0}\right)\left\langle M_{i, j}\right\rangle_{1 \leq j \leq m_{i}}^{1 \leq i \leq n}$. Suppose that $x_{i, j}$ appears in $N_{i, 0}$ as

$$
x_{i, j} \vec{L}_{i, j}
$$

where $\tau_{0}\left(x_{i, j} \vec{L}_{i, j}\right) \in \mathscr{A}_{0} .{ }^{7}$ Some elements of $\vec{L}_{i, j}$ are elements of $\vec{v}_{i}$ and some are not. Without loss of generality, we can assume that ${ }^{8}$

$$
x_{i, j} \vec{L}_{i, j}=x_{i, j} \vec{v}_{i, j} \vec{L}_{i, j}^{\prime}
$$

[^7]where every element of $\vec{v}_{i, j}$ is in $\vec{v}_{i}$ but no element of $\vec{L}_{i, j}^{\prime}$ is. Let $\vec{q}_{i, j}=\tau_{0}\left(\vec{v}_{i, j}\right)$ and $\vec{q}_{i, j}^{\prime}=\tau_{0}\left(\vec{L}_{i, j}^{\prime}\right)$. Thus, for
$$
\tau_{0}\left(x_{i, j}\right)=\vec{q}_{i, j} \rightarrow \vec{q}_{i, j}^{\prime} \rightarrow q_{i, j}=\tau_{0}\left(\vec{v}_{i, j}\right) \rightarrow \tau_{0}\left(\vec{L}_{i, j}^{\prime}\right) \rightarrow q_{i, j}
$$
every path starts from $\vec{q}_{i, j}$ ends in the type of the head occurrence of a in $M_{0}$, and no path starts from $\vec{q}_{i, j}^{\prime}$ does. Corresponding to this partition, let
$$
M_{i, j}=\lambda \vec{z}_{i, j} \vec{z}_{i, j}^{\prime} \cdot M_{i, j}^{-}
$$
where $\tau_{0}\left(\vec{z}_{i, j}\right)=\vec{q}_{i, j}, \tau_{0}\left(\vec{z}_{i, j}^{\prime}\right)=\vec{q}_{i, j}^{\prime}, \tau_{0}\left(M_{i, j}^{-}\right)=q_{i, j}$.
Let us partition $\left\{1, \ldots, m_{i}\right\}$ into two subsets
\[

$$
\begin{aligned}
I_{i} & =\left\{j \mid 1 \leq j \leq m_{i}, \vec{q}_{i, j}^{\prime} \neq \varepsilon\right\} \\
J_{i} & =\left\{j \mid 1 \leq j \leq m_{i}, \vec{q}_{i, j}^{\prime}=\varepsilon\right\}
\end{aligned}
$$
\]

For each $j \in I_{i}$, by applying Lemma 4.33 to $M_{i, j}^{-}$by regarding $\vec{z}_{i, j}$ as $Z$ and $\vec{z}_{i, j}^{\prime}$ as $Y$, we can find $k_{i j} \in \mathbb{N}, M_{i, j}^{\prime \prime}$ and $M_{i, j, h}$ for $1 \leq h \leq k_{i j}$ such that

- $M_{i, j}^{\prime \prime} \in \Gamma\left(\Sigma_{0}^{-}\right)$,
- $\vec{z}_{i, j}^{\prime}$ consists of exactly the members of $\mathrm{Fv}^{0}\left(M_{i, j}^{\prime \prime}\right)$,
- $M_{i, j, h} \in \Gamma^{\prime}\left(\Sigma_{0}^{-}\right)$for $1 \leq h \leq k_{i j}$,
- $\vec{z}_{i, j}=\vec{z}_{i, j, 1} \ldots \vec{z}_{i, j, k_{i j}}$ where $\vec{z}_{i, j, h}$ consists of exactly the members of $\operatorname{Fv}^{0}\left(M_{i, j, h}\right)$,
- $M_{i, j}^{-}=M_{i, j}^{\prime \prime}\left[M_{i, j, h} / y_{i, j, h}\right]_{1 \leq h \leq k_{i j}}$.

We let

$$
M_{i, j}^{\prime}=\lambda \vec{z}_{i, j}^{\prime} \cdot M_{i, j}^{\prime \prime} \in \Gamma^{\prime \prime}\left(\Sigma_{0}^{-}\right)
$$

so

$$
\begin{align*}
M_{i, j} & =\lambda \vec{z}_{i, j} \vec{z}_{i, j}^{\prime} \cdot M_{i, j}^{-} \\
& =\lambda \vec{z}_{i, j} \vec{z}_{i, j}^{\prime} \cdot M_{i, j}^{\prime \prime}\left[M_{i, j, h} / y_{i, j, h}\right]_{1 \leq h \leq k_{i j}} \\
& =\lambda \vec{z}_{i, j} \cdot M_{i, j}^{\prime}\left[M_{i, j, h} / y_{i, j, h}\right]_{1 \leq h \leq k_{i j}} \tag{4.16}
\end{align*}
$$

for $j \in I_{i}$.
Let $N_{i, 0}^{\prime}$ be obtained from $N_{i, 0}$ by replacing $x_{i, j} \vec{v}_{i, j}$ with $x_{i, j}^{\prime}$, i.e.,

$$
\begin{equation*}
N_{i, 0}=N_{i, 0}^{\prime}\left[x_{i, j} \vec{v}_{i, j} / x_{i, j}^{\prime}\right]_{1 \leq j \leq m_{i}}, \tag{4.17}
\end{equation*}
$$

where $\tau_{0}\left(x_{i, j}^{\prime}\right)=\vec{q}_{i, j}^{\prime} \rightarrow q_{i, j}$. If $j \in J_{i}$, then $\tau_{0}\left(x_{i, j}^{\prime}\right)=q_{i, j} \in \mathscr{A}_{0}$. Let

$$
\begin{equation*}
N_{i}^{\prime}=N_{i, 0}^{\prime}\left[M_{i, j}^{\prime} / x_{i, j}^{\prime}\right]_{j \in I_{i}} \tag{4.18}
\end{equation*}
$$

We show that $\left(\lambda\left\langle x_{i, j}^{\prime}\right\rangle_{j \in I_{i}} . N_{i, 0}^{\prime}\right)\left\langle M_{i, j}^{\prime}\right\rangle_{j \in I_{i}}$ is a good $\beta$-expansion of $N_{i}^{\prime}$. By $M_{0} \in \Lambda\left(\Sigma_{0}^{+}\right), N_{i, 0} \in \Lambda\left(\Sigma_{0}^{+}\right)$, thus $N_{i, 0}^{\prime} \in \Lambda\left(\Sigma_{0}^{+}\right)$, and $\tau_{0}\left(N_{i, 0}^{\prime}\right)=\tau_{0}\left(N_{i, 0}\right) \in \mathscr{A}_{0}$. All variables appearing in $N_{i, 0}$ other than $x_{i, j}^{\prime}$ appears in $M_{0}, N_{i, 0}$, so they have atomic types, and $x_{i, j}^{\prime}$ with $j \in J_{i}$ also have atomic types. $M_{i, j}^{\prime} \in \Gamma^{\prime \prime}\left(\Sigma_{0}^{-}\right)$ by its construction. Recall that for $\tau_{0}\left(x_{i, j}\right)=\vec{q}_{i, j} \rightarrow \vec{q}_{i, j}^{\prime} \rightarrow q_{i, j}$, all occurrences of atomic types in $\vec{q}_{i, j}$ are linked to the head occurrence of a in $M_{0}$, and all occurrences of atomic types in $\vec{q}_{i, j}^{\prime}$ are linked to some other lexical constants in $M_{0}$, i.e., those lexical constants are in $N_{i, 0}$ and thus in $N_{i, 0}^{\prime}$. This implies that, for $\tau_{0}\left(x_{i, j}^{\prime}\right)=\vec{q}_{i, j}^{\prime} \rightarrow q_{i, j}$, all occurrences of atomic types in $\vec{q}_{i, j}^{\prime}$ are linked to those lexical constants in $N_{i, 0}^{\prime}$. Therefore, $\left(\lambda\left\langle x_{i, j}^{\prime}\right\rangle_{j \in I_{i}} . N_{i, 0}^{\prime}\right)\left\langle M_{i, j}^{\prime}\right\rangle_{j \in I_{i}}$ satisfies the linking condition and thus is a good $\beta$-expansion of $N_{i}^{\prime}$. By applying the induction hypothesis to $N_{i}^{\prime}$, we get $P_{i}^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}\left(P_{i}^{\prime}\right)=\left[\tau_{0}\left(N_{i}^{\prime}\right)\right]$,

$$
\begin{equation*}
\mathscr{L}^{\prime}\left(P_{i}^{\prime}\right)=N_{i}^{\prime} \tag{4.19}
\end{equation*}
$$

and each free variable $y$ of $N_{i}^{\prime}$ has the type $\left[\tau_{0}(y)\right] \in \mathscr{A}_{0}^{\prime}$ in $P_{i}^{\prime}$. Let

- $\vec{v}_{i}^{\prime}$ be the subsequence of $\vec{v}_{i}$ consisting of variables not in $\vec{v}_{i, j}$ for any $j$,
- $\vec{x}_{i}^{\prime}$ be the sequence of variables $x_{i, j}^{\prime}$ for $j \in J_{i}$,
- $\vec{y}_{i}$ be the sequence of variables $y_{i, j, h}$ for $j \in I_{i}, 1 \leq h \leq k_{i j}$.

Indeed the variables constituting above sequences occur free in $N_{i}^{\prime}$. Let

$$
\begin{equation*}
P_{i}=\lambda \vec{v}_{i}^{\prime} \vec{x}_{i}^{\prime} \vec{y}_{i} \cdot P_{i}^{\prime} \tag{4.20}
\end{equation*}
$$

where

$$
\tau_{0}^{\prime}\left(P_{i}\right)=\left[\tau_{0}\left(\vec{v}_{i}^{\prime}\right)\right] \rightarrow\left[\tau_{0}\left(\vec{x}_{i}^{\prime}\right)\right] \rightarrow\left[\tau_{0}\left(\vec{y}_{i}\right)\right] \rightarrow\left[\tau_{0}\left(N_{i}^{\prime}\right)\right]
$$

On the other hand, the following equation holds:

$$
\begin{array}{rlr}
N_{i} & =\lambda \vec{v}_{i} \cdot N_{i, 0}\left[M_{i, j} / x_{i, j}\right]_{1 \leq j \leq m_{i}} \\
& =\lambda \vec{v}_{i} \cdot N_{i, 0}^{\prime}\left[x_{i, j} \vec{v}_{i, j} / x_{i, j}^{\prime}\right]_{1 \leq j \leq m_{i}}\left[M_{i, j} / x_{i, j}\right]_{1 \leq j \leq m_{i}} \\
& =\lambda \vec{v}_{i} \cdot N_{i, 0}^{\prime}\left[x_{i, j} \vec{v}_{i, j} / x_{i, j}^{\prime}\right]_{1 \leq j \leq m_{i}}\left[\lambda \vec{z}_{i, j} \cdot M_{i, j}^{\prime}\left[M_{i, j, h} / y_{i, j, h}\right]_{1 \leq h \leq k_{i j}} / x_{i, j}\right]_{j \in I_{i}} \\
\qquad & \quad\left[M_{i, j} / x_{i, j}\right]_{j \in J_{i}} \\
& =\lambda \vec{v}_{i} \cdot N_{i, 0}^{\prime}\left[M_{i, j}^{\prime}\left[M_{i, j, h}\left[\vec{v}_{i, j, h} / \vec{z}_{i, j, h}\right] / y_{i, j, h}\right]_{1 \leq h \leq k_{i j}} / x_{i, j}^{\prime}\right]_{j \in I_{i}}\left[M_{i, j} \vec{v}_{i, j} / x_{i, j}^{\prime}\right]_{j \in J_{i}} \\
& =\lambda \vec{v}_{i} \cdot N_{i}^{\prime}\left[M_{i, j, h}\left[\vec{v}_{i, j, h} / \vec{z}_{i, j, h}\right] / y_{i, j, h}\right]_{j \in I_{i}}^{1 \leq k_{i j}}\left[M_{i, j} \vec{v}_{i, j} / x_{i, j}^{\prime}\right]_{j \in J_{i}} \quad \text { (by (4.18} \tag{4.18}
\end{array}
$$

where $\vec{v}_{i, j, h}$ is the subsequence of $\vec{v}_{i, j}$ corresponding to the subsequence $\vec{z}_{i, j, h}$ of $\vec{z}_{i, j}$. Letting

$$
\begin{aligned}
\vec{M}_{i} & \left.=\left.\langle | M_{i, j} \vec{v}_{i, j}\right|_{\beta}\right\rangle_{j \in J_{i}} \\
\vec{M}_{i}^{\prime} & =\left\langle M_{i, j, h}\left[\vec{v}_{i, j, h} / \vec{z}_{i, j, h}\right]\right]_{j \in I_{i}}^{1 \leq h \leq k_{i j}}
\end{aligned}
$$

we have

$$
\begin{equation*}
N_{i}=\lambda \vec{v}_{i} \cdot N_{i}^{\prime}\left[\vec{M}_{i} / \vec{x}_{i}^{\prime}\right]\left[\vec{M}_{i}^{\prime} / \vec{y}_{i}\right] \tag{4.21}
\end{equation*}
$$

Note that $\left|M_{i, j} \vec{v}_{i, j}\right|_{\beta} \in \Gamma\left(\Sigma_{0}^{-}\right)$for $j \in J_{i}$, and $M_{i, j, h}\left[\vec{v}_{i, j, h} / \vec{z}_{i, j, h}\right] \in \Gamma^{\prime}\left(\Sigma_{0}^{-}\right)$for $j \in I_{i}$. Let

$$
L=\lambda \vec{y} w_{1} \ldots w_{n} \cdot \mathrm{a}\left(\lambda \vec{v}_{1} \cdot w_{1} \vec{v}_{1}^{\prime} \vec{M}_{1} \vec{M}_{1}^{\prime}\right) \ldots\left(\lambda \vec{v}_{n} \cdot w_{n} \vec{v}_{n}^{\prime} \vec{M}_{n} \vec{M}_{n}^{\prime}\right)
$$

where $\vec{y}$ consists of free variables of $\vec{M}_{i}$ and $\vec{M}_{i}^{\prime}$ other than elements of $\vec{v}_{i}$. Since $K=\lambda z . z \in \Gamma_{1}^{\prime \prime}\left(\Sigma_{0}\right)$, by the definition, we have a constant $\mathrm{A}_{K, L} \in \mathscr{C}_{0}^{\prime}$ such that

$$
\begin{aligned}
\tau_{0}^{\prime}\left(\mathrm{A}_{K, L}\right) & =\left[\tau_{0}(\vec{y})\right] \rightarrow \tau_{0}^{\prime}\left(P_{1}\right) \rightarrow \cdots \rightarrow \tau_{0}^{\prime}\left(P_{n}\right) \rightarrow[p] \\
\mathscr{L}^{\prime}\left(\mathrm{A}_{K, L}\right) & =L
\end{aligned}
$$

Thus, for $P=\mathrm{A}_{K, L} \vec{y} P_{1} \ldots P_{n} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$, we have

$$
\begin{align*}
\mathscr{L}^{\prime}(P) & =\mathscr{L}^{\prime}\left(\mathrm{A}_{K, L} \vec{y} P_{1} \ldots P_{n}\right) \\
& =\mathscr{L}^{\prime}\left(\mathrm{A}_{K, L} \vec{y}\right)\left(\lambda \vec{v}_{1}^{\prime} \vec{x}_{1}^{\prime} \vec{y}_{1} \cdot N_{1}^{\prime}\right) \ldots\left(\lambda \vec{v}_{n}^{\prime} \vec{x}_{n}^{\prime} \vec{y}_{n} \cdot N_{n}^{\prime}\right) \quad(\text { by }(4.19),(4.20)) \\
& =\mathrm{a}\left(\lambda \vec{v}_{1} \cdot N_{1}^{\prime}\left[\vec{M}_{1} / \vec{x}_{1}^{\prime}\right]\left[\vec{M}_{1}^{\prime} / \vec{y}_{1}\right]\right) \ldots\left(\lambda \vec{v}_{n} \cdot N_{n}^{\prime}\left[\vec{M}_{n} / \vec{x}_{n}^{\prime}\right]\left[\vec{M}_{n}^{\prime} / \vec{y}_{n}\right]\right) \\
& =\mathrm{a} N_{1} \ldots N_{n}  \tag{4.21}\\
& =M
\end{align*}
$$

Case 3. Suppose that $M_{0}=x_{1} M_{0}^{\prime}$ for a unary variable $x_{1}$ in $\vec{x}$. Then the head of $M_{0}^{\prime}$ is a lexical constant a by the linking condition. Let $M_{1}$ be the corresponding term in $\vec{M}$ to $x_{1}$. Since $x_{1}$ and $M_{1}$ are unary, either $M_{1}=\lambda z . z$ or $M_{1}=\lambda z \cdot y_{0} z$ for some unary variable $y_{0}$. Let $\overrightarrow{M^{\prime}}$ be $\vec{M}$ minus $M_{1}$ and $\vec{x}^{\prime}$ be $\vec{x}$ minus $x_{1}$.

If $M_{1}=\lambda z . z$, then this implies that $M$ has a good $\beta$-expansion $\left(\lambda \vec{x}^{\prime} \cdot M_{0}^{\prime}\right) \overrightarrow{M^{\prime}}$. This case reduces to Case 2.

Suppose that $M_{1}=\lambda z . y_{0} z . M$ can be represented as

$$
\begin{equation*}
M=x_{1} M_{0}^{\prime}\left[M_{1} / x_{1}, \vec{M}^{\prime} / \vec{x}^{\prime}\right]=y_{0} M_{0}^{\prime}\left[\vec{M}^{\prime} / \vec{x}^{\prime}\right]=y_{0}\left(\mathrm{a} N_{1} \ldots N_{n}\right) \tag{4.22}
\end{equation*}
$$

Applying exactly the same discussion as Case 2 to $M_{0}^{\prime}\left[\vec{M}^{\prime} / \vec{x}^{\prime}\right]$, for each $i \in$ $\{1, \ldots, n\}$, we get a $\lambda$-term $N_{i}^{\prime}$ and sequences of $\lambda$-terms $\vec{M}_{i}$ and $\vec{M}_{i}^{\prime}$ such that
$N_{i}^{\prime}$ has a good $\beta$-expansion, each element of $\vec{M}_{i}$ is in $\Gamma\left(\Sigma_{0}^{-}\right)$, each element of $\vec{M}_{i}^{\prime}$ is in $\Gamma^{\prime}\left(\Sigma_{0}^{-}\right)$, and

$$
\begin{equation*}
N_{i}=\lambda \vec{v}_{i} \cdot N_{i}^{\prime}\left[\vec{M}_{i} / \vec{x}_{i}^{\prime}\right]\left[\vec{M}_{i}^{\prime} / \vec{y}_{i}\right] \tag{4.23}
\end{equation*}
$$

and moreover, we have $P_{i} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that

$$
\begin{equation*}
\mathscr{L}^{\prime}\left(P_{i}\right)=\lambda \vec{v}_{i}^{\prime} \vec{x}_{i}^{\prime} \vec{y}_{i} \cdot N_{i}^{\prime} \tag{4.24}
\end{equation*}
$$

where $\vec{v}_{i}^{\prime}$ consists of variables in $\vec{v}_{i}$ which appear in neither $\vec{M}_{i}$ nor $\vec{M}_{i}^{\prime}$, and every free variable $y$ in $N_{i}^{\prime}$ has type $\left[\tau_{0}(y)\right] \in \mathscr{A}_{0}^{\prime}$ in $P_{i}$. Since $M_{1}=\lambda z . y_{0} z \in$ $\Gamma_{1}^{\prime \prime}\left(\Sigma_{0}^{-}\right)$, by the definition, we have a constant B such that

$$
\begin{align*}
\tau_{0}^{\prime}(\mathrm{B}) & =\left[\tau_{0}(\vec{y})\right] \rightarrow\left[\tau_{0}\left(y_{0}\right)\right] \rightarrow \tau_{0}^{\prime}\left(P_{1}\right) \rightarrow \cdots \rightarrow \tau_{0}^{\prime}\left(P_{n}\right) \rightarrow q \\
\mathscr{L}^{\prime}(\mathrm{B}) & =\lambda \vec{y} y_{0} w_{1} \ldots w_{n} \cdot y_{0}\left(\mathrm{a}\left(\lambda \vec{v}_{1} \cdot w_{1} \vec{v}_{1}^{\prime} \vec{M}_{1} \vec{M}_{1}^{\prime}\right) \ldots\left(\lambda \vec{v}_{n} \cdot w_{n} \vec{v}_{n}^{\prime} \vec{M}_{n} \vec{M}_{n}^{\prime}\right)\right) \tag{4.25}
\end{align*}
$$

where $\vec{y}$ consists of free variables of $\vec{M}_{i}$ or $\vec{M}_{i}^{\prime}$ other than elements of $\vec{v}_{i}$. Thus, for

$$
P=\mathrm{B} \vec{y} y_{0} P_{1} \ldots P_{n},
$$

we have

$$
\begin{align*}
\mathscr{L}^{\prime}(P) & =\mathscr{L}^{\prime}(\mathrm{B}) \vec{y} y_{0}\left(\lambda \vec{v}_{1}^{\prime} \vec{x}_{1}^{\prime} \vec{y}_{1} \cdot N_{1}^{\prime}\right) \ldots\left(\lambda \vec{v}_{n}^{\prime} \vec{x}_{n}^{\prime} \vec{y}_{n} \cdot N_{n}^{\prime}\right)  \tag{4.24}\\
& =y_{0}\left(\mathrm{a}\left(\lambda \vec{v}_{1} \cdot N_{1}^{\prime}\left[\vec{M}_{1} / \vec{x}_{1}^{\prime}\right]\left[\vec{M}_{1}^{\prime} / \vec{y}_{1}\right]\right) \ldots\left(\lambda \vec{v}_{n} \cdot N_{n}^{\prime}\left[\vec{M}_{n} / \vec{x}_{n}^{\prime}\right]\left[\vec{M}_{n}^{\prime} / \vec{y}_{n}\right]\right)\right)  \tag{4.25}\\
& =y_{0}\left(\mathrm{a} N_{1} \ldots N_{n}\right)  \tag{4.23}\\
& =M . \tag{4.22}
\end{align*}
$$

Case 4. Suppose that the head of $M_{0}$ is a variable $x_{1}$ in $\vec{x}$ and $x_{1}$ is a $k$-ary variable for some $k \geq 2$. Let the corresponding term $M_{1}$ in $\vec{M}$ be $M_{1}=\lambda z_{1} \vec{z} \cdot M_{1}^{\prime \prime}$ with $\left|z_{1} \vec{z}\right|=k$ and $\tau_{0}\left(M_{1}^{\prime \prime}\right) \in \mathscr{A}_{0}$. By applying Lemma 4.33 to $M_{1}^{\prime \prime}$ with respect to $Y=\operatorname{Fv}^{0}\left(M_{1}^{\prime \prime}\right)-\left\{z_{1}\right\} \neq \varnothing$ and $Z=\left\{z_{1}\right\}$, we can find $M_{1}^{\prime \prime \prime} \in \Gamma\left(\Sigma_{0}^{-}\right)$and $K \in \Gamma^{\prime}\left(\Sigma_{0}^{-}\right)$such that $M_{1}^{\prime \prime}=M_{1}^{\prime \prime \prime}\left[K / y_{0}\right] . \operatorname{By}\left|\mathrm{Fv}^{0}(K)\right|=1$, $\lambda z_{1} \cdot K \in \Gamma_{1}^{\prime \prime}\left(\Sigma_{0}^{-}\right)$. Let $M_{1}^{\prime}=\lambda \vec{z} \cdot M_{1}^{\prime \prime \prime} \in \Gamma^{\prime \prime}\left(\Sigma_{0}^{-}\right)$. We have

$$
\begin{equation*}
M_{1}=\lambda z_{1} \cdot M_{1}^{\prime}\left[K / y_{0}\right] . \tag{4.26}
\end{equation*}
$$

Let $M_{0}=x_{1} L_{1} \ldots L_{k}$ (recall that $M_{0}$ has an atomic type) and $\vec{x}$ divide into three subsequences $x_{1}, \vec{x}^{\prime}, \vec{x}^{\prime \prime}$ so that $\vec{x}^{\prime}$ consists of variables in $L_{1}, \vec{x}^{\prime \prime}$ consists of variables in $L_{j}$ for some $j \geq 2$. Corresponding to this division, $\vec{M}$
is divided into three subsequences $M_{1}, \overrightarrow{M^{\prime}}$, and $\overrightarrow{M^{\prime \prime}} . M$ can be represented as

$$
\begin{align*}
M & =M_{0}[\vec{M} / \vec{x}] \\
& =x_{1} L_{1} L_{2} \ldots L_{k}\left[M_{1} / x_{1}, \overrightarrow{M^{\prime}} / \vec{x}^{\prime}, \overrightarrow{M^{\prime \prime}} / \vec{x}^{\prime \prime}\right] \\
& =M_{1} L_{1} L_{2} \ldots L_{k}\left[\overrightarrow{M^{\prime}} / \vec{x}^{\prime}\right]\left[\overrightarrow{M^{\prime \prime}} / \vec{x}^{\prime \prime}\right] \\
& =M_{1}^{\prime}\left[K\left[L_{1}\left[\vec{M}^{\prime} / \vec{x}^{\prime}\right] / z_{1}\right] / y_{0}\right] L_{2} \ldots L_{k}\left[\vec{M}^{\prime \prime} / \vec{x}^{\prime \prime}\right] \tag{4.27}
\end{align*}
$$

by (4.26). Let

$$
\begin{align*}
& M^{\prime}=M_{1}^{\prime} L_{2} \ldots L_{k}\left[\vec{M}^{\prime \prime} / \vec{x}^{\prime \prime}\right]  \tag{4.28}\\
& M_{0}^{\prime}=x_{1}^{\prime} L_{2} \ldots L_{k} \quad\left(\text { thus } M_{0}=M_{0}^{\prime}\left[x_{1} L_{1} / x_{1}^{\prime}\right]\right)
\end{align*}
$$

where $\tau_{0}\left(x_{1}^{\prime}\right)=\tau_{0}\left(M_{1}^{\prime}\right)$. By (4.27) and (4.28),

$$
\begin{equation*}
M=M^{\prime}\left[K\left[L_{1}\left[\vec{M}^{\prime} / \vec{x}^{\prime}\right] / z_{1}\right] / y_{0}\right] . \tag{4.29}
\end{equation*}
$$

By $K \in \Gamma^{\prime}\left(\Sigma_{0}^{-}\right)$, it contains at least one occurrence of a lexical constant or a variable of $l$-ary with $l \geq 2$. Therefore, the size of $M^{\prime}$ is strictly smaller than $M_{\text {. }}$ It is not difficult to see that $M^{\prime}$ has a good $\beta$-expansion $\left(\lambda x_{1}^{\prime} \vec{x}^{\prime \prime} \cdot M_{0}^{\prime}\right) M_{1}^{\prime} \vec{M}^{\prime \prime}$, since $\left(\lambda \vec{x} . M_{0}\right) \vec{M}$ is a good $\beta$-expansion of $M$. By the induction hypothesis, we have $P^{\prime} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that $\tau_{0}^{\prime}\left(P^{\prime}\right)=\left[\tau_{0}\left(M^{\prime}\right)\right]$,

$$
\begin{equation*}
\mathscr{L}^{\prime}\left(P^{\prime}\right)=M^{\prime} \tag{4.30}
\end{equation*}
$$

and $\tau_{0}^{\prime}(y)=\left[\tau_{0}(y)\right] \in \mathscr{A}_{0}^{\prime}$ for every free variable $y$ in $M^{\prime}$. In particular, $\tau_{0}^{\prime}\left(y_{0}\right)=\left[\tau_{0}\left(y_{0}\right)\right]$.

By the linking condition on $x_{1}$, the head of $L_{j}$ is a lexical constant for each $j \in\{1, \ldots, k\}$. Particularly for $j=1$, there are $N_{1,0}, \ldots, N_{n, 0}$ such that $L_{1}$ can be written as

$$
L_{1}=\mathrm{a}\left(\lambda \vec{v}_{1} \cdot N_{1,0}\right) \ldots\left(\lambda \vec{v}_{n} \cdot N_{n, 0}\right) .
$$

Letting $N_{1}, \ldots, N_{n}$ be such that

$$
\begin{equation*}
L_{1}\left[\vec{M}^{\prime} / \vec{x}^{\prime}\right]=\mathrm{a} N_{1} \ldots N_{n} \tag{4.31}
\end{equation*}
$$

we have

$$
\begin{align*}
M & =M^{\prime}\left[K\left[L_{1}\left[\overrightarrow{M^{\prime}} / \vec{x}^{\prime}\right] / z_{1}\right] / y_{0}\right] \\
& =M^{\prime}\left[K\left[\mathrm{a} N_{1} \ldots N_{n} / z_{1}\right] / y_{0}\right] \tag{4.32}
\end{align*}
$$

by (4.29) and (4.31). It is not difficult to see that $\left(\lambda \vec{x}^{\prime} . L_{1}\right) \vec{M}^{\prime}$ is a good $\beta$ expansion of $L_{1}\left[\overrightarrow{M^{\prime}} / \vec{x}^{\prime}\right]=\mathrm{a} N_{1} \ldots N_{n}$, since $\left(\lambda \vec{x} \cdot M_{0}\right) \vec{M}$ is a good $\beta$-expansion of $M$. Applying exactly the same discussion as Case 2 to $L_{1}\left[\vec{M}^{\prime} / \vec{x}^{\prime}\right]$, for each $i \in\{1, \ldots, n\}$, we have a $\lambda$-term $N_{i}^{\prime}$ and sequences $\vec{M}_{i}$ and $\vec{M}_{i}^{\prime}$ of $\lambda$-terms such that $N_{i}^{\prime}$ has a good $\beta$-expansion, each element of $\vec{M}_{i}$ is in $\Gamma\left(\Sigma_{0}^{-}\right)$, each element of $\overrightarrow{M_{i}^{\prime}}$ is in $\Gamma^{\prime}\left(\Sigma_{0}^{-}\right)$, and

$$
\begin{equation*}
N_{i}=\lambda \vec{v}_{i} \cdot N_{i}^{\prime}\left[\vec{M}_{i} / \vec{x}_{i}^{\prime}\right]\left[\vec{M}_{i}^{\prime} / \vec{y}_{i}\right] \tag{4.33}
\end{equation*}
$$

and moreover, we have $P_{i} \in \Lambda\left(\Sigma_{0}^{\prime}\right)$ such that

$$
\begin{equation*}
\mathscr{L}^{\prime}\left(P_{i}\right)=\lambda \vec{v}_{i}^{\prime} \vec{x}_{i}^{\prime} \vec{y}_{i} \cdot N_{i}^{\prime} \tag{4.34}
\end{equation*}
$$

where $\vec{v}_{i}^{\prime}$ consists of variables in $\vec{v}_{i}$ which appear in neither $\vec{M}_{i}$ nor $\vec{M}_{i}^{\prime}$, and every free variable $y$ in $N_{i}^{\prime}$ has type $\left[\tau_{0}(y)\right]$ in $P_{i}$. By the definition, we have a constant C such that

$$
\begin{align*}
\tau_{0}^{\prime}(\mathrm{C}) & =\left[\tau_{0}(\vec{y})\right] \rightarrow \tau_{0}^{\prime}\left(P_{1}\right) \rightarrow \cdots \rightarrow \tau_{0}^{\prime}\left(P_{n}\right) \rightarrow\left[\tau_{0}(K)\right] \\
\mathscr{L}^{\prime}(\mathrm{C}) & =\lambda \vec{y} w_{1} \ldots w_{n} \cdot K\left[\mathrm{a}\left(\lambda \vec{v}_{1} \cdot w_{1} \vec{v}_{1}^{\prime} \vec{M}_{1} \vec{M}_{1}^{\prime}\right) \ldots\left(\lambda \vec{v}_{n} \cdot w_{n} \vec{v}_{n}^{\prime} \vec{M}_{n} \vec{M}_{n}^{\prime}\right) / z_{1}\right] \tag{4.35}
\end{align*}
$$

where $\vec{y}$ consists of free variables of $K, \vec{M}_{i}$, or $\vec{M}_{i}^{\prime}$ other than $z_{1}, \vec{v}_{i}$. Thus, for

$$
P=P^{\prime}\left[\mathrm{C} \vec{y} P_{1} \ldots P_{n} / y_{0}\right],
$$

we have

$$
\begin{array}{rlrl}
\mathscr{L}^{\prime}(P) & =M^{\prime}\left[\mathscr{L}^{\prime}(\mathrm{C}) \vec{y}\left(\lambda \vec{v}_{1}^{\prime} \vec{x}_{1}^{\prime} \vec{y}_{1} \cdot N_{1}^{\prime}\right) \ldots\left(\lambda \vec{v}_{n}^{\prime} \vec{x}_{n}^{\prime} \vec{y}_{n} \cdot N_{n}^{\prime}\right) / y_{0}\right] & (\text { by }(4.30),(4.34)) \\
& =M^{\prime}\left[K \left[\mathrm{a}\left(\lambda \vec{v}_{1} \cdot N_{1}^{\prime}\left[\vec{M}_{1} / \vec{x}_{1}^{\prime}\right]\left[\vec{M}_{1}^{\prime} / \vec{y}_{1}\right]\right)\right.\right. & \\
& \left.\left.\quad \ldots\left(\lambda \vec{v}_{n} \cdot N_{n}^{\prime}\left[\vec{M}_{n} / \vec{x}_{n}^{\prime}\right]\left[\vec{M}_{n}^{\prime} / \vec{y}_{n}\right]\right) / z_{1}\right] / y_{0}\right] & & (\text { by }(4.35)) \\
& =M^{\prime}\left[K\left[\mathrm{a} N_{1} \ldots N_{n} / z_{1}\right] / y_{0}\right] & \text { (by }(4.33)) \\
& =M . & \text { by }(4.32)) \tag{4.32}
\end{array}
$$

Theorem 4.37. For every semilexicalized $A C G \mathscr{G} \in \mathbf{G}(3, n)$, there is a lexicalized $A C G \mathscr{G}^{\prime} \in \mathbf{G}(3, n+1)$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\{R \in \mathcal{O}(\mathscr{G}) \mid R \text { contains a constant }\} .
$$

### 4.3.6 Schabes et al.'s Lexicalization and Ours

Although our notion of lexicalization differs from the one by Schabes et al. $[47,48]$, what we have done is not very far from their definition. They define the notion of lexicalization in a strong sense, where it preserves not only the string language but also the tree language yielding the string language. This definition requires grammars that should be lexicalized to be finitely ambiguous, i.e., the tree language contains at most finitely many trees that yield the same string. Their motivation behind the definition would be to preserve the syntactic structures as well as the surface structures. If we regard both tree languages and abstract languages as syntactic structures that yield string languages and object languages as surface structures, respectively, one might think the definition of "lexicalization of ACGs" should be strengthened in analogy with Schabes et al.'s definition so that the abstract languages of the given ACG and the resultant lexicalized ACG coincide. However, this definition seems too rigorous and there would be very few interesting subclasses of ACGs admit strong lexicalization in this sense.

We alternatively propose to define the notion of "strong lexicalization of ACGs" analogous to Schabes et al.'s definition so that an ACG $\mathscr{G}=$ $\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ is converted into a lexicalized ACG $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L} \circ \mathscr{L}^{\prime}, s^{\prime}\right\rangle$ where $\mathcal{A}(\mathscr{G})=\mathscr{L}^{\prime}\left(\mathcal{A}\left(\mathscr{G}^{\prime}\right)\right)$ holds. According to this definition, one can "recover" the original abstract language from the resultant ACG. Actually, this is what we have done in the latter half of our lexicalization method on the preprocessed ACGs which have neither nullary nor unary nonlexical constants. Moreover, when we assume that input semilexicalized ACGs are finitely ambiguous in the sense that every element of the object language has finitely many abstract terms mapped to it, it is possible to modify the preprocessing that eliminates nullary and unary nonlexical constants so that the resultant grammar can also recover the original abstract language. Conversely, one may think of the preprocessing as elimination of infinite ambiguity from the given semilexicalized ACG. This way our whole algorithm becomes a strong lexicalization method. Therefore, our lexicalization method has a close conceptual similarity to Schabes et al.'s notion of lexicalization.

In the sequel, we concretely see the relation between Schabes et al.'s lexicalization and ours through lexicalization of finitely ambiguous CFGs. Here our lexicalization method means modified lexicalization method, which recovers the original abstract languages. As Schabes et al.'s lexicalization method converts a finitely ambiguous CFG into a "strongly equivalent" lexicalized TAG in the sense that the tree language of the TAG is the set of derivation trees of the CFG, our lexicalization method for semilexicalized ACGs converts the ACG $\mathscr{G}_{G} \in \mathbf{G}_{\text {string }}(2,2)$ encoding a finitely ambiguous CFG


Figure 4.1: Our lexicalization of a CFG
$G$ into the $\mathrm{ACG} \mathscr{G}_{G}^{l} \in \mathrm{G}_{\mathrm{string}}^{\mathrm{lex}}(2,3)$ encoding a lexicalized TAG $G^{\prime}$ strongly equivalent to $G$ "as a string generator". That a string ACG $\mathscr{G}_{G}^{l}$ encodes a TAG $G^{\prime}$ as a string generator means that $\mathscr{G}_{G}^{l}$ is obtained from the ACG $\mathscr{G}_{G^{\prime}} \in \mathbf{G}_{\text {tree }}^{\text {lex }}(2,2(\mathrm{mon}))$ encoding $G^{\prime}$ by the canonical method for transforming a tree ACG into a string ACG generating the yield string language of the original tree ACG (see Proposition 5.5 for the canonical method). Since such $\mathscr{G}_{G^{\prime}}$ and $G^{\prime}$ are effectively computable from $\mathscr{G}_{G}^{l}$, our method gives alternative construction for Schabes's theorem that every finitely ambiguous CFG has a strongly equivalent lexicalized TAG. This is illustrated in Figure 4.1.

More direct comparison would be possible when modifying the definition of "lexicalized ACGs". For the accordance with Schabes et al.'s notion of lexicalized grammar, here we introduce the notion of "strongly lexicalized tree ACGs". We say that an abstract constant of a tree ACG is strongly lexical iff it is mapped to a term containing an object constant of an atomic type. Then the set $\mathscr{C}_{0}$ of abstract constants is partitioned into $\mathscr{D}_{0}$ and $\mathscr{E}_{0}$ where $\mathscr{D}_{0}$ is the set of strongly lexical abstract constants. Note that $\mathscr{C}_{0}^{-} \subseteq \mathscr{E}_{0}$ and $\mathscr{D}_{0} \subseteq \mathscr{C}_{0}^{+}$. We then call the tree ACG strongly lexicalized iff $\mathscr{E}_{0}$ is the empty set. It is not difficult to modify the first half of our lexicalization method, which eliminates nullary and unary constants in $\mathscr{C}_{0}^{-}$, so that it eliminates nullary and unary constants in $\mathscr{E}_{0}$, if the input ACG encodes a finitely ambiguous CFG in the sense of Schabes et al. Besides, recall that the second half of our lexicalization method depends only on the fact that every constant in $\mathscr{C}_{0}^{-}$ has a $k$-ary second-order type for some $k \geq 2$. Thus, replacing $\mathscr{C}_{0}^{+}$with $\mathscr{D}_{0}$ and $\mathscr{C}_{0}^{-}$with $\mathscr{E}_{0}$ in our procedure, we get a strong lexicalization method that completely satisfies Schabes et al.'s notion of lexicalization of CFGs.

### 4.4 Summary

This chapter has discussed lexicalized ACGs, in particular, lexicalization of semilexicalized ACGs. As seen in Section 4.2, to be lexicalized can be thought of as a desirable restriction on ACGs not only from the lexicalists' point of view, but also from the point of view of the computational complexity.

In section 4.3, we have presented a lexicalization method for semilexicalized ACGs, where every nonlexical abstract constant has a type of at most second-order. Our basic strategy for lexicalizing semilexicalized ACGs is as follows. As a preparation, we first eliminate nullary or unary nonlexical constants from the given grammar. Then for the resultant grammar $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$, we construct a new ACG $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{0}, \mathscr{L}^{\prime}, s^{\prime}\right\rangle$ so that $\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\mathcal{A}(\mathscr{G})$ and $\mathscr{L}^{\prime}(\mathrm{A})$ contains a lexical abstract constant in $\mathscr{C}_{0}^{+}$with some finite number of nonlexical constants in $\mathscr{C}_{0}^{-}$for every $A \in \mathscr{C}_{0}^{\prime}$. We have presented three kinds of lexicalization methods depending on the order of given ACGs. While the lexicalization method for ACGs with fourth or higher-order abstract vocabularies preserves the orders of the lexicons, our lexicalization methods for ACGs with second and third-order abstract vocabularies both increase the orders of the lexicons by one.

Our result gives another proof for the theorem that every semilexicalized ACG generates a language in NP, which has been proven by Salvati [43].

Moreover, our lexicalization method for second-order ACGs, though it does not preserves the orders of the lexicons, entails that every LCFRS has an equivalent lexicalized LCFRS modulo the empty string, where an LCFRS is said to be lexicalized iff the description of each function assigned to each production includes at least one terminal symbol. It is not difficult to modify Salvati's conversion [45] from a string second-order ACG $\mathscr{G} \in \mathbf{G}_{\text {string }}(2, n)$ into an equivalent LCFRS so that it preserves lexicalization. Therefore, the ACG $\mathscr{G} \in \mathbf{G}_{\text {string }}(2,4)$ encoding an LCFRS, which has an equivalent lexicalized $\mathrm{ACG} \mathscr{G}^{\prime} \in \mathbf{G}_{\text {string }}^{\text {lex }}(2,5)$ modulo the empty string by our lexicalization, has an equivalent lexicalized LCFRS modulo the empty string.

For the sake of generality, we have defined the notion of lexicalized ACGs so that it does not comprehend Schabes et al.'s definition of lexicalized TAGs [47, 48]. Nevertheless, as discussed in Section 4.3.6, our lexicalization method can be modified so that it entails Schabes et al.'s theorem that every finitely ambiguous CFG has a strongly equivalent lexicalized TAG.

Therefore, our lexicalization method partially generalizes the research by Schabes et al. The result of this chapter, as well as the previous chapter, strengthens de Groote and Pogodalla's view that ACGs can be the kernel of a grammatical framework; not only well known grammar formalisms themselves, but also transformations of existing grammar formalisms are encoded
in ACGs.

## Future Work

It is future work whether or not we can lexicalize second and/or third-order ACGs with preserving the orders of the lexicons. It is known that several subclasses of second-order string ACGs are closed under lexicalization, where the order of the lexicons is preserved.

Recall that $\mathbf{G}_{\text {string }}(2,2)$ is equivalent to the class of context-free grammars. According to the straightforward encoding by de Groote and Pogodalla [15, 19], if a CFG is "lexicalized" in the sense that every production contains a terminal symbol in its right-hand side, then the corresponding secondorder ACG is lexicalized, and vice versa. The fact that every CFG that does not generate the empty string has an equivalent lexicalized CFG, e.g., Greibach normal form, implies that every $\mathscr{G} \in \mathbf{G}_{\text {string }}(2,2)$ has an equivalent $\mathscr{G}^{\prime} \in \mathbf{G}_{\text {string }}^{\text {lex }}(2,2)$ modulo the empty string.

As mentioned above, every string second-order ACG $\mathscr{G} \in \mathbf{G}_{\text {string }}(2, n)$ has an equivalent lexicalized LCFRS modulo the empty string, which has an equivalent lexicalized $\mathrm{ACG} \mathscr{G}^{\prime} \in \mathbf{G}_{\text {string }}^{\mathrm{lex}}(2,4)$.

Moreover, by Schabes et al.'s work on lexicalization of TAGs [47,48], it is not hard to see that the subclass of second-order string ACGs whose lexicons map an atomic type to either str or $s t r \rightarrow s t r$ is closed under lexicalization.

The reason why we cannot preserve the order of the lexicon is that our lexicalization method is designed to be able to "recover" the original abstract language, in the sense that there is a lexicon from the new abstract vocabulary to the original abstract vocabulary such that the original abstract language is exactly the image of the new abstract language, except for the steps eliminating nullary and unary nonlexical constants. In order to lexicalize second-order ACGs with preserving the orders of the lexicons, it seems inevitable to give up recovering the original abstract languages.

If each subclass $\mathbf{G}(2, n)$ is closed under lexicalization, it might entail that some grammar formalisms, e.g., linear CFTGs, admit lexicalization, with or without help of a similar discussion in Section 4.3.6.

On the other hand, it seems difficult to lexicalize general third-order ACGs, because there is a third-order ACG that simulates vector addition systems (Proposition 2.18. See also Proposition 5.7), while every lexicalized ACG generates an NPTIME language (Proposition 4.3). ${ }^{9}$

[^8]
## Chapter 5

## Two-Dimensional Abstract Categorial Grammars

### 5.1 Introduction

If two ACGs $\mathscr{G}_{1}$ and $\mathscr{G}_{2}$ share the same abstract vocabulary and the distinguished type, i.e., $\mathcal{A}\left(\mathscr{G}_{1}\right)=\mathcal{A}\left(\mathscr{G}_{2}\right)$, each element $P_{1}$ of $\mathcal{O}\left(\mathscr{G}_{1}\right)$ is associated with terms $P_{2}$ in $\mathcal{O}\left(\mathscr{G}_{2}\right)$ through the elements $M$ of $\mathcal{A}\left(\mathscr{G}_{1}\right)=\mathcal{A}\left(\mathscr{G}_{2}\right)$ which are mapped to $P_{1}$ and $P_{2}$. This way, ACGs can represent relations of two languages. It is convenient to give a formalization to this paradigm.

Definition 5.1. A two-dimensional $A C G(2 D-A C G)$ is a sextuple $\mathscr{G}=$ $\left\langle\Sigma_{0}, \Sigma_{1}, \Sigma_{2}, \mathscr{L}_{1}, \mathscr{L}_{2}, s\right\rangle$ such that

- $\Sigma_{0}$ is a higher-order signature, called the abstract vocabulary,
- $\Sigma_{i}$ for $i \in\{1,2\}$ is a higher-order signature, called the $i$-th object vocabulary,
- $\mathscr{L}_{i}$ is a lexicon from $\Sigma_{0}$ to $\Sigma_{i}(i$-th lexicon $)$,
- $s \in \mathscr{A}_{0}$ is the distinguished type.

The $i$-th projection of $\mathscr{G}$ is defined as $\mathscr{G}_{i}=\left\langle\Sigma_{0}, \Sigma_{i}, \mathscr{L}_{i}, s\right\rangle$. The abstract language, object language and $i$-th projection of the object language are respectively defined as

$$
\begin{aligned}
\mathcal{A}(\mathscr{G}) & =\left\{|M|_{\beta \eta} \in \Lambda\left(\Sigma_{0}\right) \mid M \text { is a closed linear } \lambda \text {-term of type } s\right\} \\
\mathcal{O}(\mathscr{G}) & \left.=\left\{\left.\langle | \mathscr{L}_{1}(M)\right|_{\beta \eta},\left|\mathscr{L}_{2}(M)\right|_{\beta \eta}\right\rangle \mid M \in \mathcal{A}(\mathscr{G})\right\} \\
\mathcal{O}_{i}(\mathscr{G}) & =\left\{P_{i} \mid P_{i} \in \mathcal{O}\left(\mathscr{G}_{i}\right) \text { for } \mathscr{G}_{i} \text { the } i \text {-th projection of } \mathscr{G}\right\}
\end{aligned}
$$

Moreover, the classification of (one-dimensional) ACGs can be used for 2DACGs. For instance, $\mathbf{G}_{\text {tree,string }}(2,1(\mathrm{r}), 4)$ represents the class of ACGs whose first projection belongs to $\mathbf{G}_{\text {tree }}(2,1(\mathrm{r}))$ and second projection belongs to $\boldsymbol{G}_{\text {string }}(2,4)$.

This way we can define $k$-dimensional ACGs ( $k D-A C G s$ ) for any $k$.
Though symmetric treatment of syntax and semantics of natural language is an important appeal of ACGs, there is little mathematical research on 2DACGs. In this chapter, we discuss the expressive power on relations of 2DACG. Although it is rather important what kind of relations between strings and meaning representations can be expressed by 2D-ACGs, we have no wellknown measure to characterize the class of relations between strings and meanings. Therefore, we evaluate the expressive power of 2D-ACGs through encoding by 2D-ACGs of some well-known tree transducer formalisms that define relations between trees and trees, or trees and strings.

Section 5.2 presents some examples of 2D-ACGs, which imply some closure properties of string ACGs. A fundamental question is whether several mathematical properties of 1D-ACGs still hold for 2D-ACGs. We show that every second-order 2D-ACG generates a PTIME language as every secondorder 1D-ACG does (Theorem 2.11) in Section 5.3. Moreover, we show that every second-order string $2 \mathrm{D}-\mathrm{ACG}$ in $\mathbf{G}_{\text {string }}\left(2, n_{1}, n_{2}\right)$ has an equivalent $2 \mathrm{D}-\mathrm{ACG} \mathrm{G}_{\text {string }}(2,4,4)$, like second-order string 1D-ACGs (Theorem 2.6). This is shown in Section 5.5, after we prove that every deterministic tree walking transducer can be encoded by a second-order 2D-ACG. Section 5.4 shows that the class of linear macro tree transducers exactly corresponds to $\mathbf{G}_{\text {tree,string }}(2,1(\mathrm{sr}), 2)$.

## Related Work

For a binary relation $\mathcal{R}$, the domain $\mathcal{R}_{1}$ of $\mathcal{R}$ is defined as

$$
\mathcal{R}_{1}=\{P \mid\langle P, Q\rangle \in \mathcal{R} \text { for some } Q\}
$$

and the range $\mathcal{R}_{2}$ of $\mathcal{R}$ is defined as

$$
\mathcal{R}_{2}=\{Q \mid\langle P, Q\rangle \in \mathcal{R} \text { for some } P\} .
$$

Finite state transducers are simple extensions of finite state-automata which write some possibly empty string for each transition. The output string by a transducer for an input string is defined as the concatenation of the strings written during the transition steps. A finite state transducer translates a regular language into a regular language. For a formal definition and basic properties of finite state transducers, see a standard text book (e.g., [3]). De Groote establishes the following result.


A bimorphism $\langle\mathcal{L}, \Phi, \Psi\rangle$ is represented by $\left\langle\Sigma_{0}, \Sigma_{1}, \Sigma_{2}, \mathscr{L}_{\Phi} \circ \mathscr{L}_{\mathcal{L}}, \mathscr{L}_{\Psi} \circ \mathscr{L}_{\mathcal{L}}, s\right\rangle$, if $\mathcal{L}$ is represented by $\left\langle\Sigma_{0}, \Sigma_{\mathcal{L}}, \mathscr{L}_{\mathcal{L}}, s\right\rangle$.

Figure 5.1: Bimorphism and ACG

Proposition 5.2 (de Groote [15]). For every finite state transducer $T$, there is a $2 D-A C G \mathscr{G} \in \mathbf{G}_{\text {string,string }}(2,2,2)$ such that

$$
\mathcal{R}(T)=\mathcal{O}(\mathscr{G})
$$

where $\mathcal{R}(T)$ denotes the relation defined by $T$.
Here we would like to emphasize that even though it is known that each of the domain and range of a relation is represented by an ACG, the question whether the relation itself can be represented by a 2D-ACG is not trivial. De Groote has shown Proposition 5.2 using Nivat's Theorem [39], which establishes the equivalence between finite state transducers and bimorphisms. A bimorphism $\mathscr{B}$ is a triple $\langle\mathcal{L}, \Phi, \Psi\rangle$ where $\mathcal{L}$ is a regular language, $\Phi$ and $\Psi$ are homomorphisms. The relation defined by $\mathscr{B}$ is given as $\mathcal{R}(\mathscr{B})=\{\langle\Phi(\mathrm{w}), \Psi(\mathrm{w})\rangle \mid \mathrm{w} \in \mathcal{L}\}$. We see the straightforward correspondence between homomorphisms and first-order lexicons from string signatures to string signatures. Together with the fact that regular languages are represented by ACGs belonging to $\mathbf{G}_{\text {string }}(2,2)$, this homomorphism-lexicon correspondence and Nivat's Theorem entail Proposition 5.2 (see Figure 5.1).

Tree bimorphisms are a tree-version of bimorphisms. A tree bimorphism $\mathscr{B}=\langle\mathcal{L}, \Phi, \Psi\rangle$ consists of a regular tree language $\mathcal{L}$ and two tree homomorphisms $\Phi$ and $\Psi$. Instead of giving the definition of tree bimorphisms in the usual way, it is enough to state just that a tree homomorphism is a first-order $\lambda \mathrm{K}$-lexicon ${ }^{1}$ from a tree signature to a tree signature. The relation defined by $\mathscr{B}$ is given as $\mathcal{R}(\mathscr{B})=\{\langle\Phi(M), \Psi(M)\rangle \mid M \in \mathcal{L}\}$. As (string) bimorphisms are extended to tree bimorphisms, finite state transducers are extended to tree transducers (see [3] for instance). It is known that bottom-up tree transducers that are allowed to translate without reading the label on the current

[^9]node ( $\varepsilon$-rules) correspond to tree-bimorphisms whose first homomorphisms $\Phi$ are semi-relabeling, and bottom-up tree transducers without $\varepsilon$-rules correspond to tree-bimorphisms whose first homomorphisms $\Phi$ are relabeling. Moreover, some kinds of restrictions, such as non-duplication, non-deletion, etc., on transducers and tree-homomorphisms preserve the equivalence. In particular, if a transducer is linear (non-duplicating and non-deleting), then $\Psi$ is linear in the corresponding bimorphism. By the straightforward correspondence between tree-homomorphisms and lexicons, and between regular tree grammars and $\mathbf{G}_{\text {tree }}(2,1(\mathrm{r})$ ) (Figure 5.1 is still valid for tree cases), we have the following corollary.
Corollary 5.3. Every linear bottom-up tree transducer has an equivalent $2 D$ ACG belonging to $\mathbf{G}_{\text {tree,tree }}(2,1(\mathrm{sr}), 1)$, and vice versa. Every linear bottomup tree transducer without $\varepsilon$-rules has an equivalent $2 D-A C G$ belonging to $\mathrm{G}_{\text {tree, tree }}(2,1(\mathrm{r}), 1)$, and vice versa.

Since linear bottom-up tree transducers and linear top-down tree transducers define the same class of tree languages, the above corollary holds for linear top-down transducers. We note that though all $\mathbf{G}_{\text {tree }}(2,1(\mathrm{r}))$, $\mathbf{G}_{\text {tree }}(2,1(\mathrm{sr}))$ and $\mathbf{G}_{\text {tree }}(2,1)$ define the class of regular tree languages, we cannot drop the condition that "the first lexicon is (semi)relabeling" from the above statement.

Shieber [51] has shown that the relations defined by synchronous tree substitution grammars (STSGs) can be represented by linear tree bimorphisms, where a bimorphism is said to be linear iff both tree homomorphisms are linear. This immediately implies the following corollary.
Corollary 5.4. Synchronous tree substitution grammars can be represented by $2 D$-ACGs belonging to $\mathbf{G}_{\text {tree, tree }}(2,1,1)$.

In fact, Yamada [58] has given a direct encoding of STSGs by 2D-ACGs in $\mathbf{G}_{\text {tree,tree }}(2,1,1)$. The converse relation is open. Shieber has presented a method that constructs a "corresponding" STSG $G^{\mathscr{B}}$ for a given linear tree bimorphism $\mathscr{B}=\langle\mathcal{L}, \Phi, \Psi\rangle$ such that the tree relation defined by $\mathscr{B}$ is obtainable from the tree relation defined by the STSG by a relabeling homomorphism.

### 5.2 Examples

Proposition 5.5. For every tree $A C G \mathscr{G} \in \mathbf{G}_{\text {tree }}(m, n)$, one can find a $2 D$ $A C G \mathscr{G}^{\prime} \in \mathbf{G}_{\text {tree,string }}(m, n, n+1)$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\{\langle P, \operatorname{yield}(P)\rangle \mid P \in \mathcal{O}(\mathscr{G})\} .
$$

Proof. For $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$, let a string signature $\Sigma_{2}$ and a lexicon $\mathscr{L}_{2}$ : $\Sigma_{1} \rightarrow \Sigma_{2}$ be defined by

$$
\begin{aligned}
\mathscr{A}_{2} & =\{o\} \\
\mathscr{C}_{2} & =\left\{\mathrm{c} \in \mathscr{C}_{1} \mid \tau_{1}(\mathrm{c})=o\right\}, \\
\tau_{2}(\mathrm{c}) & =\text { str for all } \mathrm{c} \in \mathscr{C}_{2}, \\
\mathscr{L}^{\prime}(o) & =\text { str } \\
\mathscr{L}^{\prime}(\mathrm{c}) & = \begin{cases}\mathrm{c} & \text { if } \mathrm{c} \in \mathscr{C}_{1} \cap \mathscr{C}_{2} \\
\lambda x_{1}^{s t r} \ldots x_{n}^{s t r} \cdot x_{1}+\cdots+x_{n} & \text { if } \tau_{1}(\mathrm{c})=o^{n} \rightarrow o \text { with } n \geq 1 .\end{cases}
\end{aligned}
$$

The 2D-ACG $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}, \Sigma_{1}, \Sigma_{2}, \mathscr{L}, \mathscr{L}^{\prime} \circ \mathscr{L}, s\right\rangle$ satisfies the desired condition.

Proposition 5.6 (Reverse). For every string $A C G \mathscr{G} \in \mathbf{G}_{\text {string }}(m, n)$, one can find a $2 D-A C G \mathscr{G}^{\prime} \in \mathbf{G}_{\text {string,string }}(m, n, n+1)$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\left\{\left\langle/ \mathrm{w} /, / \mathrm{w}^{R} /\right\rangle \mid / \mathrm{w} / \in \mathcal{O}(\mathscr{G})\right\}
$$

where $\mathrm{w}^{R}$ denotes the reverse of w .
Proof. For $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$, let $\Sigma_{0}^{\prime}$ and $\Sigma_{1}^{\prime}$ be respectively the extensions of $\Sigma_{0}$ and $\Sigma_{1}$ such that

$$
\Sigma_{0}^{\prime}\left\{\begin{array} { l } 
{ \mathscr { L } _ { 0 } ^ { \prime } = \mathscr { A } _ { 0 } \cup \{ s ^ { \prime } \} ( s ^ { \prime } \notin \mathscr { A } _ { 0 } ) , } \\
{ \mathscr { C } _ { 0 } ^ { \prime } = \mathscr { C } _ { 0 } \cup \{ \# \} ( \# \notin \mathscr { C } _ { 0 } ) , } \\
{ \tau _ { 0 } ^ { \prime } = \tau _ { 0 } \cup \{ \# \mapsto s \rightarrow s ^ { \prime } \} , }
\end{array} \quad \Sigma _ { 1 } ^ { \prime } \left\{\begin{array}{l}
\mathscr{A}_{1}^{\prime}=\mathscr{A}_{1}=\{o\}, \\
\mathscr{C}_{1}^{\prime}=\mathscr{C}_{1} \cup\{\#\}\left(\# \notin \mathscr{C}_{1}\right), \\
\tau_{1}^{\prime}=\tau_{1} \cup\{\# \mapsto o\},
\end{array}\right.\right.
$$

and define $\mathscr{L}^{\prime}$ and $\mathscr{L}^{\prime \prime}$ as extensions of $\mathscr{L}$ by
$\mathscr{L}^{\prime}: \Sigma_{0}^{\prime} \rightarrow \Sigma_{1}\left\{\begin{array}{l}\mathscr{L}^{\prime}\left(s^{\prime}\right)=s t r, \\ \mathscr{L}^{\prime}(\#)=\lambda z^{s t r} . z,\end{array} \quad \mathscr{L}^{\prime \prime}: \Sigma_{0}^{\prime} \rightarrow \Sigma_{1}^{\prime}\left\{\begin{array}{l}\mathscr{L}^{\prime \prime}\left(s^{\prime}\right)=o, \\ \mathscr{L}^{\prime \prime}(\#)=\lambda z^{s t r} . z \# .\end{array}\right.\right.$
Thus, the ACG $\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L}^{\prime}, s^{\prime}\right\rangle$ is equivalent to $\mathscr{G}$ and the ACG $\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}^{\prime}, \mathscr{L}^{\prime \prime}, s^{\prime}\right\rangle$ generates exactly $\mathcal{O}(\mathscr{G})$ with the endmarker $\#$.

Let a new second-order lexicon $\mathscr{L}^{R}: \Sigma_{1}^{\prime} \rightarrow \Sigma_{1}$ be defined as

$$
\begin{aligned}
\mathscr{L}^{R}(o) & =\operatorname{str} \\
\mathscr{L}^{R}(\mathrm{a}) & =\lambda x^{s t r} \cdot x+\mathrm{a} \\
\mathscr{L}^{R}(\#) & =/ \varepsilon / .
\end{aligned}
$$

We easily see that

$$
\mathscr{L}^{R}\left(\mathrm{a}_{1}\left(\ldots\left(\mathrm{a}_{n} \#\right) \ldots\right)\right)=/ \mathrm{a}_{n} \ldots \mathrm{a}_{1} / .
$$

Thus, for $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \Sigma_{1}, \mathscr{L}^{\prime}, \mathscr{L}^{R} \circ \mathscr{L}^{\prime \prime}, s^{\prime}\right\rangle$, we have

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\left\{\left\langle/ \mathrm{w} /, / \mathrm{w}^{R} /\right\rangle \mid / \mathrm{w} / \in \mathcal{O}(\mathscr{G})\right\} .
$$

Suppose that all the elements of an alphabet $\mathscr{C}$ are ordered as $\left\langle c_{1}, \ldots, c_{n}\right\rangle$. The sorted string $\operatorname{Sort}(\mathrm{w})$ of a string $\mathrm{w} \in \mathscr{C}^{*}$ is defined as

$$
\operatorname{Sort}(\mathrm{w})=\mathrm{c}_{1}^{m_{1}} \ldots \mathrm{c}_{n}^{m_{n}}
$$

where $m_{i}=\# \mathrm{c}_{i}(\mathrm{w})$ for each $i \in\{1, \ldots, n\}$.
Proposition 5.7 (Sort). For every string $A C G \mathscr{G} \in \mathbf{G}_{\text {string }}(m, n)$, there is $a 2 D-A C G \mathscr{G}^{\prime} \in \mathbf{G}_{\text {string,string }}(\max \{3, m\}, n, n)$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\{\langle/ \mathrm{w} /, / \operatorname{Sort}(\mathrm{w}) /\rangle \mid / \mathrm{w} / \in \mathcal{O}(\mathscr{G})\} .
$$

Proof. Let $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle$ and suppose that the elements of the object vocabulary $\mathscr{C}_{1}$ are ordered as $\left\langle\mathrm{c}_{1}, \ldots, \mathrm{c}_{n}\right\rangle$. We construct a $2 \mathrm{D}-\mathrm{ACG} \mathscr{G}^{\prime}=$ $\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \Sigma_{1}, \mathscr{L}^{\prime}, \mathscr{L}^{S}, s_{n}\right\rangle$ as follows. Let

$$
\begin{aligned}
\mathscr{A}_{0}^{\prime} & =\mathscr{A}_{0} \cup\left\{q_{i}, s_{i} \mid 1 \leq i \leq n\right\} \quad\left(q_{i}, s_{i} \notin \mathscr{A}_{0}\right), \\
\mathscr{C}_{0}^{\prime} & =\mathscr{C}_{0} \cup\left\{\mathrm{~B}_{i}, \mathrm{C}_{i} \mid 1 \leq i \leq n\right\} \quad\left(\mathrm{B}_{i}, \mathrm{C}_{i} \notin \mathscr{A}_{0}\right), \\
\tau_{0}^{\prime}(\mathrm{a}) & =q_{1}^{k_{1}} \rightarrow \cdots \rightarrow q_{n}^{k_{n}} \rightarrow \tau_{0}(\mathrm{a}) \text { for } k_{i}=\# \mathrm{c}_{i}(\mathscr{L}(\mathrm{a})) \text { for a } \in \mathscr{C}_{0} \\
\tau_{0}^{\prime}\left(\mathrm{B}_{i}\right) & =s_{i-1} \rightarrow s_{i} \text { where } s_{0}=s, \\
\tau_{0}^{\prime}\left(\mathrm{C}_{i}\right) & =\left(q_{i} \rightarrow s_{i}\right) \rightarrow s_{i} .
\end{aligned}
$$

For $P \in \Lambda\left(\Sigma_{1}\right)$, let $\widetilde{P}$ denote the term obtained from $P$ by replacing the $j$-th occurrence of each constant $\mathrm{c}_{i}$ with a fresh variable $x_{i, j}$, i.e., $\widetilde{P}$ is the constant-free term such that

$$
P=\widetilde{P}\left[c_{i} / x_{i, j}\right]_{1 \leq i \leq n}^{1 \leq j \leq k_{i}}
$$

for the number $k_{i}$ of the occurrences of $\mathrm{c}_{i}$ in $P$. We define

$$
\begin{aligned}
& \mathscr{L}^{\prime}(p)=\mathscr{L}^{S}(p)= \begin{cases}\mathscr{L}(p) & \text { if } p \in \mathscr{A}_{0} \\
s t r & \text { otherwise }\end{cases} \\
& \mathscr{L}^{\prime}(\mathrm{a})=\mathscr{L}^{S}(\mathrm{a})=\lambda x_{1,1}^{q_{1}} \ldots x_{1, k_{1}}^{q_{1}} \ldots x_{n, 1}^{q_{n}} \ldots x_{n, k_{n}}^{q_{n}} \cdot \widetilde{\mathscr{L}(\mathrm{a})} \\
&\text { for } \left.k_{i}=\#_{\mathrm{c}_{i}} \mathscr{L}(\mathrm{a})\right) \text { for } \mathrm{a} \in \mathscr{C}_{0}
\end{aligned} \quad \begin{aligned}
& \mathscr{L}^{\prime}\left(\mathrm{B}_{i}\right)=\mathscr{L}^{S}\left(\mathrm{~B}_{i}\right)=\lambda x^{s t r} \cdot x, \\
& \mathscr{L}^{\prime}\left(\mathrm{C}_{i}\right)=\lambda x^{s t r \rightarrow s t r} \cdot x / \mathrm{c}_{i} / . \\
& \mathscr{L}^{S}\left(\mathrm{C}_{i}\right)=\lambda x^{s t r \rightarrow s t r} \cdot x / \varepsilon /+/ \mathrm{c}_{i} / .
\end{aligned}
$$

We need to show

1. $\left\langle/ \mathrm{w}_{1} /, / \mathrm{w}_{2} /\right\rangle \in \mathcal{O}\left(\mathscr{G}^{\prime}\right)$ implies $\# \mathrm{c}_{i}\left(/ \mathrm{w}_{1} /\right)=\# \mathrm{c}_{i}\left(/ \mathrm{w}_{2} /\right)$ for all $\mathrm{c}_{i} \in \mathscr{C}_{1}$,
2. $/ \mathrm{w} / \in \mathcal{O}_{2}\left(\mathscr{G}^{\prime}\right)$ implies $\mathrm{w} \in\left\{\mathrm{c}_{1}\right\}^{*} \ldots\left\{\mathrm{c}_{1}\right\}^{*}$,
3. $\mathcal{O}(\mathscr{G})=\mathcal{O}_{1}\left(\mathscr{G}^{\prime}\right)$,
(1) is obvious.
(2) If $N \in \mathcal{A}\left(\mathscr{G}^{\prime}\right)$, then $N$ has the form

$$
\begin{aligned}
N= & \mathrm{C}_{n}\left(\lambda x _ { n , 1 } ^ { q _ { n } } \cdot \mathrm { C } _ { n } \left(\lambda x _ { n , 2 } ^ { q _ { n } } \ldots \mathrm { C } _ { n } \left(\lambda x_{n, m_{n}}^{q_{n}} \cdot \mathrm{~B}_{n}( \right.\right.\right. \\
& \vdots \\
& \left.\left.\left.\left.\mathrm{C}_{1}\left(\lambda x_{1,1}^{q_{1}} \cdot \mathrm{C}_{1}\left(\lambda x_{1,2}^{q_{1}} \ldots \mathrm{C}_{1}\left(\lambda x_{1, m_{1}}^{q_{1}} \cdot \mathrm{~B}_{1} N^{\prime}\right) \ldots\right)\right) \ldots\right)\right) \ldots\right)\right) .
\end{aligned}
$$

Thus we see (2) as

$$
\mathscr{L}^{S}(N)=\mathscr{L}^{S}\left(N^{\prime}\right)\left[/ \varepsilon / / x_{i, j}\right]_{1 \leq i \leq n}^{1 \leq j \leq m_{i}}+/ \mathrm{c}_{1}^{m_{1}} /+\cdots+/ \mathrm{c}_{n}^{m_{n}} /=/ \mathrm{c}_{1}^{m_{1}} \ldots \mathrm{c}_{n}^{m_{n}} / .
$$

(3) We show $\mathcal{O}(\mathscr{G}) \subseteq \mathcal{O}_{1}\left(\mathscr{G}^{\prime}\right)$. Suppose that $M \in \mathcal{A}(\mathscr{G})$ is given. Let $k$ be the total number of all occurrences of constants in $M$. That is,

$$
M=M^{\prime}\left[\mathrm{a}_{1} / z_{1}, \ldots, \mathrm{a}_{k} / z_{k}\right]
$$

for a constant-free term $M^{\prime}$ with $\operatorname{Fv}\left(M^{\prime}\right)=\left\{z_{1}, \ldots, z_{k}\right\}$. For $1 \leq i \leq n$ and $1 \leq j \leq k$, let

$$
\begin{aligned}
l_{j, i} & =\sum_{1 \leq h \leq j} \#_{\mathrm{c}_{i}}\left(\mathscr{L}\left(\mathrm{a}_{h}\right)\right) \\
m_{i} & =l_{k, i}=\#_{\mathrm{c}_{i}}(\mathscr{L}(M)) .
\end{aligned}
$$

Let
$N^{\prime}=M^{\prime}\left[\mathrm{a}_{1}\left\langle x_{i, j}^{q_{i}}\right\rangle_{1 \leq i \leq n}^{1 \leq j \leq l_{1, i}} / z_{1}, \mathrm{a}_{2}\left\langle x_{i, j}^{q_{i}}\right\rangle_{1 \leq i \leq n}^{l_{1, i}+1 \leq j \leq l_{2, i}} / z_{2}, \ldots, \mathrm{a}_{k}\left\langle x_{i, j}^{q_{i}}\right\rangle_{1 \leq i \leq n}^{l_{k-1, i}+1 \leq j \leq l_{k, i}} / z_{k}\right]$.
Indeed $N^{\prime}$ is well-typed on $\Sigma_{0}^{\prime}$ and

$$
\mathscr{L}(M)=\mathscr{L}^{\prime}\left(N^{\prime}\right)\left[\mathrm{c}_{i} / x_{i, j}\right]_{1 \leq i \leq n}^{1 \leq j \leq m_{i}} .
$$

For

$$
\begin{aligned}
N= & \mathrm{C}_{n}\left(\lambda x _ { n , 1 } ^ { q _ { n } } \cdot \mathrm { C } _ { n } \left(\lambda x _ { n , 2 } ^ { q _ { n } } \ldots \mathrm { C } _ { n } \left(\lambda x_{n, m_{n}}^{q_{n}} \cdot \mathrm{~B}_{n}( \right.\right.\right. \\
& \vdots \\
& \left.\left.\left.\left.\mathrm{C}_{1}\left(\lambda x_{1,1}^{q_{1}} \cdot \mathrm{C}_{1}\left(\lambda x_{1,2}^{q_{1}} \ldots \mathrm{C}_{1}\left(\lambda x_{1, m_{1}}^{q_{1}} \cdot \mathrm{~B}_{1} N^{\prime}\right) \ldots\right)\right) \ldots\right)\right) \ldots\right)\right),
\end{aligned}
$$

we have $N \in \mathcal{A}\left(\mathscr{G}^{\prime}\right)$ and $\mathscr{L}^{\prime}(N)=\mathscr{L}(M)$.
The converse direction $\mathcal{O}_{1}\left(\mathscr{G}^{\prime}\right) \subseteq \mathcal{O}(\mathscr{G})$ is shown exactly symmetrically. We easily see that every occurrence of a constant a in $N \in \mathcal{A}(\mathscr{G})$ accompanies variables as a $\vec{x}_{1} \ldots \vec{x}_{n}$ where the length of each $\vec{x}_{i}$ is $\#_{c_{i}}(\mathscr{L}(\mathrm{a}))$.

Proposition 5.8 (Reordering). For every string $A C G \mathscr{G} \in \mathbf{G}_{\text {string }}(m, n)$, there is a $2 D-A C G \mathscr{G}^{\prime} \in \mathbf{G}_{\text {string,string }}(\max \{3, m\}, n, n)$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\left\{\langle/ \mathrm{w} /, / \mathrm{v} /\rangle \mid / \mathrm{w} / \in \mathcal{O}(\mathscr{G}), \#_{\mathrm{c}_{i}}(\mathrm{w})=\#_{\mathrm{c}_{i}}(\mathrm{v}) \text { for all } \mathrm{c}_{i} \in \mathscr{C}_{1}\right\} .
$$

Proof. Replace each $s_{i}$ in the proof of Proposition 5.7 with $s=s_{0}$.

### 5.3 One-Dimensionality and Two-Dimensionality

One may wonder whether a 1D-ACG can encode a 2D-ACG. Formally putting, given a 2D-ACG $\mathscr{G}$, can we construct an ACG $\mathscr{G}^{\prime}$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\left\{\lambda w \cdot w P_{1} P_{2} \mid\left\langle P_{1}, P_{2}\right\rangle \in \mathcal{O}(\mathscr{G})\right\} ?
$$

This is an open problem at this moment. We have just a partial answer to the question.

Proposition 5.9. For every $2 D-A C G \mathscr{G} \in \mathbf{G}\left(2, n_{1}, n_{2}\right)$, there is an $A C G$ $\mathscr{G}^{\prime} \in \mathbf{G}(2, n)$ for $n=\max \left\{n_{1}, n_{2}\right\}+2$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\left\{\lambda w \cdot w P_{1} P_{2} \mid\left\langle P_{1}, P_{2}\right\rangle \in \mathcal{O}(\mathscr{G})\right\} .
$$

Proof. Let $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \Sigma_{2}, \mathscr{L}_{1}, \mathscr{L}_{2}, s\right\rangle \in \mathbf{G}\left(2, n_{1}, n_{2}\right)$. We define the corresponding ACG $\mathscr{G}^{\prime} \in \mathbf{G}(2, n)$ with $n=\max \left\{n_{1}, n_{2}\right\}+2$ to have the same abstract vocabulary as $\mathscr{G}$. The new lexicon $\mathscr{L}$ maps each atomic type $p$ to

$$
\mathscr{L}(p)=\left(\mathscr{L}_{1}(p) \rightarrow \mathscr{L}_{2}(p) \rightarrow o\right) \rightarrow o .
$$

For each a $\in \mathscr{C}_{0}$ of type $p_{1} \rightarrow \cdots \rightarrow p_{k} \rightarrow q$, if $\mathscr{L}_{1}(\mathrm{a})=\lambda x_{1} \ldots x_{k} \cdot P_{1}$ and $\mathscr{L}_{2}(\mathrm{a})=\lambda y_{1} \ldots y_{k} \cdot P_{2}$, then we define $\mathscr{L}(\mathrm{a})$ as

$$
\mathscr{L}(\mathrm{a})=\lambda z_{1} \ldots z_{k} w \cdot z_{1}\left(\lambda x_{1} y_{1} \cdot z_{2}\left(\lambda x_{2} y_{2} \ldots z_{k}\left(\lambda x_{k} y_{k} \cdot w P_{1} P_{2}\right) \ldots\right)\right)
$$

It is easy to check that for every variable-free $M \in \Lambda\left(\Sigma_{0}\right)$ of an atomic type, it holds that

$$
\mathscr{L}(M)=\lambda w^{\tau_{1}\left(\mathscr{L}_{1}(M)\right) \rightarrow \tau_{2}\left(\mathscr{L}_{2}(M)\right) \rightarrow o} . w \mathscr{L}_{1}(M) \mathscr{L}_{2}(M) .
$$

It is easy to extend the above proposition for second-order $k$ D-ACGs for $k \geq 2$.

Corollary 5.10. Every second-order $k D-A C G \mathscr{G} \in \mathbf{G}\left(2, n_{1}, \ldots, n_{k}\right)$ generates a PTIME language.

Proof. By Proposition 5.9 and Theorem 2.11.
Corollary 5.11. For every finite state transducer $T$, there is a linear contextfree rewriting system $G^{T}$ such that

$$
\mathcal{L}\left(G^{T}\right)=\left\{\mathrm{w}_{1} \# \mathrm{w}_{2} \mid\left\langle\mathrm{w}_{1}, \mathrm{w}_{2}\right\rangle \in \mathcal{R}(T)\right\}
$$

where $\#$ is a new symbol not in $T$ and $\mathcal{L}\left(G^{T}\right)$ denotes the language defined by $G^{T}$.
Proof. For a 2D-ACG $\mathscr{G}^{T} \in \mathbf{G}_{\text {string,string }}(2,2,2)$ representing the relation defined by the finite state transducer $T$, let $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle \in \mathbf{G}_{\text {string }}(2,4)$ be the one-dimensional representation of $\mathscr{G}^{T}$ constructed by Proposition 5.9.

$$
\mathcal{O}(\mathscr{G})=\left\{\lambda w \cdot w / \mathbf{w}_{1} / / \mathrm{w}_{2} / \mid\left\langle\mathrm{w}_{1}, \mathrm{w}_{2}\right\rangle \in \mathcal{R}(T)\right\} .
$$

Let $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}^{\prime}, \mathscr{L}^{\prime}, s^{\prime}\right\rangle \in \mathbf{G}_{\text {string }}(2,4)$ be the extension of $\mathscr{G}$ obtained by adding the lexical entry

$$
\left\langle \#, s \rightarrow s^{\prime}, \lambda w \cdot w(\lambda x y \cdot x+\#+y)\right\rangle
$$

where $s^{\prime}$ is a new abstract atomic type mapped to $s t r$. Then,

$$
\mathcal{O}\left(\mathscr{G}^{\prime}\right)=\left\{/ \mathrm{w}_{1} \# \mathrm{w}_{2} / \mid\left\langle\mathrm{w}_{1}, \mathrm{w}_{2}\right\rangle \in \mathcal{R}(T)\right\} .
$$

The equivalence between LCFRSs and $\mathbf{G}_{\text {string }}(2,4)$ (Theorem 2.6) completes the proof.

### 5.4 Linear Macro Tree Transducers

One well-known extension of top-down tree transducers is macro tree transducers (MTTs) [2,6], which are obtained by adding the feature of context-free tree grammars to top-down tree transducers. Their output tree languages for fixed input form context-free tree languages. Even though RTGs and linear CFTGs are equivalent to ACGs belonging to $\mathbf{G}_{\text {tree }}(2,1)$ and $\mathbf{G}_{\text {tree }}(2,2)$ respectively, it is not obvious whether linear MTTs are also equivalent to $\mathbf{G}_{\text {tree, tree }}(2,1,2)$. In fact we show the following proposition in this section.
Theorem 5.12. Linear $\varepsilon$-free macro tree transducers are equivalent to $2 D$ ACGs belonging to $\mathbf{G}_{\text {tree,tree }}(2,1(\mathrm{r}), 2)$. Linear macro tree transducers are equivalent to $2 D$-ACGs belonging to $\mathbf{G}_{\text {tree, tree }}(2,1(\mathrm{sr}), 2)$.

Our translation method is essentially same as the proof of the correspondence between CFTGs and MTTs by Engelfriet and Vogler [7]. While the CFTG obtained by their construction from a given MTT may have infinitely many nonterminal symbols and production rules, the CFTG obtained from a given linear MTT through our method is a usual linear CFTG consisting of finitely many production rules.

## Definition

In this section, we provide two fixed special disjoint sets of variables, $\mathscr{X}=$ $\left\{x_{1}, \ldots,\right\}$ and $\mathscr{Y}=\left\{y_{1}, \ldots\right\}$, and let $\mathscr{X}_{k}=\left\{x_{1}, \ldots, x_{k}\right\}, \mathscr{Y}_{k}=\left\{y_{1}, \ldots, y_{k}\right\}$. For usual variables, we use letters $z, z_{0}, z_{1}, z_{2}, \cdots \in \mathscr{Z}$.

For fixed ranked alphabets $Q$ and $\Sigma$, let $R H S(Q, \Sigma, \mathscr{X}, \mathscr{Y}) \subseteq \mathbb{T}(Q \cup \Sigma \cup$ $\mathscr{X} \cup \mathscr{Y})$ be the smallest set $R H S$ such that

- $\mathscr{Y} \subseteq R H S$,
- if $\mathrm{f} \in \Sigma^{(m)}$ and $M_{1}, \ldots, M_{m} \in R H S$, then $\mathrm{f} M_{1} \ldots M_{m} \in R H S$,
- if $q \in Q^{(1+n)}, x_{i} \in \mathscr{X}$, and $N_{1}, \ldots, N_{n} \in R H S$, then $q x_{i} N_{1} \ldots N_{n} \in$ RHS.

A macro tree transducer (MTT) is a quintuple $T=\left\langle\Sigma_{1}, \Sigma_{2}, Q, \Delta, q_{s}\right\rangle$ where

- $\Sigma_{1}, \Sigma_{2}$ and $Q$ are ranked alphabets such that $\left(\Sigma_{1} \cup \Sigma_{2}\right) \cap Q=\varnothing$ and $Q^{(0)}=\varnothing$,
- $q_{s} \in Q^{(1)}$ is called the initial state,
- $\Delta$ is the set of transition rules each of which is of the form

$$
\begin{aligned}
& -q\left(\mathrm{f} x_{1} \ldots x_{m}\right) y_{1} \ldots y_{n} \rightarrow M, \text { or } \\
& -q x_{1} y_{1} \ldots y_{n} \rightarrow N \quad(\text { called an } \varepsilon \text {-rule })
\end{aligned}
$$

where $\mathrm{f} \in \Sigma_{1}^{(m)}, q \in Q^{(1+n)}, M \in \operatorname{RHS}\left(Q, \Sigma_{2}, \mathscr{X}_{m}, \mathscr{Y}_{n}\right)$ and $N \in$ $\operatorname{RHS}\left(Q, \Sigma_{2}, \mathscr{X}_{1}, \mathscr{Y}_{n}\right)$.

An MTT is $\varepsilon$-free if it has no $\varepsilon$-rule.
For $M, M^{\prime} \in \mathbb{T}\left(Q \cup \Sigma_{1} \cup \Sigma_{2}\right)$ and $\rho=q\left(f x_{1} \ldots x_{m}\right) y_{1} \ldots y_{n} \rightarrow K \in \Delta$, we write $M \vdash_{T}^{\rho} M^{\prime}$ iff there is $M_{0} \in \mathbb{T}\left(Q \cup \Sigma_{1} \cup \Sigma_{2} \cup\{z\}\right)$ such that

- $z$ occurs exactly once in $M_{0}$,
- $M=M_{0}\left[q\left(\mathrm{f} M_{1} \ldots M_{m}\right) N_{1} \ldots N_{n} / z\right]$,
- $M^{\prime}=M_{0}\left[K\left[M_{1} / x_{1}, \ldots, M_{m} / x_{m}, N_{1} / y_{1}, \ldots, N_{n} / y_{n}\right] / z\right]$.

For an $\varepsilon$-rule $\rho=q x_{1} y_{1} \ldots y_{n} \rightarrow K \in \Delta$, we write $M \vdash_{T}^{\rho} M^{\prime}$ iff there is $M_{0} \in \mathbb{T}\left(Q \cup \Sigma_{1} \cup \Sigma_{2} \cup\{z\}\right)$ such that

- $z$ occurs exactly once in $M_{0}$,
- $M=M_{0}\left[q M_{1} N_{1} \ldots N_{n} / z\right]$,
- $M^{\prime}=M_{0}\left[K\left[M_{1} / x_{1}, N_{1} / y_{1}, \ldots, N_{n} / y_{n}\right] / z\right]$.
$\vdash_{T}$ is defined to be the union of $\vdash_{T}^{\rho}$ for all $\rho \in \Delta$ and $\vdash_{T}^{*}$ is the reflexive and transitive closure of $\vdash_{T}$. The relation $\mathcal{R}(T)$ determined by $T$ is given as

$$
\mathcal{R}(T)=\left\{\langle M, N\rangle \mid M \in \mathbb{T}\left(\Sigma_{1}\right), N \in \mathbb{T}\left(\Sigma_{2}\right), \text { and } q_{s} M \vdash_{T}^{*} N\right\} .
$$

We say that an MTT $T$ is linear iff

- for every $q\left(\mathrm{f} x_{1} \ldots x_{m}\right) y_{1} \ldots y_{n} \rightarrow K \in \Delta$, each $x_{i}$ and $y_{j}$ appears in $K$ exactly once for $1 \leq i \leq m$ and $1 \leq j \leq n$ and
- for every $q x_{1} y_{1} \ldots y_{n} \rightarrow K \in \Delta$, each $x_{1}$ and $y_{j}$ appears in $K$ exactly once for $1 \leq j \leq n$.

If $T$ is linear, then for a rule $\rho=q\left(\mathrm{f} x_{1} \ldots x_{m}\right) y_{1} \ldots y_{n} \rightarrow K \in \Delta$, each $x_{i}$ has the unique parent $q_{i} \in Q$ in $K$. That is, for some $K_{0} \in \mathbb{T}\left(\Sigma_{2} \cup \mathscr{Y} \cup \mathscr{Z}\right)$ and $\vec{K}_{i}$ consisting of trees in $\mathbb{T}\left(\Sigma_{2} \cup \mathscr{Y} \cup \mathscr{Z}\right), K$ can be represented as

$$
K=K_{0}\left[q_{k_{1}} x_{k_{1}} \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[q_{k_{m}} x_{k_{m}} \vec{K}_{k_{m}} / z_{m}\right]
$$

where $\left\langle k_{1}, \ldots, k_{m}\right\rangle$ is a permutation of $\langle 1, \ldots, m\rangle$. Note that the above does not represent a simultaneous substitution, but a sequential substitution; e.g., $z_{2}$ may appear in $\vec{K}_{k_{1}}$. For instance, if

$$
K \equiv \mathrm{a}\left(q_{1} x_{1} \mathbf{b}\left(\mathbf{a}\left(\mathbf{a}\left(q_{2} x_{2} \mathbf{b}\right)\right) y_{1}\right)\right)\left(q_{3} x_{3} \mathbf{b}\right),
$$

where $\mathrm{a} \in \Sigma_{2}^{(2)}, \mathrm{b} \in \Sigma_{2}^{(0)}, q_{1} \in Q^{(3)}, q_{2}, q_{3} \in Q^{(2)}$, then $K$ can be represented as

$$
K \equiv \mathrm{a} z_{1} z_{3}\left[q_{1} x_{1} \mathrm{~b}\left(\mathrm{a}\left(\mathrm{a} z_{2}\right) y_{1}\right) / z_{1}\right]\left[q_{2} x_{2} \mathrm{~b} / z_{2}\right]\left[q_{3} x_{3} \mathrm{~b} / z_{3}\right] .
$$

## Technical Lemmas

For some technical reason, we extend the definition of $\vdash_{T} \subseteq \mathbb{T}\left(Q \cup \Sigma_{1} \cup \Sigma_{2}\right) \times$ $\mathbb{T}\left(Q \cup \Sigma_{1} \cup \Sigma_{2}\right)$ to define the relation on $\mathbb{T}\left(Q \cup \Sigma_{1} \cup \Sigma_{2} \cup \mathscr{Z}\right) \times \mathbb{T}\left(Q \cup \Sigma_{1} \cup \Sigma_{2} \cup \mathscr{Z}\right)$ in the obvious way.

Definition 5.13. For fixed ranked alphabets $Q, \Sigma_{1}, \Sigma_{2}$, the set of sentential forms $S F\left(Q, \Sigma_{1}, \Sigma_{2}, \mathscr{Z}\right) \subseteq \mathbb{T}\left(Q \cup \Sigma_{1} \cup \Sigma_{2} \cup \mathscr{Z}\right)$ is defined as the smallest set $S F$ such that

- $\mathscr{Z} \subseteq S F$,
- if $\mathrm{f} \in \Sigma_{2}^{(m)}$ and $N_{1}, \ldots, N_{m} \in S F$, then $\mathrm{f} N_{1} \ldots N_{m} \in S F$,
- if $q \in Q^{(1+n)}, M \in \mathbb{T}\left(\Sigma_{1}\right)$, and $N_{1}, \ldots, N_{n} \in S F$, then $q M N_{1} \ldots N_{n} \in$ $S F$.

We simply write $S F$ for $\operatorname{SF}\left(Q, \Sigma_{1}, \Sigma_{2}, \mathscr{Z}\right)$ if $Q, \Sigma_{1}, \Sigma_{2}$ are obvious from the context.

Lemma 5.14 (Engelfriet and Vogler [7]). Let $T=\left\langle\Sigma_{1}, \Sigma_{2}, Q, q_{s}, \Delta\right\rangle$ be an MTT. For all $M \in \operatorname{SF}\left(Q, \Sigma_{1}, \Sigma_{2}, \mathscr{Z}\right)$ and $M^{\prime}, M \vdash_{T} M^{\prime}$ implies $M^{\prime} \in S F\left(Q, \Sigma_{1}, \Sigma_{2}, \mathscr{Z}\right)$.

Clearly $q_{s} M \in S F$ for $M \in \mathbb{T}\left(\Sigma_{1}\right)$, so we can ignore trees that are not in the sentential form.

If $L \in S F$, then $L$ can be represented as

$$
L=K_{0}\left[q_{1} M_{1} \vec{K}_{1} / z_{1}\right] \ldots\left[q_{m} M_{m} \vec{K}_{m} / z_{m}\right]
$$

where each $\vec{K}_{i}$ is a sequence of trees in $\mathbb{T}\left(\Sigma_{2} \cup \mathscr{Z}\right), K_{0} \in \mathbb{T}\left(\Sigma_{2} \cup \mathscr{Z}\right), M_{i} \in$ $\mathbb{T}\left(\Sigma_{1}\right)$ and each $z_{j}$ with $j \in\{1, \ldots, m\}$ occurs in $K_{0}, \vec{K}_{1}, \ldots, \vec{K}_{m}$ exactly once. (We assume that $z_{1}, \ldots, z_{m}$ do not occur in $L$.)

Lemma 5.15. Let $T=\left\langle\Sigma_{1}, \Sigma_{2}, Q, q_{s}, \Delta\right\rangle$ be a linear MTT. Suppose that $L \in S F$ is represented as

$$
L=z_{1}\left[P_{1} / z_{1}\right] \ldots\left[P_{k} / z_{k}\right]
$$

where $P_{i} \in S F, z_{i}$ occurs in $z_{1}, P_{1}, \ldots, P_{i-1}$ exactly once, and $z_{i}$ does not occur in $P_{i}, \ldots, P_{k}$ for each $i \in\{1, \ldots, k\}$. If $L \vdash_{T}^{\rho} L^{\prime}$, then there exist $j \in\{1, \ldots, k\}$ and $P_{j}^{\prime} \in S F$ such that

1. $P_{j} \vdash_{T}^{\rho} P_{j}^{\prime}$,
2. $L^{\prime}=z_{1}\left[P_{1}^{\prime} / z_{1}\right] \ldots\left[P_{k}^{\prime} / z_{k}\right]$ where $P_{i}^{\prime}=P_{i}$ for $i \neq j$,
3. $P_{i}^{\prime} \in S F$ for each $i \in\{1, \ldots, k\}$.
4. $z_{i}$ occurs in $z_{1}, P_{1}^{\prime}, \ldots, P_{i-1}^{\prime}$ exactly once, and $z_{i}$ does not occur in $P_{i}^{\prime}, \ldots, P_{k}^{\prime}$ for each $i \in\{1, \ldots, k\}$.

Proof. If $L \vdash_{T}^{\rho} L^{\prime}$, we have two cases.
Case 1. Suppose that $\rho$ is not an $\varepsilon$-rule.

$$
\rho=q(\mathrm{f} \vec{x}) \vec{y} \rightarrow K
$$

By the definition of a derivation, $L$ and $L^{\prime}$ are written as

$$
\begin{aligned}
L & =L_{0}[q(f \vec{M}) \vec{N} / z] \\
L^{\prime} & =L_{0}[K[\vec{M} / \vec{x}, \vec{N} / \vec{y}] / z] .
\end{aligned}
$$

Let $j$ be such that the focused occurrence of $q$ in $L$ appears in $P_{j}$. Since $z_{j}$ occurs exactly once in $z_{1}, P_{1}, \ldots, P_{j-1}$, the occurrence of $q$ in $P_{j}$ corresponds to exactly one occurrence of $q$ in $L . P_{j}$ can be represented as

$$
P_{j}=P_{j, 0}\left[q(\mathrm{f} \vec{M}) \overrightarrow{N^{\prime}} / z\right]
$$

since $P_{j} \in S F$. For

$$
P_{j}^{\prime}=P_{j, 0}\left[K\left[\vec{M} / \vec{x}, \overrightarrow{N^{\prime}} / \vec{y}\right] / z\right],
$$

we have $P_{j} \vdash_{T}^{\rho} P_{j}^{\prime}$. (1) is proved. By Lemma 5.14, $P_{j}^{\prime} \in S F$ and thus (3) is proved.

Let $\phi$ and $\psi$ denote the sequential substitutions $\left[P_{1} / z_{1}\right] \ldots\left[P_{j-1} / z_{j-1}\right]$ and $\left[P_{j+1} / z_{j+1}\right] \ldots\left[P_{k} / z_{k}\right]$, respectively. Since $z_{j}$ occurs in $z_{1}, P_{1}, \ldots, P_{j-1}$ exactly once, we have

$$
\begin{aligned}
L_{0} & =z_{1} \phi\left[P_{j, 0} / z_{j}\right] \psi \\
\vec{N} & =\vec{N}^{\prime} \psi,
\end{aligned}
$$

where each element of $z_{j+1}, \ldots, z_{k}$ appears in either $z_{1} \phi\left[P_{j, 0} / z_{j}\right]$ or $\vec{N}^{\prime}$. Therefore,

$$
\begin{aligned}
L^{\prime} & =L_{0}[K[\vec{M} / \vec{x}, \vec{N} / \vec{y}] / z] \\
& =z_{1} \phi\left[P_{j, 0} / z_{j}\right] \psi[K[\vec{M} / \vec{x}, \vec{N} / \vec{y}] / z] \\
& =z_{1} \phi\left[P_{j, 0} / z_{j}\right]\left[K\left[\vec{M} / \vec{x}, \overrightarrow{N^{\prime}} / \vec{y}\right] / z\right] \psi \\
& =z_{1} \phi\left[P_{j}^{\prime} / z_{j}\right] \psi .
\end{aligned}
$$

(2) is proved.
(4) is clear by the linearity.

Case 2. If $\rho$ is an $\varepsilon$-rule, similarly we can show the lemma.
Corollary 5.16. Let $T=\left\langle\Sigma_{1}, \Sigma_{2}, Q, q_{s}, \Delta\right\rangle$ be a linear MTT. Suppose that $L \in S F$ is represented as

$$
L=z_{1}\left[P_{1} / z_{1}\right] \ldots\left[P_{k} / z_{k}\right]
$$

where $P_{i} \in S F, z_{i}$ occurs in $P_{1}, \ldots, P_{i-1}$ exactly once, and $z_{i}$ does not occur in $P_{i}, \ldots, P_{k}$ for each $i \in\{1, \ldots, k\}$. If

$$
L \vdash_{T}^{\rho_{1}} \cdots \vdash_{T}^{\rho_{l}} L^{\prime}
$$

there is a partition $\left\langle S_{1}, \ldots, S_{k}\right\rangle$ of $\{1, \ldots, l\}$ such that

- $S_{i}=\left\{n_{i, 1}, \ldots, n_{i, m_{i}}\right\}$ with $n_{i, j}<n_{i, j+1}$,
- $P_{i} \vdash_{\rho_{n_{i, 1}}} \cdots \vdash_{\rho_{n_{i, m_{i}}}} P_{i}^{\prime}$,
- $L^{\prime}=z_{1}\left[P_{1}^{\prime} / z_{1}\right] \ldots\left[P_{k}^{\prime} / z_{k}\right]$.


## Transformation

Here we present a method of transforming a linear MTT $T=\left\langle\Sigma_{1}, \Sigma_{2}, Q, \Delta, q_{s}\right\rangle$ into a $2 \mathrm{D}-\mathrm{ACG} \mathscr{G}^{T}=\left\langle\Sigma_{0}, \Sigma_{1}, \Sigma_{2}, \mathscr{L}_{1}, \mathscr{L}_{2}, s\right\rangle$. The set of abstract atomic types is defined as the set of states, $\mathscr{A}_{0}=Q$ and define $\mathscr{L}_{1}(q)=o$ and $\mathscr{L}_{2}(q)=o^{n} \rightarrow o$ for each $q \in Q^{(1+n)}$. For each rule $\rho \in \Delta$, we put the following lexical entries into $\mathscr{G}^{T}$ :

For $\rho=q\left(\mathrm{f} x_{1} \ldots x_{m}\right) y_{1} \ldots y_{n} \rightarrow K_{0}\left[q_{k_{1}} x_{k_{1}} \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[q_{k_{m}} x_{k_{m}} \vec{K}_{k_{m}} / z_{m}\right]$, where $\left\langle k_{1}, \ldots, k_{m}\right\rangle$ is a permutation of $\langle 1, \ldots, m\rangle$,

$$
\begin{aligned}
\llbracket \rho \rrbracket & \in \mathscr{C}_{0}, \\
\tau_{0}(\llbracket \rho \rrbracket) & =q_{1} \rightarrow \cdots \rightarrow q_{m} \rightarrow q \\
\mathscr{L}_{1}(\llbracket \rho \rrbracket) & =\mathrm{f} \\
\mathscr{L}_{2}(\llbracket \rho \rrbracket) & =\lambda x_{1} \ldots x_{m} y_{1} \ldots y_{n} . K_{0}\left[x_{k_{1}} \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[x_{k_{m}} \vec{K}_{k_{m}} / z_{m}\right]
\end{aligned}
$$

and for an $\varepsilon$-rule $\rho=q x_{1} y_{1} \ldots y_{n} \rightarrow K_{0}\left[q_{1} x_{1} \vec{K}_{1} / z_{1}\right]$,

$$
\begin{aligned}
\llbracket \rho \rrbracket & \in \mathscr{C}_{0}, \\
\tau_{0}(\llbracket \rho \rrbracket) & =q_{1} \rightarrow q, \\
\mathscr{L}_{1}(\llbracket \rho \rrbracket) & =\lambda z^{o} . z, \\
\mathscr{L}_{2}(\llbracket \rho \rrbracket) & =\lambda x_{1} y_{1} \ldots y_{n} \cdot K_{0}\left[x_{1} \vec{K}_{1} / z_{1}\right] .
\end{aligned}
$$

Since $T$ is linear, $\mathscr{L}_{2}(\llbracket \rho \rrbracket)$ is indeed a well-typed linear $\lambda$-term.
Lemma 5.17. $\mathcal{O}\left(\mathscr{G}^{T}\right) \subseteq \mathcal{R}(T)$
Proof. We show by induction on $M$ that if $M \in \Lambda\left(\Sigma_{0}\right)$ is variable-free and has an atomic type $q \in Q^{(1+n)}$, then

$$
q\left|\mathscr{L}_{1}(M)\right|_{\beta} N_{1} \ldots N_{n} \vdash_{T}^{*}\left|\mathscr{L}_{2}(M) N_{1} \ldots N_{n}\right|_{\beta}
$$

for every $N_{i} \in \mathbb{T}\left(\Sigma_{2} \cup \mathscr{Z}\right)$. This implies that for any $M \in \mathcal{A}\left(\mathscr{G}^{T}\right)$, we have $q_{s}\left|\mathscr{L}_{1}(M)\right|_{\beta} \vdash_{T}^{*}\left|\mathscr{L}_{2}(M)\right|_{\beta}$.

Let the head of $M$ be $\llbracket \rho \rrbracket$. We have two cases.
Case 1. Suppose that $\rho \in \Delta$ is not an $\varepsilon$-rule,

$$
\rho=q\left(\mathrm{f} x_{1} \ldots x_{m}\right) y_{1} \ldots y_{n} \rightarrow K_{0}\left[q_{k_{1}} x_{k_{1}} \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[q_{k_{m}} x_{k_{m}} \vec{K}_{k_{m}} / z_{m}\right]
$$

where $\left\langle k_{1}, \ldots, k_{m}\right\rangle$ is a permutation of $\langle 1, \ldots, m\rangle$. By $\tau_{0}(\llbracket \rho \rrbracket)=q_{1} \rightarrow \cdots \rightarrow$ $q_{m} \rightarrow q, M$ has the form

$$
M=\llbracket \rho \rrbracket M_{1} \ldots M_{m}
$$

for some $M_{i}$ with $\tau_{0}\left(M_{i}\right)=q_{i}$ and $\tau_{0}(M)=q$. Applying the induction hypothesis to each $M_{i}$, we have

$$
\begin{equation*}
q_{i}\left|\mathscr{L}_{1}\left(M_{i}\right)\right|_{\beta} \vec{K}_{i} \vdash_{T}^{*}\left|\mathscr{L}_{2}\left(M_{i}\right) \vec{K}_{i}\right|_{\beta} . \tag{5.1}
\end{equation*}
$$

Let $\vec{x}=x_{1} \ldots x_{m}, \vec{y}=y_{1} \ldots y_{n}, \vec{N}=N_{1} \ldots N_{n}$. By the definition,

$$
\begin{aligned}
\mathscr{L}_{1}(\llbracket \rho \rrbracket) & =\mathrm{f} \\
\mathscr{L}_{2}(\llbracket \rho \rrbracket) & =\lambda \vec{x} \vec{y} \cdot K_{0}\left[x_{k_{1}} \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[x_{k_{m}} \vec{K}_{k_{m}} / z_{m}\right] .
\end{aligned}
$$

Since

$$
\begin{aligned}
\mathscr{L}_{2}(M) \vec{N} & =\mathscr{L}_{2}(\llbracket \rho \rrbracket) \mathscr{L}_{2}\left(M_{1}\right) \ldots \mathscr{L}_{2}\left(M_{m}\right) \vec{N} \\
& =\left(\lambda \vec{x} \vec{y} \cdot K_{0}\left[x_{k_{1}} \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[x_{k_{m}} \vec{K}_{k_{m}} / z_{m}\right]\right) \mathscr{L}_{2}\left(M_{1}\right) \ldots \mathscr{L}_{2}\left(M_{m}\right) \vec{N} \\
& \rightarrow_{\beta} K_{0}\left[\mathscr{L}_{2}\left(M_{k_{1}}\right) \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[\mathscr{L}_{2}\left(M_{k_{m}}\right) \vec{K}_{k_{m}} / z_{m}\right][\vec{N} / \vec{y}],
\end{aligned}
$$

we have

$$
\begin{align*}
q \mathscr{L}_{1}(M) & N_{1} \ldots N_{n}=q\left(\mathrm{f}\left|\mathscr{L}_{1}\left(M_{1}\right)\right|_{\beta} \ldots\left|\mathscr{L}_{1}\left(M_{m}\right)\right|_{\beta}\right) \vec{N} \\
& \vdash_{T}^{\rho} K_{0}\left[q_{k_{1}}\left|\mathscr{L}_{1}\left(M_{k_{1}}\right)\right|_{\beta} \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[q_{k_{m}}\left|\mathscr{L}_{1}\left(M_{k_{m}}\right)\right|_{\beta} \vec{K}_{k_{m}} / z_{m}\right][\vec{N} / \vec{y}] \\
& \vdash_{T}^{*} K_{0}\left[\left|\mathscr{L}_{2}\left(M_{k_{1}}\right) \vec{K}_{k_{1}}\right|_{\beta} / z_{1}\right] \ldots\left[\left|\mathscr{L}_{2}\left(M_{k_{m}}\right) \vec{K}_{k_{m}}\right|_{\beta} / z_{m}\right][\vec{N} / \vec{y}] \tag{5.1}
\end{align*}
$$

$$
={ }_{\beta} \mathscr{L}_{2}(M) N_{1} \ldots N_{n} .
$$

Case 2. Suppose that $\rho$ is an $\varepsilon$-rule of the form

$$
q x_{1} y_{1} \ldots y_{n} \rightarrow K_{0}\left[q_{1} x_{1} \vec{K}_{1}\right] .
$$

Then, $\tau_{0}(\llbracket \rho \rrbracket)=q_{1} \rightarrow q, M=\llbracket \rho \rrbracket M_{1}$ for some $M_{1}$ with $\tau_{0}\left(M_{1}\right)=q_{1}$, and $\tau_{0}(M)=q$. Applying the induction hypothesis to $M_{1}$, we have

$$
q_{1}\left|\mathscr{L}_{1}\left(M_{1}\right)\right|_{\beta} \vec{K}_{1} \vdash_{T}^{*}\left|\mathscr{L}_{2}\left(M_{1}\right) \vec{K}_{1}\right|_{\beta} .
$$

By the definition,

$$
\begin{aligned}
\mathscr{L}_{1}(\llbracket \rho \rrbracket) & =\lambda z^{o} . z \\
\mathscr{L}_{2}(\llbracket \rho \rrbracket) & =\lambda x_{1} y_{1} \ldots y_{n} \cdot K_{0}\left[x_{1} \vec{K}_{1} / z_{1}\right] .
\end{aligned}
$$

Let $\vec{y}=y_{1} \ldots y_{n}$ and $\vec{N}=N_{1} \ldots, N_{n}$. Since

$$
\begin{aligned}
\mathscr{L}_{2}(M) \vec{N} & =\mathscr{L}_{2}(\llbracket \rho \rrbracket) \mathscr{L}_{2}\left(M_{1}\right) \vec{N} \\
& =\left(\lambda x_{1} \vec{y} \cdot K_{0}\left[x_{1} \vec{K}_{1} / z_{1}\right]\right) \mathscr{L}_{2}\left(M_{1}\right) \vec{N} \\
& \rightarrow_{\beta} K_{0}\left[\mathscr{L}_{2}\left(M_{1}\right) K_{1} / z_{1}\right][\vec{N} / \vec{y}],
\end{aligned}
$$

we have

$$
\begin{aligned}
q \mathscr{L}_{1}(M) \vec{N} & \rightarrow_{\beta} q\left|\mathscr{L}_{1}\left(M_{1}\right)\right|_{\beta} \vec{N} \\
& \vdash_{T}^{\rho} K_{0}\left[q_{1}\left|\mathscr{L}_{1}\left(M_{1}\right)\right|_{\beta} \vec{K}_{1} / z_{1}\right][\vec{N} / \vec{y}] \\
& \vdash_{T}^{*} K_{0}\left[\left|\mathscr{L}_{2}\left(M_{1}\right) \vec{K}_{1}\right|_{\beta} / z_{1}\right][\vec{N} / \vec{y}] \\
& ={ }_{\beta} \mathscr{L}_{2}(M) \vec{N} .
\end{aligned}
$$

Lemma 5.18. $\mathcal{R}(T) \subseteq \mathcal{O}\left(\mathscr{G}^{T}\right)$.
Proof. We show by induction on $k$ that if

$$
q M N_{1} \ldots N_{n} \vdash_{T}^{\rho_{0}} P_{1} \vdash_{T}^{\rho_{1}} \cdots \vdash_{T}^{\rho_{k}} P_{k}
$$

where $q \in Q^{(1+n)}, M \in \mathbb{T}\left(\Sigma_{1}\right)$ and $N_{1}, \ldots, N_{n}, P_{k} \in \mathbb{T}\left(\Sigma_{2} \cup \mathscr{Z}\right)$, then there is $L \in \Lambda\left(\Sigma_{0}\right)$ of type $q$ such that $\left|\mathscr{L}_{1}(L)\right|_{\beta} \equiv M$ and $\left|\mathscr{L}_{2}(L) N_{1} \ldots N_{n}\right|_{\beta} \equiv P_{k}$.

Case 1. Suppose that $\rho_{0}$ is not an $\varepsilon$-rule. Then, $M, \rho_{0}, P_{1}$ can be written as

$$
\begin{aligned}
M & =\mathrm{f} M_{1} \ldots M_{m} \\
\rho_{0} & =q\left(\mathrm{f} x_{1} \ldots x_{m}\right) y_{1} \ldots y_{n} \rightarrow K_{0}\left[q_{k_{1}} x_{k_{1}} \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[q_{k_{m}} x_{k_{m}} \vec{K}_{k_{m}} / z_{m}\right] \\
P_{1} & =K_{0}\left[q_{k_{1}} M_{k_{1}} \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[q_{k_{m}} M_{k_{m}} \vec{K}_{k_{m}} / z_{m}\right][\vec{N} / \vec{y}]
\end{aligned}
$$

where $\vec{N}=N_{1} \ldots N_{n}, \vec{y}=y_{1} \ldots y_{n}$, and $\left\langle k_{1}, \ldots, k_{m}\right\rangle$ is a permutation of $\langle 1, \ldots, m\rangle$. By Lemma 5.16, we have a partition $\left\langle S_{1}, \ldots, S_{m}\right\rangle$ of $\{1, \ldots, k\}$ such that

- $S_{i}=\left\{n_{i, 1}, \ldots, n_{i, l_{i}}\right\}$ with $n_{i, j}<n_{i, j+1}$,
- $q_{i} M_{i} \vec{K}_{i}[\vec{N} / \vec{y}] \vdash_{T}^{\rho_{i, 1}} \cdots \vdash_{T}^{\rho_{n_{i, l_{i}}}} R_{i}$,
- $P_{k}=K_{0}\left[R_{k_{1}} / z_{1}\right] \ldots\left[R_{k_{m}} / z_{m}\right]$.

By the induction hypothesis, there is $L_{i} \in \Lambda\left(\Sigma_{0}\right)$ of type $q_{i}$ such that $\mathscr{L}_{1}\left(L_{i}\right)=M_{i}$ and

$$
\begin{equation*}
\mathscr{L}_{2}\left(L_{i}\right) \vec{K}_{i}[\vec{N} / \vec{y}]=R_{i} . \tag{5.2}
\end{equation*}
$$

For $L=\llbracket \rho_{0} \rrbracket L_{1} \ldots L_{m}$, we see that

$$
\mathscr{L}_{1}(L)=\mathscr{L}_{1}\left(\llbracket \rho_{0} \rrbracket\right) \mathscr{L}_{1}\left(L_{1}\right) \ldots \mathscr{L}_{1}\left(L_{m}\right)=\mathrm{f} M_{1} \ldots M_{m}=M
$$

and

$$
\begin{align*}
& \mathscr{L}_{2}(L) N_{1} \ldots N_{n} \\
& =\mathscr{L}_{2}\left(\llbracket \rho_{0} \rrbracket\right) \mathscr{L}_{2}\left(L_{1}\right) \ldots \mathscr{L}_{2}\left(L_{m}\right) \vec{N} \\
& =\left(\lambda x_{1} \ldots x_{m} \vec{y} \cdot K_{0}\left[x_{k_{1}} \vec{K}_{k_{1}} / z_{1}\right] \ldots\left[x_{k_{m}} \vec{K}_{k_{m}} / z_{m}\right]\right) \mathscr{L}_{2}\left(L_{1}\right) \ldots \mathscr{L}_{2}\left(L_{m}\right) \vec{N} \\
& =K_{0}\left[\mathscr{L}_{2}\left(L_{k_{1}}\right) \vec{K}_{k_{1}}[\vec{N} / \vec{y}] / z_{1}\right] \ldots\left[\mathscr{L}_{2}\left(L_{k_{m}}\right) \vec{K}_{k_{m}}[\vec{N} / \vec{y}] / z_{m}\right] \\
& =K_{0}\left[R_{k_{1}} / z_{1}\right] \ldots\left[R_{k_{m}} / z_{m}\right]  \tag{5.2}\\
& =P_{k} .
\end{align*}
$$

Case 2. Suppose that $\rho_{0}$ is an $\varepsilon$-rule. Then, $\rho_{0}$ and $P_{1}$ can be written as

$$
\begin{aligned}
& \rho_{0}=q x_{1} y_{1} \ldots y_{n} \rightarrow K_{0}\left[q_{1} x_{1} \vec{K}_{1} / z_{1}\right] \\
& P_{1}=K_{0}\left[q_{1} M \vec{K}_{1} / z_{1}\right][\vec{N} / \vec{y}]
\end{aligned}
$$

where $\vec{N}=N_{1} \ldots N_{n}$ and $\vec{y}=y_{1} \ldots y_{n}$. There is $R_{1} \in \mathbb{T}\left(\Sigma_{2} \cup \mathscr{Z}\right)$ such that

- $q_{1} M \vec{K}_{1}[\vec{N} / \vec{y}] \vdash_{T}^{\rho_{1}} \cdots \vdash_{T}^{\rho_{k}} R_{1}$,
- $P_{k}=K_{0}\left[R_{1} / z_{1}\right]$.

By the induction hypothesis, there is $L_{1} \in \Lambda\left(\Sigma_{0}\right)$ of type $q_{1}$ such that $\mathscr{L}_{1}\left(L_{1}\right)=M$ and

$$
\left|\mathscr{L}_{2}\left(L_{1}\right) \vec{K}_{1}[\vec{N} / \vec{y}]\right|_{\beta}=R_{1}
$$

For $L=\llbracket \rho_{0} \rrbracket L_{1}$, we see that

$$
\mathscr{L}_{1}(L)=\mathscr{L}_{1}\left(\llbracket \rho_{0} \rrbracket\right) \mathscr{L}_{1}\left(L_{1}\right)=\left(\lambda z^{o} . z\right) M=M,
$$

and

$$
\begin{aligned}
& \mathscr{L}_{2}(L) N_{1} \ldots N_{n} \\
& =\mathscr{L}_{2}\left(\llbracket \rho_{0} \rrbracket\right) \mathscr{L}_{2}\left(L_{1}\right) N_{1} \ldots N_{n} \\
& =\left(\lambda x_{1} y_{1} \ldots y_{n} \cdot K_{0}\left[x_{1} \vec{K}_{1} / z_{1}\right]\right) \mathscr{L}_{2}\left(L_{1}\right) N_{1} \ldots N_{n} \\
& =K_{0}\left[\mathscr{L}_{2}\left(L_{1}\right) \vec{K}_{1}[\vec{N} / \vec{y}] / z_{1}\right] \\
& =K_{0}\left[R_{1} / z_{1}\right] \\
& =P_{k} .
\end{aligned}
$$

Corollary 5.19. Every linear $\varepsilon$-free macro tree transducer has an equivalent $2 D-A C G$ belonging to $\mathbf{G}_{\text {tree,tree }}(2,1(\mathrm{r}), 2)$. Every linear macro tree transducers has an equivalent $2 D$-ACGs belonging to $\mathbf{G}_{\text {tree, tree }}(2,1(\mathrm{sr}), 2)$.

It is not difficult to establish the converse direction. We define a linear MTT $T^{\mathscr{G}}=\left\langle\Sigma_{1}, \Sigma_{2}, Q, q_{s}, \Delta\right\rangle$ equivalent to a given 2D-ACG $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \Sigma_{2}\right.$, $\left.\mathscr{L}_{1}, \mathscr{L}_{2}, s\right\rangle \in \mathbf{G}_{\text {tree, tree }}(2,1(\mathrm{sr}), 2)$. Let

$$
Q^{(1+n)}=\left\{q \in \mathscr{A}_{0} \mid \mathscr{L}_{2}(q)=o^{n} \rightarrow o\right\}
$$

and $q_{s}=s$. For each a $\in \mathscr{C}_{0}$ of type $q_{1} \rightarrow \cdots \rightarrow q_{m} \rightarrow q$, if $\mathscr{L}_{2}(q)$ is $n$-ary, then put the transition rule

$$
q\left(\left|\mathscr{L}_{1}(\mathrm{a}) x_{1} \ldots x_{m}\right|_{\beta}\right) y_{1} \ldots y_{n} \rightarrow\left|\mathscr{L}_{2}(\mathrm{a})\left(q_{1} x_{1}\right) \ldots\left(q_{m} x_{m}\right) y_{1} \ldots y_{n}\right|_{\beta}
$$

into $\Delta$. It is easy to check that the above method is exactly the inverse of our encoding of linear MTTs by ACGs. Thus the resultant linear MTT $T^{\mathscr{G}}$ is equivalent to the given $2 \mathrm{D}-\mathrm{ACG} \mathscr{G}$. This shows the following statement.

Theorem 5.20. Every linear $\varepsilon$-free macro tree transducer has an equivalent $2 D-A C G$ belonging to $\mathbf{G}_{\text {tree, tree }}(2,1(\mathrm{r}), 2)$ and vice versa. Every linear macro tree transducers has an equivalent $2 D$-ACGs belonging to $\mathbf{G}_{\text {tree, tree }}(2,1(\mathrm{sr}), 2)$ and vice versa.

## Synchronous Tree Adjoining Grammars and 2D-ACGs

The notion of tree bimorphisms can be extended by replacing each homomorphism with general mappings, as lexicons are allowed to be higher-order. Shieber [52] has shown that the relation defined by synchronous tree adjoining grammars (STAGs), which are an extension of STSGs, can be represented by tree bimappings $\mathscr{B}=\langle\mathcal{L}, \Phi, \Psi\rangle$ where $\mathcal{L}$ is a regular tree language and both mappings $\Phi$ and $\Psi$ are defined by linear monadic $\varepsilon$-free MTTs. In a monadic MTT, $Q=Q^{(1)} \cup Q^{(2)}$ holds. Actually what Theorem 5.20 states is that linear MTTs can be represented by $2 D$-ACGs, not by lexicons. Nevertheless, the following Lemma entails that STAGs are also encodable by 2D-ACGs.

Lemma 5.21. Let two $2 D$-ACGs $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \Sigma_{2}, \mathscr{L}_{1}, \mathscr{L}_{2}, s\right\rangle \in \mathrm{G}(2,1(\mathrm{r}), n)$ and $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \Sigma_{2}^{\prime}, \mathscr{L}_{1}^{\prime}, \mathscr{L}_{2}^{\prime}, s^{\prime}\right\rangle \in \mathbf{G}\left(2,1(\mathrm{r}), n^{\prime}\right)$ are given. There is a $2 D$ $A C G \mathscr{G}^{\prime \prime} \in \mathbf{G}\left(2, n, n^{\prime}\right)$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime \prime}\right)=\left\{\langle P, Q\rangle \mid\langle M, P\rangle \in \mathcal{O}(\mathscr{G}) \text { and }\langle M, Q\rangle \in \mathcal{O}\left(\mathscr{G}^{\prime}\right)\right\} .
$$



Figure 5.2: Bimapping and ACG (Lemma 5.21)
Proof. Let $\Sigma_{0}^{\prime \prime}, \mathscr{L}_{1}^{\prime \prime}: \Sigma_{0}^{\prime \prime} \rightarrow \Sigma_{1}, \mathscr{L}: \Sigma_{0}^{\prime \prime} \rightarrow \Sigma_{2}, \mathscr{L}^{\prime}: \Sigma_{0}^{\prime \prime} \rightarrow \Sigma_{2}^{\prime}$ be defined as

$$
\begin{aligned}
& \mathscr{A}_{0}^{\prime \prime}=\left\{[p, q] \mid p \in \mathscr{A}_{0}, q \in \mathscr{L}_{0}^{\prime}, \mathscr{L}_{1}(p)=\mathscr{L}_{1}^{\prime}(q)\right\}, \\
& \mathscr{C}_{0}^{\prime \prime}=\left\{\llbracket \mathrm{a}, \mathrm{~b} \rrbracket \mid \mathrm{a} \in \mathscr{C}_{0}, \mathrm{~b} \in \mathscr{C}_{0}^{\prime}, \mathscr{L}_{1}(\mathrm{a})=\mathscr{L}_{1}^{\prime}(\mathrm{b})\right\}, \\
& \tau_{0}^{\prime \prime}(\llbracket \mathrm{a}, \mathrm{~b} \rrbracket)=\left[p_{1}, q_{1}\right] \rightarrow \cdots \rightarrow\left[p_{k}, q_{k}\right] \rightarrow\left[p_{0}, q_{0}\right] \\
& \text { if } \tau_{0}(\mathrm{a})=p_{1} \rightarrow \cdots \rightarrow p_{k} \rightarrow p_{0}, \tau_{0}(\mathrm{~b})=q_{1} \rightarrow \cdots \rightarrow q_{k} \rightarrow q_{0}, \\
& \mathscr{L}([p, q])=\mathscr{L}_{2}(p), \quad \mathscr{L}^{\prime}([p, q])=\mathscr{L}_{2}^{\prime}(q), \\
& \mathscr{L}(\llbracket \mathrm{a}, \mathrm{~b} \rrbracket)=\mathscr{L}_{2}(\mathrm{a}), \quad \mathscr{L}^{\prime}(\llbracket \mathrm{a}, \mathrm{~b} \rrbracket)=\mathscr{L}_{2}^{\prime}(\mathrm{b}) .
\end{aligned}
$$

For $\mathscr{G}^{\prime \prime}=\left\langle\Sigma_{0}^{\prime \prime}, \Sigma_{2}, \Sigma_{2}^{\prime}, \mathscr{L}, \mathscr{L}^{\prime},\left[s, s^{\prime}\right]\right\rangle$, we have

$$
\mathcal{O}\left(\mathscr{G}^{\prime \prime}\right)=\left\{\langle P, Q\rangle \mid\langle M, P\rangle \in \mathcal{O}(\mathscr{G}) \text { and }\langle M, Q\rangle \in \mathcal{O}\left(\mathscr{G}^{\prime}\right)\right\} .
$$

Corollary 5.22. Synchronous tree adjoining grammars can be represented by $2 D$-ACGs belonging to $\mathbf{G}_{\text {tree,tree }}(2,2(\mathrm{mon}), 2(\mathrm{mon}))$.
Proof. For a given STAG $G$, let $\mathscr{B}^{G}=\left\langle\mathcal{L}, T_{1}, T_{2}\right\rangle$ be the corresponding bimapping where $T_{1}$ and $T_{2}$ are linear monadic $\varepsilon$-free MTTs. By our construction, we have $\mathscr{G}^{T_{i}} \in \mathbf{G}_{\text {tree, tree }}(2,1(\mathrm{r}), 2(\mathrm{mon}))$ such that

$$
\mathcal{O}\left(\mathscr{G}^{T_{i}}\right)=\left\{\langle M, N\rangle \mid M \vdash_{T_{i}} N\right\}
$$

By applying Lemma 5.21 to $\mathscr{G}^{T_{1}}$ and $\mathscr{G}^{T_{2}}$, we get $\mathscr{G} \in \mathbf{G}_{\text {tree,tree }}(2,2($ mon $), 2(\mathrm{mon}))$ such that

$$
\mathcal{O}(\mathscr{G})=\left\{\left\langle N_{1}, N_{2}\right\rangle \mid M \vdash_{T_{1}} N_{1}, M \vdash_{T_{2}} N_{2}\right\}=\mathcal{R}\left(\mathscr{B}^{G}\right) .
$$

Remark 5.23. Though STAGs generate PTIME languages by Corollaries 5.10 and 5.22 , the universal membership problem for STAGs is NPComplete. The membership of NP is obvious. The NP-hardness can be
shown by reduction from the string rearrangement problem (see Definition 6.4 and Theorem 6.5). For the definition of STAGs and their derivations, see [50].

For an instance $\left\langle\mathrm{w},\left\langle\mathrm{w}_{1}, \ldots, \mathrm{w}_{n}\right\rangle\right\rangle$ of the string rearrangement problem, let an STAG consists of the following initial tree pair and auxiliary tree pairs:



It is clear that $\left\langle w_{1}, \ldots, w_{n}\right\rangle$ has a solution iff the following tree pair is derivable by the above STAG. The order of adjoining trees corresponds to the permutation on $\left\langle w_{1}, \ldots, w_{n}\right\rangle$.




This result makes a contrast to the fact that the universal membership problem for TAGs is in P .

### 5.5 Deterministic Tree Walking Transducers

A deterministic tree walking transducer [1] (DTWT) is a kind of finite state automaton that takes a tree as input and walks on the tree from a node to a node starting from the root node. Depending on the current state and the label on the current node, it deterministically decides which adjacent node it
should go to and which state it should enter. Each transition, it writes some string. When it gets out of the tree with a final state, the concatenation of the written strings is output. Thus, a DTWT defines a relation between trees and strings. A DTWT becomes a tree recognizer or string generator if we disregard its output or input, respectively. It is known that DTWTs as tree recognizers define a proper subclass of the regular tree languages [23] and that DTWTs as string generators are equivalent to LCFRSs [57]. Consequently it would be expected that the relation defined by DTWTs are definable by 2D-ACG belonging to $\mathbf{G}_{\text {tree,string }}(2,1,4)$, though it is not a corollary to the above results. In this section, we show that the expectation is correct by borrowing the idea from Weir's work [57].

## Definitions and Notations

For each tree $M$, we give the set $\operatorname{path}(M) \subseteq \mathbb{N}^{*}$ of paths of $M$ as

$$
\operatorname{path}\left(f M_{1} \ldots M_{n}\right)=\{\varepsilon\} \cup\left\{i \pi \mid \pi \in \operatorname{path}\left(M_{i}\right)\right\}
$$

and the subtree specified by $\pi \in \mathbb{N}^{*}$ is defined as

$$
\mathrm{f} M_{1} \ldots M_{n}: \pi= \begin{cases}\mathrm{f} M_{1} \ldots M_{n} & \text { if } \pi=\varepsilon \\ M_{i}: \pi^{\prime} & \text { if } \pi=i \pi^{\prime} .\end{cases}
$$

A deterministic tree walking transducer is a sextuple $T=\left\langle\Sigma_{1}, V, Q, \Delta, q_{s}, F\right\rangle$ where

- $\Sigma_{1}$ is a ranked alphabet, called the input alphabet,
- $V$ is an unranked alphabet, called the output alphabet,
- $Q$ is the set of states,
- $q_{s} \in Q$ is the initial state,
- $Q_{f} \subseteq Q$ is the set of final states,
- $\Delta$ is a partial function from $\Sigma_{1} \times Q$ to $V^{*} \times \mathbb{N} \times Q$, called the set of transition rules, such that if $\Delta(\mathrm{A}, q)=\left\langle\mathrm{w}, d, q^{\prime}\right\rangle$, then $0 \leq d \leq \operatorname{rank}(\mathrm{A})$.

A configuration of $T$ is a quadruple $\langle K, \pi, q, \mathbf{w}\rangle$ where $K \in \mathbb{T}\left(\Sigma_{1}\right)$ is the tree under consideration, $\pi \in \operatorname{path}(K) \cup\{\uparrow\}$ specifies a node in $K$ or is $\uparrow$ (where $\uparrow$ can be thought of as the parent of the root of $K$ ), $q \in Q$ is the current state, and $w \in V^{*}$ is the output string produced up to that point in the computation. We call a state $q$ an up-state for A if $\Delta(\mathrm{A}, q)=\left\langle\mathrm{v}, 0, q^{\prime}\right\rangle$ for
some $q^{\prime}$ and v and a down-state for A if $\Delta(\mathrm{A}, q)=\left\langle\mathrm{v}, d, q^{\prime}\right\rangle$ for some $d>0$. We write

$$
\langle K, \pi, q, \mathrm{w}\rangle \vdash_{T}\left\langle K, \pi^{\prime}, q^{\prime}, \mathrm{ww}^{\prime}\right\rangle
$$

if

- the head of $K: \pi$ is $\mathrm{A} \in \Sigma_{1}$,
- $\Delta(\mathrm{A}, q)=\left\langle\mathrm{w}^{\prime}, d, q^{\prime}\right\rangle$,
- either $\pi=\pi^{\prime} m$ for some $m$ or $\pi=\varepsilon \wedge \pi^{\prime}=\uparrow$ when $d=0$,
- $\pi^{\prime}=\pi d$ when $d>0$.

The relation defined by $T$ is
$\mathcal{R}(T)=\left\{\langle K, \mathbf{w}\rangle \mid K \in \mathbb{T}\left(\Sigma_{1}\right)\right.$ and $\left\langle K, \varepsilon, q_{s}, \varepsilon\right\rangle \vdash_{T}^{*}\langle K, \uparrow, q, \mathbf{w}\rangle$ for some $\left.q \in F\right\}$.
If $K$ is in the domain of $\mathcal{R}(T)$, we say that $K$ is accepted.
Example 5.24. Let a DTWT $T=\left\langle\Sigma_{1}, V, Q, \Delta, q_{0},\left\{q_{f}\right\}\right\rangle$ have the following transition rules in $\Delta$ :

$$
\begin{aligned}
& \left\langle\mathrm{B}, q_{0}\right\rangle \xrightarrow{\varepsilon}\left\langle 1, q_{0}\right\rangle, \quad\left\langle\mathrm{A}, q_{0}\right\rangle \xrightarrow{\mathrm{a}}\left\langle 1, q_{0}\right\rangle, \quad\left\langle\mathrm{C}, q_{0}\right\rangle \xrightarrow{\varepsilon}\left\langle 0, q_{1}\right\rangle, \quad\left\langle\mathrm{A}, q_{1}\right\rangle \xrightarrow{\varepsilon}\left\langle 0, q_{1}\right\rangle, \\
& \left\langle\mathrm{B}, q_{1}\right\rangle \xrightarrow{\boldsymbol{\varepsilon}}\left\langle 1, q_{2}\right\rangle,
\end{aligned}\left\langle\mathrm{A}, q_{2}\right\rangle \xrightarrow{\varepsilon}\left\langle 2, q_{2}\right\rangle,\left\langle\mathrm{B}, q_{2}\right\rangle \xrightarrow{\varepsilon}\left\langle 1, q_{2}\right\rangle,\left\langle\mathrm{D}, q_{2}\right\rangle \xrightarrow{\boldsymbol{c}}\left\langle 0, q_{3}\right\rangle,
$$

$T$ translates the following tree into the string aaabbcbcbcddd.

$\xrightarrow{T} \quad$ aaabbcbcbcddd

## Encoding of DTWTs by 2D-ACGs

First we construct a 1 D -ACG $\mathscr{G}_{1}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}_{1}, s\right\rangle$ whose language is the domain of the relation defined by a given DTWT, disregarding the output strings. Suppose that a tree $K$ is accepted by $T$. We can decorate each node of $K$ with the sequence of states of $T$ which $T$ enters when $T$ visits that


Figure 5.3: Tree decorated with valid histories
node. The tree in Example 5.24 is decorated as in Figure 5.3 according to the walking of $T$. We can give any decoration on a tree, but at most one of them represents the real walking of the automaton on the tree, since the automaton is deterministic. We call such a decoration correct. A tree admits exactly one correct decoration if the tree is accepted, and otherwise, it has no correct decoration. Though checking the correctness of a decoration on a tree can be done by walking the whole tree, here we present a method of checking the correctness of a decoration locally.

If a node labeled with A in a tree $K \in \mathbb{T}\left(\Sigma_{1}\right)$ is decorated with a sequence $\vec{p}$ of states, then $\vec{p}$ must end with an up-state $p$ for A , since $T$ terminates at the upper node $\uparrow$ of the root node. Moreover, by the determinacy of $T, \vec{p}$ contains no overlapped occurrences of the same state. If a state $p$ appears in $\vec{p}$ twice or more, then $T$ falls into a loop. We call a possibly empty sequence $\vec{p}$ of states a valid history of A if

- for every element $p$ of $\vec{p}, \Delta(\mathrm{~A}, p)$ is defined,
- no state occurs twice or more in the sequences,
- the last element (if exists) of $\vec{p}$ is an up-state.

In particular, if A has rank 0 , then every valid history $\vec{p}$ on A consists of up-states only. There are finitely many valid histories of $A$ for each $A$, since any state can appear in a valid history at most once.

Let us look at the following fragment of the decorated tree in Figure 5.3.


Now you can forget other parts of the tree. Here $q_{0} q_{1} q_{2} q_{3} q_{4}, q_{0} q_{2}, q_{2} q_{3}$ are respectively valid histories of A, C, B. Since the automaton starts from the root node, the automaton visits the parent node A before its children C , B. Let us analyze the relation between those three valid histories starting from the state $q_{0}$ on the node A . According to the transition rule $\left\langle\mathrm{A}, q_{0}\right\rangle \xrightarrow{\mathrm{a}}$ $\left\langle 1, q_{0}\right\rangle \in \Delta$, the automaton must go down to the first child C and the state $q_{0}$ is unchanged. Indeed, the first state of the decoration on the node $C$ is $q_{0}$. It is consistent. Then the transition rule $\left\langle\mathrm{C}, q_{0}\right\rangle \xrightarrow{\epsilon}\left\langle 0, q_{1}\right\rangle$ demands the automaton to go up to the parent A and to change the state into $q_{1}$. Indeed, the second state on A is $q_{1}$. Here the automaton should obey the rule $\left\langle\mathrm{A}, q_{1}\right\rangle \xrightarrow{\varepsilon}\left\langle 0, q_{1}\right\rangle$, and it disappears from our view. Though we cannot know whether the automaton comes back here unless we look at other parts of the tree, we assume that it comes back here with the third state $q_{2}$ on A , and continue checking the consistency as described so far. By $\left\langle\mathrm{A}, q_{2}\right\rangle \xrightarrow{\varepsilon}\left\langle 2, q_{2}\right\rangle$, the automaton arrives at B with the state $q_{2}$. Then it next goes down by the rule $\left\langle\mathrm{B}, q_{2}\right\rangle \xrightarrow{\varepsilon}\left\langle 1, q_{2}\right\rangle$. Again we assume the automaton comes back here with the next state $q_{3}$ on B , though we do not known whether the automaton really comes back here with that state. This consistency checking is illustrated by arrows as below, where all the states appearing in the valid histories are connected in sequence. This case we say it is a locally consistent combination of valid histories.


The following two are examples of inconsistent combinations of valid histories.
[C,

[C,


In the example on the left, though the sequence $q_{2} q_{0}$ is a valid history of C , it contradicts the transition rule $\left\langle\mathrm{A}, q_{0}\right\rangle \xrightarrow{\mathrm{a}}\left\langle 1, q_{0}\right\rangle$. In the example on the right, we cannot reach the last state $q_{2}$ on the node $\mathbf{C}$. These are inconsistent combinations of valid histories. For $A \in \Sigma_{1}^{(0)}$, which cannot have children, we also define a locally consistent combination of valid histories to be a valid history of A.

Let us turn our attention to the root node. It is easy to see that if the root node of an accepted tree is decorated as $\left[\mathrm{A}, q_{1} \ldots q_{k}\right]$, then

- the first element $q_{1}$ must be the initial state,
- if $i \neq k$, then $q_{i}$ is a down state of A ,
- $\left\langle\mathrm{A}, q_{k}\right\rangle \xrightarrow{\vee}\left\langle 0, q_{f}\right\rangle \in \Delta$ for some $q_{f} \in Q_{f}$.

We call a valid history $\vec{q}$ on A satisfying those conditions a root history of A. As we will prove later, for a given decorated input tree, if each fragment of the tree of height one is decorated by consistent history and the root node is labeled with a root history, then the whole decoration on the tree is correct, and thus the undecorated tree is accepted by the automaton. For a given DTWT $T$, we construct a 2D-ACG whose abstract language represents the set of correctly decorated trees on $\Sigma_{1}$ and whose first lexicon strips off the decoration of the tree.

Definition 5.25. For $\mathrm{A} \in \Sigma_{1}^{(n)}$ and $\mathrm{B}_{1}, \ldots, \mathrm{~B}_{n} \in \Sigma_{1}$, a locally consistent combination of valid histories of $\left\langle\mathrm{A}, \mathrm{B}_{1}, \ldots, \mathrm{~B}_{n}\right\rangle$ is a sequence $\sigma$ of pairs whose first component is a symbol in $\left\{\mathrm{A}, \mathrm{B}_{1}, \ldots, \mathrm{~B}_{n}\right\}$ and whose second component is a state in $Q$ such that

$$
\begin{aligned}
\sigma & =\sigma_{1} \ldots \sigma_{k} \\
\sigma_{i} & =\sigma_{i, 1} \ldots \sigma_{i, k_{i}}\left\langle\mathrm{~A}, p_{i, k_{i}+1}\right\rangle \\
\sigma_{i, j} & =\left\langle\mathrm{A}, p_{i, j}\right\rangle\left\langle\mathrm{B}_{d_{i, j}, j}, q_{i, j, 1}\right\rangle \ldots\left\langle\mathrm{B}_{d_{i, j}}, q_{i, j, k_{i, j}+1}\right\rangle
\end{aligned}
$$

where

- the sequence $\vec{p}=p_{1,1} \ldots p_{1, k_{1}+1} \ldots p_{k, 1} \ldots p_{k, k_{k}+1}$ is a valid history of A for each $1 \leq i \leq k$,
- $\vec{q}_{h}=\left\langle q_{i, j, g} \mid d_{i, j}=h, 1 \leq g \leq k_{i, j}+1\right\rangle^{2}$ (elements of $\vec{q}_{h}$ are paired with $\left.\mathrm{B}_{h}\right)$ is a valid history of $\mathrm{B}_{h}$

[^10]and moreover, for each $1 \leq i \leq k, 1 \leq j \leq k_{i}, 1 \leq g \leq k_{i, j}$,

- $\Delta\left(\mathrm{A}, p_{i, k_{i}+1}\right)=\langle\mathrm{v}, 0, r\rangle$ for some $\mathrm{v} \in V^{*}$ and $r \in Q$,
- $\Delta\left(\mathrm{A}, p_{i, j}\right)=\left\langle\mathrm{v}_{i, j}, d_{i, j}, q_{i, j, 1}\right\rangle$ with $1 \leq d_{i, j} \leq n$ and $\mathrm{v}_{i, j} \in V^{*}$,
- $\Delta\left(\mathrm{B}_{d_{i, j}}, q_{i, j, g}\right)=\langle\mathrm{v}, d, r\rangle$ for some $\mathrm{v} \in V^{*}, d>0$, and $r \in Q$,
- $\Delta\left(\mathrm{B}_{d_{i, j}}, q_{i, j, k_{i, j}+1}\right)=\left\langle\mathrm{v}_{i, j}^{\prime}, 0, p_{i, j+1}\right\rangle$ with $\mathrm{v}_{i, j}^{\prime} \in V^{*}$.

Now we define a higher-order signature $\Sigma_{0}$ and a lexicon $\mathscr{L}_{1}: \Sigma_{0} \rightarrow \Sigma_{1}$. Let
$\mathscr{A}_{0}=\{[\mathrm{A}, \vec{p}] \mid \vec{p}$ is a valid history of A$\} \cup\{s\}$
$\mathscr{C}_{0}=\left\{\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket \mid \sigma\right.$ is a locally consistent combination of valid
histories of $\left.A B_{1} \ldots B_{n}\right\}$
$\cup\{\llbracket \mathrm{A}, \vec{q} \rrbracket \mid \vec{q}$ is a root history of A$\}$
and $\tau_{0}$ be defined as

$$
\tau_{0}\left(\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket\right)=\left[\mathrm{B}_{1}, \vec{q}_{1}\right] \rightarrow \cdots \rightarrow\left[\mathrm{B}_{n}, \vec{q}_{n}\right] \rightarrow[\mathrm{A}, \vec{p}]
$$

where each $\vec{q}_{i}$ and $\vec{p}$ are determined as in Definition 5.25,

$$
\tau_{0}(\llbracket \mathrm{~A}, \vec{q} \rrbracket)=[\mathrm{A}, \vec{q}] \rightarrow s
$$

Let

$$
\begin{aligned}
\mathscr{L}_{1}([\mathrm{~A}, \vec{q}]) & =o, \\
\mathscr{L}_{1}(\llbracket \mathrm{AB}, \sigma \rrbracket) & =\mathrm{A}, \\
\mathscr{L}_{1}(\llbracket \mathrm{~A}, \vec{q} \rrbracket) & =\lambda z^{o} . z,
\end{aligned}
$$

We have established the correspondence between correctly decorated trees on $\mathbb{T}\left(\Sigma_{1}\right)$ and the abstract language of $\mathscr{G}_{1}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}_{1}, s\right\rangle$.

Now, let turn our attention to the output of the transducer. We say that a valid history $\vec{q}$ on A is a unitary history iff $\vec{q}$ contains exactly one up-state for A. If a valid history $\vec{q}$ contains $k$ up-states, then $\vec{q}$ is uniquely partitioned into $k$ unitary histories. Suppose that a subtree $K$ of a correctly decorated tree is rooted by $[\mathrm{A}, \vec{q}]$ such that $\vec{q}$ is partitioned into $k$ unitary histories. This means that the automaton visits the subtree $K$ exactly $k$ times. The automaton outputs some string $\mathrm{w}_{i}$ during the walking of the $i$-th visit. Let $M \in \Lambda\left(\Sigma_{0}\right)$ represent $K$. We want the second lexicon $\mathscr{L}_{2}: \Sigma_{0} \rightarrow \Sigma_{V}$ to map $M$ to

$$
\lambda x^{s t r^{k} \rightarrow s t r} \cdot x / \mathrm{w}_{1} / \ldots / \mathrm{w}_{k} / .
$$

This can be realized by defining $\mathscr{L}_{2}$ as follows.

$$
\begin{aligned}
\mathscr{L}_{2}(s) & =s t r, \\
\mathscr{L}_{2}([\mathrm{~A}, \vec{p}]) & =\left(s t r^{k} \rightarrow s t r\right) \rightarrow s t r \\
& \text { if } \vec{p} \text { is partitioned into } k \text { unitary histories of A. }
\end{aligned}
$$

Let $\mathrm{A}, \mathrm{B}_{1}, \ldots, \mathrm{~B}_{n}$, and $\sigma$ be as in Definition 5.25.

$$
\mathscr{L}_{2}\left(\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket\right)=\lambda y_{1} \ldots y_{n} x \cdot y_{1}\left(\lambda \vec{z}_{1} \cdot y_{2}\left(\lambda \vec{z}_{2} \ldots y_{n}\left(\lambda \vec{z}_{n} \cdot x N_{1} \ldots N_{k}\right) \ldots\right)\right)
$$

where $\vec{z}_{h}=\left\langle z_{i, j} \mid d_{i, j}=h\right\rangle$ and

$$
N_{i}=/ \mathrm{v}_{i, 1} /+z_{i, 1}+/ \mathrm{v}_{i, 1}^{\prime} /+\cdots+/ \mathrm{v}_{i, k_{i}} /+z_{i, k_{i}}+/ \mathrm{v}_{i, k_{i}}^{\prime} /
$$

For root histories, let

$$
\mathscr{L}_{2}(\llbracket \mathrm{~A}, \vec{q} \rrbracket)=\lambda y^{(s t r \rightarrow s t r) \rightarrow s t r} \cdot y\left(\lambda z^{s t r} \cdot z+/ \mathrm{v} /\right)
$$

if $\Delta(\mathrm{A}, q)=\left\langle\mathrm{v}, 0, q_{f}\right\rangle$ for some $q_{f} \in Q_{f}$ for the last element $q$ of $\vec{q}$.
We then define an ACG $\mathscr{G}^{T} \in \mathbf{G}_{\text {tree,string }}(2,1(\mathrm{sr}), 4)$ as

$$
\mathscr{G}^{T}=\left\langle\Sigma_{0}, \Sigma_{1}, \Sigma_{V}, \mathscr{L}_{1}, \mathscr{L}_{2}, s\right\rangle .
$$

Lemma 5.26. $\mathcal{O}\left(\mathscr{G}^{T}\right)=\mathcal{R}(T)$.
Proof. $\left[\mathcal{O}\left(\mathscr{G}^{T}\right) \subseteq \mathcal{R}(T)\right]$
First we show the following claim by induction on $M$ :
Suppose that $M \in \mathbb{T}\left(\Sigma_{0}\right)$ has type $[\mathrm{A}, \vec{p}] \in \mathscr{A}_{0}-\{s\}$. Let $\vec{p}$ be partitioned into $k$ unitary histories $\vec{p}_{1}, \ldots, \vec{p}_{k}$ of A . Then, there are $\mathrm{w}_{1}, \ldots, \mathrm{w}_{k} \in V^{*}$ such that

- $\mathscr{L}_{2}(M)=\lambda x^{\operatorname{str} \rightarrow s t r} \cdot x / \mathrm{w}_{1} / \ldots / \mathrm{w}_{k} /$,
- $\left\langle\mathscr{L}_{1}(M), \varepsilon, p_{i, 1}, \varepsilon\right\rangle \vdash_{T}^{*}\left\langle\mathscr{L}_{1}(M), \varepsilon, p_{i, k_{i}+1}, \mathrm{w}_{i}\right\rangle$ where $\vec{p}_{i}=p_{i, 1} \ldots p_{i, k_{i}+1}$.

Basis. Suppose that $M=\llbracket \mathrm{A},\left\langle\mathrm{A}, p_{1}\right\rangle \ldots\left\langle\mathrm{A}, p_{k}\right\rangle \rrbracket \in \mathscr{C}_{0}$ with $\mathrm{A} \in \Sigma_{1}^{(0)}$. Every $p_{i}$ is an up-state and thus a unitary history of A . By the definition,

- $\mathscr{L}_{2}\left(\llbracket \mathrm{~A},\left\langle\mathrm{~A}, p_{1}\right\rangle \ldots\left\langle\mathrm{A}, p_{k}\right\rangle \rrbracket\right)=\lambda x^{s t r^{k} \rightarrow s t r} \cdot x \overbrace{\mid \varepsilon / \ldots / \varepsilon /}^{k \text { times }}$,
- $\left\langle\mathrm{A}, \varepsilon, p_{i}, \varepsilon\right\rangle \vdash_{T}^{0}\left\langle\mathrm{~A}, \varepsilon, p_{i}, \varepsilon\right\rangle$.

Step. Suppose that $M$ is of the form $\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket M_{1} \ldots M_{n}$ and

$$
\tau_{0}\left(\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket\right)=\left[\mathrm{B}_{1}, \vec{q}_{1}\right] \rightarrow \cdots \rightarrow\left[\mathrm{B}_{n}, \vec{q}_{n}\right] \rightarrow[\mathrm{A}, \vec{p}] .
$$

$\sigma$ must be of the form

$$
\begin{aligned}
\sigma & =\sigma_{1} \ldots \sigma_{k} \\
\sigma_{i} & =\sigma_{i, 1} \ldots \sigma_{i, k_{i}}\left\langle\mathrm{~A}, p_{i, k_{i}+1}\right\rangle \\
\sigma_{i, j} & =\left\langle\mathrm{A}, p_{i, j}\right\rangle\left\langle\mathrm{B}_{d_{i, j}}, q_{i, j, 1}\right\rangle \ldots\left\langle\mathrm{B}_{d_{i, j}}, q_{i, j, k_{i, j}+1}\right\rangle
\end{aligned}
$$

such that

- the sequence $\vec{p}=p_{1,1} \ldots p_{1, k_{1}+1} \ldots p_{k, 1} \ldots p_{k, k_{k}+1}$ is a valid history of A , where each $p_{i, 1} \ldots p_{i, k_{i}+1}$ is a unitary history of A ,
- $\vec{q}_{h}=\left\langle q_{i, j, g} \mid d_{i, j}=h, 1 \leq g \leq k_{i, j}+1\right\rangle$ is a valid history of $\mathrm{B}_{h}$, where each $q_{i, j, 1} \ldots q_{i, j, k_{i, j}+1}$ is a unitary history of $\mathrm{B}_{h}$,
and moreover, for each $1 \leq i \leq k$ and $1 \leq j \leq k_{i}$,

$$
\begin{align*}
& \Delta\left(\mathrm{A}, p_{i, j}\right)=\left\langle\mathrm{v}_{i, j}, d_{i, j}, q_{i, j, 1}\right\rangle \text { with } 1 \leq d_{i, j} \leq n,  \tag{5.3}\\
& \Delta\left(\mathrm{~B}_{d_{i, j}}, q_{i, j, k_{i, j}+1}\right)=\left\langle\mathrm{v}_{i, j}^{\prime}, 0, p_{i, j+1}\right\rangle . \tag{5.4}
\end{align*}
$$

We apply the induction hypothesis to $M_{h}$ of type $\left[\mathrm{B}_{h}, \vec{q}_{h}\right]$ for each $1 \leq h \leq n$. If $\vec{q}_{h}$ is partitioned into $l_{h}$ unitary histories of $\mathrm{B}_{h}$, then there are strings $\mathrm{u}_{h, 1}, \ldots, \mathrm{u}_{h, l_{h}} \in V^{*}$ such that

$$
\begin{equation*}
\mathscr{L}_{2}\left(M_{h}\right)=\lambda x^{s t r^{l_{h} h \rightarrow s t r} \cdot x / \mathrm{u}_{h, 1} / \ldots / \mathrm{u}_{h, l_{h}} /, ., ~ . ~} \tag{5.5}
\end{equation*}
$$

and for each $i, j$ with $d_{i, j}=h$,

$$
\begin{equation*}
\left\langle\mathscr{L}_{1}\left(M_{h}\right), \varepsilon, q_{i, j, 1}, \varepsilon\right\rangle \vdash_{T}^{*}\left\langle\mathscr{L}_{1}\left(M_{h}\right), \varepsilon, q_{i, j, k_{i, j}+1}, \mathbf{w}_{i, j}\right\rangle \tag{5.6}
\end{equation*}
$$

where $\mathbf{w}_{i, j}=\mathrm{u}_{h, g}$ for the $g$-th pair $\langle i, j\rangle$ such that $d_{i, j}=h$, i.e., $\left\langle\mathbf{u}_{h, 1}, \ldots, \mathbf{u}_{h, l_{h}}\right\rangle=$ $\left\langle\mathrm{w}_{i, j} \mid d_{i, j}=h\right\rangle$. Thus, for each $j \leq k_{i}$ we get

$$
\begin{align*}
\left\langle\mathscr{L}_{1}(M), \varepsilon, p_{i, j}, \varepsilon\right\rangle & \vdash_{T}\left\langle\mathscr{L}_{1}(M), d_{i, j}, q_{i, j, 1}, \mathrm{v}_{i, j}\right\rangle  \tag{5.3}\\
& \vdash_{T}^{*}\left\langle\mathscr{L}_{1}(M), d_{i, j}, q_{i, j, k_{i, j}+1}, \mathrm{v}_{i, j} \mathbf{w}_{i, j}\right\rangle  \tag{5.6}\\
& \vdash_{T}\left\langle\mathscr{L}_{1}(M), \varepsilon, p_{i, j+1}, \mathrm{v}_{i, j} \mathbf{w}_{i, j} \mathrm{v}_{i, j}^{\prime}\right\rangle, \tag{5.4}
\end{align*}
$$

and therefore

$$
\left\langle\mathscr{L}_{1}(M), \varepsilon, p_{i, 1}, \varepsilon\right\rangle \vdash_{T}^{*}\left\langle\mathscr{L}_{1}(M), \varepsilon, p_{i, k_{i}+1}, \mathrm{w}_{i}\right\rangle
$$

for $\mathrm{w}_{i}=\mathrm{v}_{i, 1} \mathrm{w}_{i, 1} \mathrm{v}_{i, 1}^{\prime} \ldots \mathrm{v}_{i, k_{i}} \mathrm{w}_{i, k_{i}} \mathrm{v}_{i, k_{i}}^{\prime}$. On the other hand, by the definition of $\mathscr{L}_{2}$,

$$
\mathscr{L}_{2}\left(\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket\right)=\lambda y_{1} \ldots y_{n} x \cdot y_{1}\left(\lambda \vec{z}_{1} \cdot y_{2}\left(\lambda \vec{z}_{2} \ldots y_{n}\left(\lambda \vec{z}_{n} \cdot x N_{1} \ldots N_{k}\right) \ldots\right)\right)
$$

where $\vec{z}_{h}=\left\langle z_{i, j} \mid d_{i, j}=h\right\rangle$ and

$$
N_{i}=/ \mathrm{v}_{i, 1} /+z_{i, 1}+/ \mathrm{v}_{i, 1}^{\prime} /+\cdots+/ \mathrm{v}_{i, k_{i}} /+z_{i, k_{i}}+/ \mathrm{v}_{i, k_{i}}^{\prime} /
$$

Let $/ \overrightarrow{\mathbf{u}}_{h} /=\left\langle/ \mathbf{u}_{h, 1} /, \ldots, / \mathbf{u}_{h, l_{h}} /\right\rangle=\left\langle/ \mathbf{w}_{i, j} / \mid d_{i, j}=h\right\rangle$. Then,

$$
\begin{align*}
\mathscr{L}_{2}(M) & =\mathscr{L}_{2}\left(\left[\mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket\right)\left(\lambda x^{s t r^{l} \rightarrow s t r} \cdot x / \overrightarrow{\mathrm{u}}_{1} /\right) \ldots\left(\lambda x^{s t r^{l n} \rightarrow s t r} \cdot x / \overrightarrow{\mathrm{u}_{n}} /\right)\right.  \tag{5.5}\\
& (\text { by }(5 . \\
& =\lambda x \cdot x N_{1} \ldots N_{k}\left[/ \overrightarrow{\mathrm{u}}_{1} / / \vec{z}_{1}\right] \ldots\left[/ \overrightarrow{\mathrm{u}}_{n} / / \vec{z}_{n}\right] \\
& =\lambda x \cdot x N_{1} \ldots N_{k}\left[/ \mathrm{w}_{i, j} / / z_{i, j}\right]_{1 \leq j \leq i \leq k}^{11 \leq k_{i}} \\
& =\lambda x \cdot x / \mathrm{w}_{1} / \ldots / \mathrm{w}_{n} / .
\end{align*}
$$

We have completed the proof of the claim.
Suppose that $M \in \mathcal{A}\left(\mathscr{G}^{T}\right)$. Then, $M$ has the form $M=\llbracket \mathrm{A}, \vec{q} \rrbracket M^{\prime}$ for some $M^{\prime}$ of type $[\mathrm{A}, \vec{q}]$, where $\vec{q}$ is a root history of A of the form $\vec{q}=q_{s} q_{2} \ldots q_{k+1}$ such that

$$
\Delta\left(\mathrm{A}, q_{k+1}\right)=\langle\mathrm{v}, 0, q\rangle
$$

for some $q \in Q_{f}$ and $\mathbf{v} \in V^{*}$. By applying the above claim to $M^{\prime}$, we get

- $\mathscr{L}_{2}\left(M^{\prime}\right)=\lambda x^{s t r \rightarrow s t r} \cdot x / \mathrm{w} /$,
- $\left\langle\mathscr{L}_{1}\left(M^{\prime}\right), \varepsilon, q_{s}, \varepsilon\right\rangle \vdash_{T}^{*}\left\langle\mathscr{L}_{1}\left(M^{\prime}\right), \varepsilon, q_{k+1}, \mathrm{w}\right\rangle$.

Combining the above derivation with respect to $\mathscr{L}_{1}\left(M^{\prime}\right)$ and $\Delta\left(\mathrm{A}, q_{k+1}\right)=$ $\langle\mathbf{v}, 0, q\rangle$, we get

$$
\left\langle\mathscr{L}_{1}\left(M^{\prime}\right), \varepsilon, q_{s}, \varepsilon\right\rangle \vdash_{T}^{*}\left\langle\mathscr{L}_{1}\left(M^{\prime}\right), \uparrow, q, \mathrm{wv}\right\rangle,
$$

and thus we have $\left\langle\mathscr{L}_{1}\left(M^{\prime}\right), \mathrm{wv}\right\rangle \in \mathcal{R}(T)$. Since $\mathscr{L}_{1}(\llbracket \mathrm{~A}, \vec{q} \rrbracket)=\lambda x . x$ and $\mathscr{L}_{2}(\llbracket \mathrm{~A}, \vec{q} \rrbracket)=\lambda y \cdot y(\lambda z \cdot z+/ \mathrm{v} /)$, we get

$$
\begin{aligned}
& \mathscr{L}_{1}\left(\llbracket \mathrm{~A}, \vec{q} \rrbracket M^{\prime}\right)=\mathscr{L}_{1}\left(M^{\prime}\right) \\
& \mathscr{L}_{2}\left(\llbracket \mathrm{~A}, \vec{q} \rrbracket M^{\prime}\right)=/ \mathrm{wv} / .
\end{aligned}
$$

We obtain $\mathcal{O}\left(\mathscr{G}^{T}\right) \subseteq \mathcal{R}(T)$.
$\left[\mathcal{R}(T) \subseteq \mathcal{O}\left(\mathscr{G}^{T}\right)\right]$

Let $K^{\prime}$ be a subtree of the correctly decorated tree of an accepted tree, [A,$\vec{p}]$ the label of the root of $K^{\prime}, \vec{p}=\vec{p}_{1} \ldots \vec{p}_{k}$, and $\vec{p}_{i}=p_{i, 1} \ldots p_{i, k_{i}+1}$ where each $\vec{p}_{i}$ is a unitary history of A . Then, for each $1 \leq i \leq k$ we have

$$
\left\langle K, \varepsilon, p_{i, 1}, \varepsilon\right\rangle \vdash_{T}^{*}\left\langle K, \varepsilon, p_{i, k_{i}+1}, \mathrm{w}_{i}\right\rangle
$$

where $K \in \Lambda\left(\Sigma_{1}\right)$ is the tree obtained from $K^{\prime}$ by stripping off the decoration with states. We first prove that for such $K, \vec{p}$, and $\mathrm{w}_{1}, \ldots, \mathrm{w}_{k}$, there is $M \in \Lambda\left(\Sigma_{0}\right)$ such that

- $\tau_{0}(M)=[\mathrm{A}, \vec{p}]$,
- $\mathscr{L}_{1}(M)=K$,
- $\mathscr{L}_{2}(M)=\lambda x^{s t r^{k} \rightarrow s t r} \cdot x / \mathrm{w}_{1} / \ldots / \mathrm{w}_{k} /$
by induction on $K$.
Basis. If $K=\mathrm{A} \in \Sigma_{1}^{(0)}$ and $K^{\prime}=[\mathrm{A}, \vec{p}]$ where $\vec{p}=p_{1} \ldots p_{k}$ is a valid history of $A$, then we have

$$
\left\langle K, \varepsilon, p_{i}, \varepsilon\right\rangle \vdash_{T}^{0}\left\langle K, \varepsilon, p_{i}, \varepsilon\right\rangle
$$

for $1 \leq i \leq k$. By the definition, we have an abstract constant

$$
\begin{aligned}
& \llbracket \mathrm{A},\left\langle\mathrm{~A}, p_{1}\right\rangle \ldots\left\langle\mathrm{A}, p_{k}\right\rangle \rrbracket \in \mathscr{C}_{0} \text { such that } \\
& \tau_{0}\left(\llbracket \mathrm{~A},\left\langle\mathrm{~A}, p_{1}\right\rangle \ldots\left\langle\mathrm{A}, p_{k}\right\rangle \rrbracket\right)=[\mathrm{A}, \vec{p}] \\
& \mathscr{L}_{1}\left(\llbracket \mathrm{~A},\left\langle\mathrm{~A}, p_{1}\right\rangle \ldots\left\langle\mathrm{A}, p_{k}\right\rangle \rrbracket\right)=\mathrm{A}=K, \\
& \mathscr{L}_{2}\left(\llbracket \mathrm{~A},\left\langle\mathrm{~A}, p_{1}\right\rangle \ldots\left\langle\mathrm{A}, p_{k}\right\rangle \rrbracket\right)=\lambda x^{\text {str }} \rightarrow \text { str } x \underbrace{\mid \varepsilon / \ldots / \varepsilon /}_{k \text { times }} .
\end{aligned}
$$

Step. Suppose that a tree $K \in \mathbb{T}\left(\Sigma_{1}\right)$ and a sequence $\vec{p}$ of states are such that $K=\mathrm{A}\left(\mathrm{B}_{1} \vec{K}_{1}\right) \ldots\left(\mathrm{B}_{n} \vec{K}_{n}\right), \vec{p}=p_{1,1} \ldots p_{1, k_{1}+1} \ldots p_{k, 1} \ldots p_{k, k_{k}+1}$ where each $p_{i, 1} \ldots p_{i, k_{i}+1}$ is a unitary history of A , and

$$
\begin{equation*}
\left\langle K, \varepsilon, p_{i, 1}, \varepsilon\right\rangle \vdash_{T}^{*}\left\langle K, \varepsilon, p_{i, k_{i}+1}, \mathrm{w}_{i}\right\rangle \tag{5.7}
\end{equation*}
$$

for each $1 \leq i \leq k$.
For each $j \leq k_{i}$, since $p_{i, j}$ is a down-state for A, we have $\Delta\left(\mathrm{A}, p_{i, j}\right)=$ $\left\langle\mathrm{v}_{i, j}, d_{i, j}, r_{i, j}\right\rangle$ for some $\mathrm{v}_{i, j} \in V^{*}, d_{i, j}>0, r_{i, j} \in Q$, and $T$ goes down to the $d_{i, j}$-th child of A.

$$
\begin{equation*}
\left\langle K, \varepsilon, p_{i, j}, \varepsilon\right\rangle \vdash_{T}\left\langle K, d_{i, j}, r_{i, j}, v_{i, j}\right\rangle \tag{5.8}
\end{equation*}
$$

Since $T$ must come back from $\mathrm{B}_{d_{i, j}}$ to the node A with the state $p_{i, j+1}$, there are $r_{i, j}^{\prime} \in Q$ and $\mathrm{v}_{i, j}^{\prime} \in V^{*}$ such that $\Delta\left(\mathrm{B}_{d_{i, j}}, r_{i, j}^{\prime}\right)=\left\langle\mathrm{v}_{i, j}^{\prime}, 0, p_{i, j+1}\right\rangle$, i.e.,

$$
\begin{equation*}
\left\langle K, d_{i, j}, r_{i, j}^{\prime}, \varepsilon\right\rangle \vdash_{T}\left\langle K, \varepsilon, p_{i, j+1}, \mathrm{v}_{i, j}^{\prime}\right\rangle \tag{5.9}
\end{equation*}
$$

Consider the move of $T$ from when it is at $\mathrm{B}_{d_{i, j}}$ with the state $r_{i, j}$ to when it is at $\mathrm{B}_{d_{i, j}}$ with the state $r_{i, j}^{\prime}$. That is,

$$
\begin{equation*}
\left\langle K, d_{i, j}, r_{i, j}, \varepsilon\right\rangle \vdash_{T}^{*}\left\langle K, d_{i, j}, r_{i, j}^{\prime}, \mathbf{w}_{i, j}\right\rangle \tag{5.10}
\end{equation*}
$$

for some $\mathrm{w}_{i, j} \in V^{*}$. By (5.7)-(5.10), we get the equation

$$
\begin{equation*}
\mathbf{w}_{i}=\mathbf{v}_{i, 1} \mathbf{w}_{i, 1} \mathbf{v}_{i, 1}^{\prime} \ldots \mathrm{v}_{i, k_{i}} \mathbf{w}_{i, k_{i}} \mathrm{v}_{i, k_{i}}^{\prime} \tag{5.11}
\end{equation*}
$$

Suppose that the $h$-th child of A is decorated as $\left[\mathrm{B}_{h}, \vec{q}_{h}\right]$ in $K^{\prime} . \vec{q}_{d_{i, j}}$ must contain as a subsequence a unitary history of $\mathrm{B}_{d_{i, j}}$ corresponding to (5.10). Let the unitary history be $\vec{q}_{i, j}=q_{i, j, 1} \ldots q_{i, j, k_{i, j}+1}$, where $q_{i, j, 1}=r_{i, j}$, $q_{i, j, k_{i, j}+1}=r_{i, j}^{\prime}$. We see $\vec{q}_{h}=\left\langle q_{i, j, g} \mid d_{i, j}=h, 1 \leq g \leq k_{i, j}+1\right\rangle$. By applying the induction hypothesis to $\mathrm{B}_{h} \vec{K}_{h}$ and $\vec{q}_{h}$, we get $M_{h} \in \Lambda\left(\Sigma_{0}\right)$ such that

$$
\begin{align*}
\tau_{0}\left(M_{h}\right) & =\left[\mathrm{B}_{h}, \vec{q}_{h}\right], \\
\mathscr{L}_{1}\left(M_{h}\right) & =\mathrm{B}_{h} \vec{K}_{h}, \\
\mathscr{L}_{2}\left(M_{h}\right) & =\lambda x \cdot x\left\langle/ \mathrm{w}_{i, j} /\right\rangle_{d_{i, j}=h} . \tag{5.12}
\end{align*}
$$

Let $\sigma$ be defined as

$$
\begin{aligned}
\sigma & =\sigma_{1} \ldots \sigma_{k} \\
\sigma_{i} & =\sigma_{i, 1} \ldots \sigma_{i, k_{i}}\left\langle\mathrm{~A}, p_{i, k_{i}+1}\right\rangle \\
\sigma_{i, j} & =\left\langle\mathrm{A}, p_{i, j}\right\rangle\left\langle\mathrm{B}_{d_{i, j}}, q_{i, j, 1}\right\rangle \ldots\left\langle\mathrm{B}_{d_{i, j}}, q_{i, j, k_{i, j}+1}\right\rangle .
\end{aligned}
$$

Since $\sigma$ is a locally consistent combination of valid histories of $\left\langle\mathrm{A}, \mathrm{B}_{1}, \ldots, \mathrm{~B}_{n}\right\rangle$, we have an abstract constant $\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket$ such that

$$
\begin{aligned}
& \tau_{0}\left(\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket\right)=\left[\mathrm{B}_{1}, \vec{q}_{1}\right] \rightarrow \cdots \rightarrow\left[\mathrm{B}_{n}, \vec{q}_{n}\right] \rightarrow[\mathrm{A}, \vec{p}] \\
& \mathscr{L}_{1}\left(\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket\right)=\mathrm{A} \\
& \mathscr{L}_{2}\left(\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket\right)=\lambda y_{1} \ldots y_{n} x \cdot y_{1}\left(\lambda \vec{z}_{1} \ldots y_{n}\left(\lambda \vec{z}_{n} \cdot x N_{1} \ldots N_{k}\right) \ldots\right) \\
& \text { where } N_{i}=/ \mathrm{v}_{i, 1} /+z_{i, 1}+/ \mathrm{v}_{i, 1}^{\prime} /+\cdots+/ \mathrm{v}_{i, k_{i}} /+z_{i, k_{i}}+/ \mathrm{v}_{i, k_{i}}^{\prime} / \\
& \quad \text { and } \vec{z}_{h}=\left\langle z_{i, j} \mid d_{i, j}=h\right\rangle .
\end{aligned}
$$

Thus, for the abstract term $M=\llbracket \mathrm{AB}_{1} \ldots \mathrm{~B}_{n}, \sigma \rrbracket M_{1} \ldots M_{n} \in \Lambda\left(\Sigma_{0}\right)$, we have

$$
\begin{align*}
\tau_{0}(M) & =[\mathrm{A}, \vec{p}] \\
\mathscr{L}_{1}(M) & =\mathrm{A}\left(\mathrm{~B}_{1} \vec{K}_{1}\right) \ldots\left(\mathrm{B}_{n} \vec{K}_{n}\right) \\
\mathscr{L}_{2}(M) & =\lambda x \cdot \mathscr{L}_{2}\left(M_{1}\right)\left(\lambda \vec{z}_{1} \ldots \mathscr{L}_{2}\left(M_{n}\right)\left(\lambda \vec{z}_{n} \cdot x N_{1} \ldots N_{k}\right) \ldots\right) \\
& =\lambda x \cdot x N_{1} \ldots N_{k}\left[/ \mathrm{w}_{i, j} / / z_{i, j}\right]_{1 \leq i \leq k}^{1 \leq i \leq k_{i}}  \tag{5.12}\\
& =\lambda x \cdot x / \mathrm{w}_{1} / \ldots / \mathrm{w}_{k} / . \tag{5.11}
\end{align*}
$$

Suppose that $\langle K, w\rangle \in \mathcal{R}(T)$ for some tree $K \in \mathbb{T}\left(\Sigma_{1}\right)$ and $\mathbf{w} \in V^{*}$. Let $\vec{q}=q_{s} q_{2} \ldots q_{k+1}$ be the root history of the root label A of $K$ obtained from the derivation

$$
\left\langle K, \varepsilon, q_{s}, \varepsilon\right\rangle \vdash_{T}^{*}\left\langle K, \varepsilon, q_{k+1}, \mathrm{w}^{\prime}\right\rangle \vdash_{T}\left\langle K, \uparrow, q, \mathrm{w}^{\prime} \mathrm{v}\right\rangle
$$

where $q \in Q_{f}$ and $\mathbf{w}=w^{\prime} \mathbf{v}$. By applying the above claim to the derivation

$$
\left\langle K, \varepsilon, q_{s}, \varepsilon\right\rangle \vdash_{T}^{*}\left\langle K, \varepsilon, q_{k+1}, \mathrm{w}^{\prime}\right\rangle,
$$

we get a $\lambda$-term $M$ such that

- $\tau_{0}(M)=[\mathrm{A}, \vec{q}]$,
- $\mathscr{L}_{1}(M)=K$,
- $\mathscr{L}_{2}(M)=\lambda x \cdot x / \mathrm{w}^{\prime} /$.

Recall that we have the abstract constant $\llbracket \mathrm{A}, \vec{q} \rrbracket$ such that

- $\tau_{0}(\llbracket \mathrm{~A}, \vec{q} \rrbracket)=[\mathrm{A}, \vec{q}] \rightarrow s$,
- $\mathscr{L}_{1}(\llbracket \mathrm{~A}, \vec{q} \rrbracket)=\lambda x . x$,
- $\mathscr{L}_{2}(\llbracket \mathrm{~A}, \vec{q} \rrbracket)=\lambda y \cdot y(\lambda z \cdot z+/ \mathrm{v} /)$.

Thus, for $\llbracket \mathrm{A}, \vec{q} \rrbracket M \in \mathcal{A}\left(\mathscr{G}^{T}\right)$, we get $\mathscr{L}_{1}(\llbracket \mathrm{~A}, \vec{q} \rrbracket M)=K$ and $\mathscr{L}_{2}(\llbracket \mathrm{~A}, \vec{q} \rrbracket M)=$ $/ \mathrm{w}^{\prime} \mathrm{v} /=/ \mathrm{w} /$. This completes the proof.

Proposition 5.27. For every deterministic tree walking transducer $T$, there is a $2 D-A C G \mathscr{G}^{T} \in \mathbf{G}_{\text {tree,string }}(2,1(\mathrm{r}), 4)$ such that $\mathcal{O}\left(\mathscr{G}^{T}\right)=\mathcal{R}(T)$.
Proof. The lexicon $\mathscr{L}_{1}$ defined above is a semi-relabeling. Abstract constants representing root histories are mapped to the identity: $\mathscr{L}_{1}(\llbracket \mathrm{~A}, \vec{q} \rrbracket)=\lambda x^{o} . x$. These constants can be eliminated as we have done in Section 4.3.3 to eliminate unary nonlexical constants from second-order ACGs. For each constant $\llbracket \mathrm{AB}, \sigma \rrbracket$ of type $\vec{\alpha} \rightarrow[\mathrm{A}, \vec{q}]$, we add a new constant of type $\vec{\alpha} \rightarrow s$ that is mapped to $\mathscr{L}_{i}(\lambda \vec{x} \cdot \llbracket \mathrm{~A}, \vec{q} \rrbracket(\llbracket \mathrm{AB}, \sigma \rrbracket \vec{x}))$ by the lexicon $\mathscr{L}_{i}$. Then we can get rid of $\llbracket A, \vec{q} \rrbracket$ from the grammar.
Remark 5.28. In the original definition [1], a DTWT has a context-free grammar as its parameter and takes only derivation trees of the CFG as input. The above proposition still holds if input is restricted to elements of a regular tree language $\mathcal{L}$. Let $\mathscr{G}^{\mathcal{L}}=\left\langle\Sigma_{0}, \Sigma_{1}, \mathscr{L}, s\right\rangle \in \mathbf{G}_{\text {tree }}(2,1(\mathrm{r}))$ represent the regular tree language $\mathcal{L}$ and $\mathscr{G}^{T} \in \mathbf{G}_{\text {tree,string }}(2,1(\mathrm{r}), 4)$ encoding the tree transducer whose input is not restricted. Applying Lemma 5.21 to the 2DACGs $\mathscr{G} \mathcal{L L}=\left\langle\Sigma_{0}, \Sigma_{1}, \Sigma_{1}, \mathscr{L}, \mathscr{L}, s\right\rangle$ and $\mathscr{G}^{T} \in \mathbf{G}_{\text {tree,string }}(2,1(\mathrm{r}), 4)$, we obtain the desired 2D-ACG.

Now we obtain the two-dimensional version of Salvati's Theorem (Theorem 2.6). It is easy to extend the following result to second-order $k \mathrm{D}$-ACGs for any $k \in \mathbb{N}$.

Corollary 5.29. For every second-order $2 D-A C G \mathscr{G} \in \mathbf{G}_{\text {string,string }}\left(2, n_{1}, n_{2}\right)$, there is an equivalent $2 D-A C G \mathscr{G}^{\prime} \in \mathbf{G}_{\text {string,string }}(2,4,4)$.

Proof. In the proof of Salvati's Theorem, for a given second-order string 1 D-ACG $\mathscr{G}_{0} \in \mathbf{G}_{\text {string }}(2, n)$, he constructs a DTWT $T_{0}$ as a string language generator whose output language coincides with the object language of $\mathscr{G}_{0}$, where input trees for $T_{0}$ are restricted to elements of $\mathcal{A}\left(\mathscr{G}_{0}\right) .^{3}$

Let $\mathscr{G}=\left\langle\Sigma_{0}, \Sigma_{1}, \Sigma_{2}, \mathscr{L}_{1}, \mathscr{L}_{2}, s\right\rangle \in \mathrm{G}_{\text {string }}\left(2, n_{1}, n_{2}\right)$. For the $i$-th projection $\mathscr{G}_{i}=\left\langle\Sigma_{0}, \Sigma_{i}, \mathscr{L}_{i}, s\right\rangle$ of $\mathscr{G}$, let $T_{i}=\left\langle\Sigma_{0}, \Sigma_{i}, Q_{i}, q_{i}, F\right\rangle$ be the DTWT obtained by Salvati's method whose input is restricted to a regular tree language $\mathcal{A}(\mathscr{G})$. That is,

$$
\mathcal{R}\left(T_{i}\right)=\left\{\left\langle M, P_{i}\right\rangle \mid M \in \mathcal{A}(\mathscr{G}), P_{i}=\mathscr{L}_{i}(M)\right\} .
$$

By applying our method to DTWTs $T_{i}$, we get 2D-ACGs $\mathscr{G}_{i}^{\prime} \in \mathrm{G}_{\text {tree,string }}$ $(2,1(r), 4)$ such that

$$
\mathcal{O}\left(\mathscr{G}_{i}\right)=\mathcal{R}\left(T_{i}\right) .
$$

By Lemma 5.21 , we get a $2 \mathrm{D}-\mathrm{ACG} \mathscr{G}^{\prime} \in \mathbf{G}(2,4,4)$ such that

$$
\begin{aligned}
\mathcal{O}\left(\mathscr{G}^{\prime}\right) & =\left\{\left\langle P_{1}, P_{2}\right\rangle \mid\left\langle M, P_{i}\right\rangle \in \mathcal{O}\left(\mathscr{G}_{i}^{\prime}\right)=\mathcal{R}\left(T_{i}\right)\right\} \\
& =\left\{\left\langle P_{1}, P_{2}\right\rangle \mid M \in \mathcal{A}(\mathscr{G}), P_{i}=\mathscr{L}_{i}(M)\right\}=\mathcal{O}(\mathscr{G}) .
\end{aligned}
$$

The converse of Proposition 5.27 does not hold, since there is a regular tree language that cannot be recognized by any DTWT [23].

### 5.6 Linearization of Affine Two-Dimensional ACGs

In Chapter 3, we gave two linearization methods for affine 1D-ACG (Theorem 3.16). One of them is valid for $2 \mathrm{D}-\mathrm{ACGs}$ as well.

Theorem 5.30. For every affine $2 D$-dimensional $A C G \mathscr{G} \in \mathbf{G}^{\text {aff }}\left(m, n_{1}, n_{2}\right)$, there is a linear $2 D$-dimensional $A C G \mathscr{G}^{\prime \prime} \in \mathbf{G}^{\operatorname{lin}}\left(m, \max \left\{2, n_{1}\right\}, \max \left\{2, n_{2}\right\}\right)$ such that $\mathcal{O}\left(\mathscr{G}^{\prime \prime}\right)=\mathcal{O}(\mathscr{G}) \cap\left(\Lambda^{\operatorname{lin}}\left(\Sigma_{1}\right) \times \Lambda^{\operatorname{lin}}\left(\Sigma_{2}\right)\right)$.

[^11]Proof. Let $\mathscr{G}_{i}=\left\langle\Sigma_{0}, \Sigma_{i}, \mathscr{L}_{i}, s\right\rangle$ be the $i$-th projection of the given affine ACG $\mathscr{G}$, and $\mathscr{G}_{i}^{\prime \prime}=\left\langle\Sigma_{i, 0}, \Sigma_{i}, \mathscr{L}_{i}^{\prime \prime},\left[s, \mathscr{L}_{i}(s)\right]\right\rangle$ the linearized form of $\mathscr{G}_{i}$ obtained by our method. Recall that we have a relabeling lexicon $\mathscr{L}_{0, i}: \Sigma_{0, i} \rightarrow \Sigma_{0}$ that satisfies the equation (3.5), i.e.,

$$
\left\{M \in \mathcal{A}\left(\mathscr{G}_{i}\right)\left||\mathscr{L}(M)|_{\beta} \text { is linear }\right\}=\left\{\mathscr{L}_{0, i}(N) \mid N \in \mathcal{A}\left(\mathscr{G}_{i}^{\prime \prime}\right)\right\} .\right.
$$

Let $\mathscr{G}_{i}^{\prime}=\left\langle\Sigma_{i, 0}, \Sigma_{0}, \Sigma_{i}, \mathscr{L}_{0, i}, \mathscr{L}_{i}^{\prime \prime},\left[s, \mathscr{L}_{i}(s)\right]\right\rangle$ for $i=1,2$. By applying Lemma 5.21 to two linear 2D-ACGs $\mathscr{G}_{i}^{\prime}$ with $i=1,2$, we get a 2 D -ACG $\mathscr{G}^{\prime \prime}$ such that

$$
\mathcal{O}\left(\mathscr{G}^{\prime \prime}\right)=\mathcal{O}(\mathscr{G}) \cap\left(\Lambda^{\operatorname{lin}}\left(\Sigma_{1}\right) \times \Lambda^{\operatorname{lin}}\left(\Sigma_{2}\right)\right)
$$

We easily apply this method to any affine $k \mathrm{D}$-ACGs with $k \in \mathbb{N}$.
On the other hand, Theorem 3.8 does not hold for two-dimensional secondorder affine ACGs, as the following counterexample shows. Let $\mathscr{G} \in \mathrm{G}_{\text {tree,tree }}^{\text {aff }}$ $(2,1,1)$ consist of the following lexical entries:

| $\mathrm{x} \in \mathscr{C}_{0}$ | $\tau_{0}(\mathrm{x})$ | $\mathscr{L}_{1}(\mathrm{x})$ | $\mathscr{L}_{2}(\mathrm{x})$ |
| :---: | :---: | :---: | :---: |
| A | $s$ | $\#$ | $\#$ |
| B | $s \rightarrow s \rightarrow s$ | $\lambda x^{\circ} y^{\circ} . \#$ | $\lambda x^{\circ} y^{\circ} . \mathrm{c} x y$ |

Then,

$$
\mathcal{O}(\mathscr{G})=\{\#\} \times \mathbb{T}\left(\Sigma_{2}\right) .
$$

Suppose that there is a linear ACG $\mathscr{G}^{\prime} \in \mathbf{G}_{\text {tree, tree }}^{\operatorname{lin}}(2,1,1)$ generating this language. Without loss of generality, we assume that each vocabulary contains neither useless constants nor useless atomic types. For $\mathscr{G}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \Sigma_{2}, \mathscr{L}_{1}^{\prime}\right.$, $\left.\mathscr{L}_{2}^{\prime}, s^{\prime}\right\rangle$, let $\mathscr{G}_{1}^{\prime}=\left\langle\Sigma_{0}^{\prime}, \Sigma_{1}, \mathscr{L}_{1}^{\prime}, s^{\prime}\right\rangle$. By $\mathcal{O}\left(\mathscr{G}_{1}^{\prime}\right)=\{\#\}$ and thus $\mathscr{C}_{1}^{\prime}=\{\#\}$, we easily see that every abstract constant a $\in \mathscr{C}_{0}^{\prime}$ has nullary or unary type, because the only closed linear terms in $\Lambda^{\operatorname{lin}}\left(\Sigma_{1}^{\prime}\right)$ of second-order types are $\lambda x^{o} . x$ and \#. Let

$$
\begin{array}{r}
h=\max \left(\left\{\operatorname{height}\left(\left|\mathscr{L}_{2}^{\prime}(\mathrm{a})\right|_{\beta}\right) \mid \mathrm{a} \in \mathscr{C}_{0}^{\prime} \text { and } \tau_{0}^{\prime}(\mathrm{a})=p \in \mathscr{A}_{0}^{\prime}\right\}\right. \\
\left.\cup\left\{\operatorname{height}\left(\left|\mathscr{L}_{2}^{\prime}(\mathrm{a}) \#\right|_{\beta}\right) \mid \mathrm{a} \in \mathscr{C}_{0}^{\prime} \text { and } \tau_{0}^{\prime}(\mathrm{a})=p \rightarrow q\right\}\right)
\end{array}
$$

where the height of a tree is defined as usual. It is easy to see that $\mathrm{c} P P \in$ $\mathbb{T}\left(\Sigma_{2}\right)-\mathcal{O}_{2}\left(\mathscr{G}^{\prime}\right)$ for any tree $P \in \mathbb{T}\left(\Sigma_{2}\right)$ of height $h$.

### 5.7 Summary

Although symmetric treatment of syntax and semantics of natural language is an important appeal of the ACG formalism as de Groote [15] states in his first paper on the ACG, the mathematical properties of two-dimensional extensions of ACGs have not been studied well so far. This chapter has investigated the mathematical properties of 2D-ACGs, particularly the generative capacity of second-order 2D-ACGs. We have evaluated the generative capacity of 2D-ACGs by encoding existing two-dimensional formalisms, namely, linear macro tree transducers (linear MTTs), synchronous tree adjoining grammars and deterministic tree walking transducers.

Table 5.1 summarizes the generative capacity of second-order 2D-ACGs revealed in this chapter and preceding research, which has already shown that finite state transducers [15] and synchronous tree substitution grammars [58] are encodable by 2D-ACGs. The fact that several transducers, which define relations of two languages in asymmetric ways, are encoded by ACGs, which define relations of two languages in symmetric ways, can be thought of as a variation of Nivat's characterization of finite transducers by bimorphisms. Through simulation of transducers, which model syntax-directed semantics, by 2D-ACGs, we have shown how rich the expressive power of 2D-ACGs is.

The encoding methods for linear tree transducers, linear MTTs, and synchronous tree substitution grammars are indeed straightforward as well as de Groote and Pogodalla's encodings for the variety of context-free formalisms. Therefore, ACGs generalize two-dimensional formalisms as well as one-dimensional ones.

There are several subclasses of 2D-ACGs whose generative capacity remains unclear. It is future work to characterize the expressive power of those subclasses.

Table 5.1: Hierarchy of second-order two-dimensional ACGs

| Finite State Transducers | $\subseteq$ | $\mathbf{G}_{\text {string,string }}(2,2,2)$ |
| :---: | :---: | :--- |
| Linear Tree Transducers | $=$ | $\mathbf{G}_{\text {tree,tree }}(2,1(\mathrm{sr}), 1)$ |
| $\varepsilon$-free Linear Tree Transducers | $=$ | $\mathbf{G}_{\text {tree,tree }}(2,1(\mathrm{r}), 1)$ |
| Linear Macro Tree Transducers | $=$ | $\mathbf{G}_{\text {tree,tree }}(2,1(\mathrm{sr}), 2)$ |
| $\varepsilon$-free Linear Macro Tree Transducers | $=$ | $\mathbf{G}_{\text {tree,tree }}(2,1(\mathrm{r}), 2)$ |
| Synchronous Tree Substitution Grammars | $\subseteq$ | $\mathbf{G}_{\text {tree,tree }}(2,1,1)$ |
| Synchronous Tree Adjoining Grammars | $\subseteq$ | $\mathbf{G}_{\text {tree,tree }}(2,2(\mathrm{mon}), 2(\mathrm{mon}))$ |
| Deterministic Tree Walking Transducers | $\subseteq$ | $\mathbf{G}_{\text {tree,string }}(2,1(\mathrm{r}), 4)$ |
| $\mathbf{G}_{\text {string,string }}\left(2, n_{1}, n_{2}\right)$ for $n_{1}, n_{2} \geq 4$ | $=$ | $\mathbf{G}_{\text {string,string }}(2,4,4)$ |

## Chapter 6

## Higher-Order Interpolation in the Linear Lambda Calculus

The main result of this chapter has been published as [59].

### 6.1 Introduction

## Parsing with ACGs and Interpolation

As pointed out by de Groote [15] and Pogodalla [42], parsing with ACGs is closely related to higher-order interpolation in the linear lambda calculus, though they are still different problems. Let $M$ be a member of the abstract language $\mathcal{A}(\mathscr{G})$ of an ACG $\mathscr{G}$, and $M_{0}$ be obtained by replacing all occurrences of constants by fresh variables $x_{1}, \ldots, x_{n}$. Then, $M$ can be represented as

$$
\begin{equation*}
M=\left(\lambda x_{1} \ldots x_{n} \cdot M_{0}\right) \mathrm{a}_{1} \ldots \mathrm{a}_{n} \in \mathcal{A}(\mathscr{G}) \tag{6.1}
\end{equation*}
$$

for constants $a_{1}, \ldots, a_{n}$ appearing in $M$ (multiple occurrences of the same constant are allowed). By the lexicon $\mathscr{L}$ of $\mathscr{G}$, the equation (6.1) becomes

$$
\begin{equation*}
\mathscr{L}(M)=P=P_{0} \mathscr{L}\left(\mathrm{a}_{1}\right) \ldots \mathscr{L}\left(\mathrm{a}_{n}\right) \in \mathcal{O}(\mathscr{G}) \tag{6.2}
\end{equation*}
$$

where $P_{0}$ is identical to $\mathscr{L}\left(\lambda x_{1} \ldots x_{n} . M_{0}\right)$ except for the type assignment. The problem of checking whether $P \in \Lambda\left(\Sigma_{1}\right)$ belongs to $\mathcal{O}(\mathscr{G})$ thus becomes that of finding a linear combinator $\lambda x_{1} \ldots x_{n} . M_{0}$ and constants $\mathrm{a}_{1}, \ldots, \mathrm{a}_{n}$ satisfying (6.2). If each $\mathrm{a}_{i}$ and thus $P_{i}=\mathscr{L}\left(\mathrm{a}_{i}\right)$ is known, it reduces to finding a linear combinator $X$ such that

$$
P=X P_{1} \ldots P_{n} .
$$

This problem is called higher-order interpolation in the linear lambda calculus, which is a restricted form of higher-order matching. For instance, if a string ACG has no lexical ambiguity, that is, for each object constant c, there is a unique abstract constant a such that $\mathscr{L}(\mathrm{a})$ contains c , then we can uniquely determine the constants $\mathrm{a}_{1}, \ldots, \mathrm{a}_{n}$ that should appear in $M \in \mathcal{A}(\mathscr{G})$, immediately from the given object term $P$. De Groote [14] has shown the matching problem in the linear lambda calculus is in NP, while it was open in the general case when he introduced ACGs. This gives an explanation why de Groote [15] defines the ACG formalism on linear lambda calculus. Motivated by this perspective, this chapter is devoted to investigating the complexity of the higher-order interpolation problem in the linear lambda calculus. We discuss the problem beyond the application of parsing with ACGs.

## Higher-Order Matching in the Linear Lambda Calculus

While the second-order unification problem modulo $\beta$ and $\beta \eta$ [12] and the sixth-order matching problem modulo $\beta$ are undecidable [32], Stirling [54] recently shows that matching problem modulo $\beta \eta$ is decidable. On the other hand, there are some results on the complexity of matching in the linear lambda calculus. These results hold for both $\beta$ and $\beta \eta$-matching. De Groote [14] shows that matching in the linear lambda calculus is NPcomplete, and by extending his result, Dougherty and Wierzbicki [5] show that matching in the affine lambda calculus is NP-complete. We call matching linear if the linearity is imposed on occurrences of unknowns, i.e., each unknown occurs exactly once in the problem instances. Salvati and de Groote [46] prove that NP-hardness of the second-order linear matching problem in the linear lambda calculus. Besides, Salvati [44] give an algorithm for solving linear matching in the linear lambda calculus. There is another result on matching involving the linearity given by Levy [30]. He gives some conditions under which second-order unification with the restriction that a solution must substitute linear $\lambda$-terms for unknowns is decidable.

In Salvati and de Groote's paper [46], they also discuss the interpolation problem. Regrettably, Salvati and de Groote's proof of NP-completeness of third-order interpolation in the linear lambda calculus contains an error. In this chapter we correct the flaw and prove NP-completeness of third-order interpolation in the linear lambda calculus.

While no free variable occurs twice or more in a linear $\lambda$-term, the number of occurrences of constants is not constrained. Since constants behave like free variables, it is natural to asks whether NP-hardness still holds when we exclude constants from problem instances. This chapter shows that fourth-
order interpolation in the linear lambda calculus is NP-complete even in the absence of constants. Therefore, multiple occurrences of constants do not play an essential role for NP-hardness of higher-order matching in the linear lambda calculus.

### 6.2 NP-Complete Problems

Definition 6.1. Let $\mathcal{V}$ be a finite set of Boolean variables. Introducing a twin $\neg \mathcal{V}=\{\neg v \mid v \in \mathcal{V}\}$ of $\mathcal{V}$, we call elements of $\mathcal{V}$ and $\neg \mathcal{V}$ positive literals and negative literals respectively. We then define the set of literals as $\mathcal{V} \cup \neg \mathcal{V}$. A formula in conjunctive normal form (CNF) $\mathcal{F}$ on $\mathcal{V}$ is a collection of clauses which are non-empty subsets of $\mathcal{V} \cup \neg \mathcal{V}$. A valuation $\psi$ on $\mathcal{V}$ is a mapping from $\mathcal{V} \cup \neg \mathcal{V}$ to $\{0,1\}$ such that $\psi(v)+\psi(\neg v)=1$ for all $v \in \mathcal{V}$. A clause $\mathcal{C} \in \mathcal{F}$ is satisfied by a valuation $\psi$ via a literal $w \in \mathcal{V} \cup \neg \mathcal{V}$ iff $w \in \mathcal{C}$ and $\psi(w)=1$. A CNF $\mathcal{F}$ is satisfied by a valuation $\psi$ iff every $\mathcal{C} \in \mathcal{F}$ is satisfied by $\psi$. A CNF $\mathcal{F}$ is satisfiable iff there is $\psi$ that satisfies $\mathcal{F}$.

The satisfiability problem is the problem of deciding whether a given CNF $\mathcal{F}$ is satisfiable or not. It is well known that the satisfiability problem is NPcomplete [4].

Definition 6.2. An $m P-C N F \mathcal{F}$ is a CNF such that each positive literal $v \in \mathcal{V}$ occurs exactly $m$ times in $\mathcal{F}$. An $n N-C N F \mathcal{F}$ is a CNF such that each negative literal $\neg v \in \neg \mathcal{V}$ occurs exactly $n$ times in $\mathcal{F}$.

A CNF $\mathcal{F}$ is polarized iff each clause $\mathcal{C} \in \mathcal{F}$ contains only positive literals or only negative literals. We call a clause $\mathcal{C}$ positive if $\mathcal{C} \subseteq \mathcal{V}$, and negative if $\mathcal{C} \subseteq \neg \mathcal{V}$.

By combining the above definitions we define an $m \mathrm{P} n \mathrm{~N}-\mathrm{CNF}$ to be a CNF which is at the same time an $m \mathrm{P}-\mathrm{CNF}$ and an $n \mathrm{~N}-\mathrm{CNF}$.

We call variants of the satisfiability problem whose instances are restricted to CNFs in a special form with the modifier expressing that form.

Theorem 6.3. The polarized 2P1N-satisfiability problem is NP-complete.
Proof. For a given CNF $\mathcal{F}$ on $\mathcal{V}$, we may assume that for each $v \in \mathcal{V}$, the positive literal $v$ occurs exactly the same number of times as the negative literal $\neg v$ in $\mathcal{F}$. Otherwise, if $v$ occurs $m$ times more than $\neg v$, we add appropriate number of clauses $\left\{\neg v, u_{i}, \neg u_{i}\right\}$ to $\mathcal{F}$ for $1 \leq i \leq m$ where $u_{i}$ are new Boolean variables. If $v$ occurs $m$ times more than $\neg v$, add clauses $\left\{v, u_{i}, \neg u_{i}\right\}$ for $1 \leq i \leq m$. Now, for a given CNF $\mathcal{F}$ on $\mathcal{V}=\left\{v_{1}, \ldots, v_{l}\right\}$ such that the positive literal $v_{i}$ occurs exactly the same number of times as the
negative literal $\neg v_{i}$ for each Boolean variable $v_{i} \in \mathcal{V}$, we construct a polarized 2P1N-CNF $\mathcal{F}^{\prime}$ on $\mathcal{V}^{\prime}$ such that $\mathcal{F}$ is satisfiable iff $\mathcal{F}^{\prime}$ is satisfiable. Let $m_{i}$ be the number of occurrences of the positive literal $v_{i}$ (thus of the negative literal $\neg v_{i}$ ) for each $v_{i} \in \mathcal{V}$.

1. Let $\mathcal{V}^{\prime}=\left\{v_{i, j}, \bar{v}_{i, j} \mid 1 \leq i \leq l, 1 \leq j \leq m_{i}\right\}$.
2. Replace the $j$-th occurrence of the positive literal $v_{i}$ in $\mathcal{F}$ with $v_{i, j}$, and the $j$-th occurrence of the negative literal $\neg v_{i}$ in $\mathcal{F}$ with $\bar{v}_{i, j}$ for $1 \leq j \leq m_{i}$.
3. Add the clauses $\left\{v_{i, j}, \bar{v}_{i, j}\right\}$ for $1 \leq j \leq m_{i},\left\{\neg v_{i, j}, \neg \bar{v}_{i, j+1}\right\}$ for $1 \leq j<$ $m_{i}$, and $\left\{\neg v_{i, m_{i}}, \neg \bar{v}_{i, 1}\right\}$.

By the construction, $\mathcal{F}^{\prime}$ is a polarized $2 \mathrm{P} 1 \mathrm{~N}-\mathrm{CNF}$. It is clear that if $\mathcal{F}$ is satisfied by a valuation $\psi$, then $\mathcal{F}^{\prime}$ is also satisfied by $\psi^{\prime}$ such that $\psi^{\prime}\left(v_{i, j}\right)=$ $\psi\left(v_{i}\right)$ and $\psi^{\prime}\left(\bar{v}_{i, j}\right)=\psi\left(\neg v_{i}\right)$ for all $j \in\left\{1, \ldots, m_{i}\right\}$.

Conversely, suppose that $\mathcal{F}^{\prime}$ is satisfied by a valuation $\psi^{\prime}$. We show that $\psi^{\prime}\left(v_{i, j}\right)=\psi^{\prime}\left(v_{i, 1}\right)$ and $\psi^{\prime}\left(\bar{v}_{i, j}\right)=\psi^{\prime}\left(\neg v_{i, 1}\right)$ for all $j \in\left\{1, \ldots, m_{i}\right\}$. Then we easily see that the valuation $\psi$ such that $\psi\left(v_{i}\right)=\psi^{\prime}\left(v_{i, 1}\right)$ satisfies $\mathcal{F}$.

If $\psi^{\prime}\left(v_{i, 1}\right)=1$, then $\psi^{\prime}\left(\bar{v}_{i, 2}\right)=0$ by $\left\{\neg v_{i, 1}, \neg \bar{v}_{i, 2}\right\} \in \mathcal{F}^{\prime} . \operatorname{By}\left\{v_{i, 2}, \bar{v}_{i, 2}\right\} \in \mathcal{F}^{\prime}$, we see $\psi^{\prime}\left(v_{i, 2}\right)=1$. This way we see that $\psi^{\prime}\left(v_{i, j}\right)=1$ and $\psi^{\prime}\left(\bar{v}_{i, j}\right)=0$ for $2 \leq j \leq m_{i}$. By $\left\{\neg v_{i, m_{i}}, \neg \bar{v}_{i, 1}\right\} \in \mathcal{F}^{\prime}$, we get $\psi^{\prime}\left(\bar{v}_{i, 1}\right)=0$.

If $\psi^{\prime}\left(v_{i, 1}\right)=0$, then $\psi^{\prime}\left(\bar{v}_{i, 1}\right)=1$ by $\left\{v_{i, 1}, \bar{v}_{i, 1}\right\} \in \mathcal{F}^{\prime}$. By $\left\{\neg v_{i, m_{i}}, \neg \bar{v}_{i, 1}\right\} \in$ $\mathcal{F}^{\prime}$, we see $\psi^{\prime}\left(v_{i, m_{i}}\right)=0$. Similarly to the previous case, we see that $\psi^{\prime}\left(v_{i, j}\right)=$ 0 and $\psi^{\prime}\left(\bar{v}_{i, j}\right)=1$ for all $j \in\left\{1, \ldots, m_{i}\right\}$.

Definition 6.4. The string rearrangement problem is the problem of determining whether there is a permutation $\pi$ on $\{1, \ldots, m\}$ such that $w=$ $\mathrm{w}_{\pi(1)} \ldots \mathrm{w}_{\pi(m)}$ for a given string w and a sequence of string $\left\langle\mathrm{w}_{1}, \ldots, \mathrm{w}_{m}\right\rangle$.

Theorem 6.5. The string rearrangement problem is NP-complete.
Proof. Clearly the problem is in NP. The NP-hardness is shown by a reduction from the polarized ${ }^{1} 2 \mathrm{P} 1 \mathrm{~N}$-satisfiability problem. Let $\mathcal{F}=\left\{\mathcal{C}_{1}, \ldots, \mathcal{C}_{m}\right.$, $\left.\mathcal{D}_{1}, \ldots, \mathcal{D}_{n}\right\}$ be a polarized $2 \mathrm{P} 1 \mathrm{~N}-\mathrm{CNF}$ on $\mathcal{V}=\left\{v_{1}, \ldots, v_{l}\right\}$ where each $\mathcal{C}_{j}$ is a positive clause and each $\mathcal{D}_{k}$ is a negative clause. Since for any valuation $\psi$, for each variable $v_{i} \in \mathcal{V}$, there are at most two clauses that are satisfied by $\psi$ via $v_{i}$ or $\neg v_{i}$, we can assume that $2 l \geq m+n$. If $2 l<m+n$ (it is decidable in polynomial time), $\mathcal{F}$ is not satisfiable.

[^12]We define strings $\mathrm{w}, \mathrm{x}_{i}, \mathrm{y}_{j}$ for $1 \leq i \leq m, 1 \leq j \leq n$ on the alphabet $V=\left\{\mathrm{c}_{1}, \ldots, \mathrm{c}_{m}, \mathrm{~d}_{1}, \ldots, \mathrm{~d}_{n}, \mathrm{f}\right\}$ by

$$
\mathrm{w}=\mathrm{v}_{1} \ldots \mathrm{v}_{l} \text { with } \mathrm{v}_{i}=\mathrm{fc}_{\mu(i, 1)} \mathrm{fd}_{\nu(i)} \mathrm{f}_{\mu(i, 2)} \mathrm{f}
$$

where $\mu(i, h)=j$ if the $h$-th occurrence of $v_{i}$ is in $\mathcal{C}_{j}$ for $h=1,2$,
and $\nu(i)=j$ if the unique occurrence of $\neg v_{i}$ is in $\mathcal{D}_{j}$,

$$
\begin{aligned}
\mathrm{x}_{i} & =\mathrm{f}_{i} \mathrm{f} \text { for } 1 \leq i \leq m, \\
\mathrm{y}_{i} & =\mathrm{fd}_{i} \mathrm{f} \text { for } 1 \leq i \leq n .
\end{aligned}
$$

We see that

$$
\#_{\mathrm{a}}(\mathrm{w}) \geq \sum_{1 \leq i \leq m} \#_{\mathrm{a}}\left(\mathrm{x}_{i}\right)+\sum_{1 \leq i \leq n} \#_{\mathrm{a}}\left(\mathrm{y}_{i}\right)
$$

for all $\mathrm{a} \in V$. If a is either $\mathrm{c}_{j}$ or $\mathrm{d}_{k}$, then $\sum_{1 \leq i \leq m} \#_{\mathrm{a}}\left(\mathrm{x}_{i}\right)+\sum_{1 \leq i \leq n} \#_{\mathrm{a}}\left(\mathrm{y}_{i}\right)=$ 1. Since each clause is not empty, $\#_{a}(\mathrm{w}) \geq 1$. If $\mathrm{a}=\mathrm{f}$, then $\#_{\mathrm{a}}(\mathrm{w})=$ $4 l \geq \sum_{1 \leq i \leq m} \#_{\mathrm{a}}\left(\mathrm{x}_{i}\right)+\sum_{1 \leq i \leq n} \#_{\mathrm{a}}\left(\mathrm{y}_{i}\right)=2(m+n)$ by the assumption that $2 l \geq m+n$. Therefore, because the length of $w$ is $7 l$ and the sum of the lengths of $\mathrm{x}_{i}$ and $\mathrm{y}_{j}$ for all $i \in\{1, \ldots, m\}$ and $j \in\{1, \ldots, n\}$ is $3(m+n)$, one can find $\mathrm{z}_{1}, \ldots, \mathrm{z}_{7 l-3(m+n)} \in V$ such that

$$
\#_{\mathrm{a}}(\mathrm{w})=\sum_{1 \leq i \leq m} \#_{\mathrm{a}}\left(\mathrm{x}_{i}\right)+\sum_{1 \leq i \leq n} \#_{\mathrm{a}}\left(\mathrm{y}_{i}\right)+\sum_{1 \leq i \leq 7 l-3(m+n)} \#_{\mathrm{a}}\left(\mathrm{z}_{i}\right)
$$

for all $\mathrm{a} \in V$.
We show that there is a valuation $\psi$ that satisfies $\mathcal{F}$ iff there is a permutation on the sequence $\left\langle\mathrm{x}_{1}, \ldots, \mathrm{x}_{m}, \mathrm{y}_{1}, \ldots, \mathrm{y}_{n}, \mathrm{z}_{1}, \ldots, \mathrm{z}_{7 l-3(m+n)}\right\rangle$ that rearranges it to coincide with $w$.

First we show the "if" direction. Suppose that $\left\langle\mathrm{x}_{1}, \ldots, \mathrm{x}_{m}, \mathrm{y}_{1}, \ldots, \mathrm{y}_{n}, \mathrm{z}_{1}\right.$, $\left.\ldots, \mathrm{z}_{7 l-3(m+n)}\right\rangle$ can be reordered to be equivalent to w . Then, each $\mathrm{x}_{j}$ for $1 \leq j \leq m$ is used as a substring of $\mathrm{v}_{i}$ for some $i$. and each $\mathrm{y}_{j}$ for $1 \leq j \leq n$ is used as a substring of $\mathrm{v}_{i}$ for some $i$. We define a valuation $\psi$ on $\mathcal{V}$ as follows:

$$
\psi\left(v_{i}\right)= \begin{cases}1 & \text { if } \mathbf{x}_{j} \text { is used as a substring of } \mathbf{v}_{i} \text { for some } j \\ 0 & \text { if } \mathbf{y}_{k} \text { is used as a substring of } \mathbf{v}_{i} \text { for some } k, \\ \text { any value } & \text { otherwise. }\end{cases}
$$

Indeed $\psi$ is well-defined. We cannot use both $\mathrm{x}_{j}$ and $\mathrm{y}_{k}$ as substrings of $\mathrm{v}_{i}$ at the same time, because of the occurrences of the symbol f . Each clause $\mathcal{C}_{j}$ is satisfied by $\psi$ with the literal $v_{i}$ if $\mathbf{x}_{j}$ is used as a substring of $\mathbf{v}_{i}$, and $\mathcal{D}_{k}$ is satisfied by $\psi$ with the literal $\neg v_{i}$ if $\mathbf{y}_{k}$ is used as a substring of $\mathbf{v}_{i}$.

Conversely, if $\psi$ satisfies $\mathcal{F}$, one can find $\phi: \mathcal{F} \rightarrow \mathcal{V} \cup \neg \mathcal{V}$ such that $\phi(\mathcal{C}) \in \mathcal{C}$ and $\psi(\phi(\mathcal{C}))=1$ for all $\mathcal{C} \in\left\{\mathcal{C}_{1}, \ldots, \mathcal{C}_{m}, \mathcal{D}_{1}, \ldots, \mathcal{D}_{n}\right\}$. It never happens that $\phi\left(\mathcal{C}_{j}\right)=v_{i}$ and $\phi\left(\mathcal{D}_{k}\right)=\neg v_{i}$. Thus we can use each string $\mathrm{x}_{j}$ as a substring of $\mathrm{v}_{i}$ if $\phi\left(\mathcal{C}_{j}\right)=v_{i}$, and each string $\mathrm{y}_{k}$ as a substring of $\mathrm{v}_{i}$ if $\phi\left(\mathcal{D}_{k}\right)=v_{i}$. By the definition of $\mathbf{z}_{i}$ for $1 \leq i \leq 7 l-3(m+n)$, we can reorder $\mathrm{x}_{1}, \ldots, \mathrm{x}_{m}, \mathrm{y}_{1}, \ldots, \mathrm{y}_{n}, \mathrm{z}_{1}, \ldots, \mathrm{z}_{7 l-3(m+n)}$ to be equivalent to w .

### 6.3 Definitions

In this chapter, we extend the notion of $\lambda$-term introducing new atomic terms, called unknowns. In order to distinguish the extended class of $\lambda$ terms from usual unknown-free $\lambda$-terms, we call the latter pure terms. The set of unknowns are denoted by $\mathscr{U}$ and the letters $\mathbf{X}$ and $\mathbf{Y}$ are used for unknowns. Unknowns can be considered as a kind of variables that are never bound, and should be substituted for by pure terms.

The definitions of $\Lambda(\Sigma)$ and the type $\tau(M)$ of a term $M \in \Lambda(\Sigma)$ are now extended as follows:

- For every $a \in \mathscr{C}, a \in \Lambda(\Sigma)$.
- For every $x \in \mathscr{X}$ and $\alpha \in \mathcal{T}(\mathscr{A}), x^{\alpha} \in \Lambda(\Sigma)$ and $\tau\left(x^{\alpha}\right)=\alpha$.
- For every $\mathbf{X} \in \mathscr{U}$ and $\alpha \in \mathcal{T}(\mathscr{A}), \mathbf{X}^{\alpha} \in \Lambda(\Sigma)$ and $\tau\left(\mathbf{X}^{\alpha}\right)=\alpha$.
- For $M, N \in \Lambda(\Sigma)$, if $\tau(M)=(\alpha \rightarrow \beta), \tau(N)=\alpha$, then $(M N) \in \Lambda(\Sigma)$ and $\tau((M N))=\beta$.
- For $x \in \mathscr{X}, \alpha \in \mathcal{T}(\mathscr{A})$ and $M \in \Lambda(\Sigma),\left(\lambda x^{\alpha} \cdot M\right) \in \Lambda(\Sigma)$ and $\tau\left(\left(\lambda x^{\alpha} \cdot M\right)\right)=(\alpha \rightarrow \tau(M))$.

Other notions related to $\lambda$-terms, such as closed terms, linear terms, etc., are defined in the same way as in Chapter 2. Note that a closed term can contain unknowns and that unknowns may occur any number of times in a linear term.

Definition 6.6. A unification equation is a pair of closed terms $\langle L, R\rangle$ of the same type. Let $\overrightarrow{\mathbf{X}}$ be the unknowns in $\langle L, R\rangle$. A substitution $\sigma=[\vec{N} / \overrightarrow{\mathbf{X}}]$ where $\vec{N}$ is a sequence of pure terms is a solution for $\langle L, R\rangle$ modulo $\beta$ iff $L \sigma={ }_{\beta} R \sigma . \sigma$ is a solution for $\langle L, R\rangle$ modulo $\beta \eta$ iff $L \sigma={ }_{\beta \eta} R \sigma$. A matching equation is a unification equation $\langle L, R\rangle$ such that $R$ is a pure term. An interpolation equation is a matching equation $\langle L, R\rangle$ such that $L$ has just one occurrence of just one unknown as its head, i.e., $L \equiv \mathbf{X} L_{1} \ldots L_{m}$, where
each $L_{i}$ contains no unknowns. The order of a unification equation is defined to be the maximum of the orders of the types of the unknowns appearing in the equation. The unification, matching, and interpolation problem modulo $\beta(\beta \eta)$ are decision problems which ask whether or not there is a solution modulo $\beta(\beta \eta)$ for given unification, matching, interpolation equations respectively. The unification, matching, and interpolation problem in the linear lambda calculus allow only linear $\lambda$-terms as problem instances and solutions.

### 6.4 Interpolation in the Linear Lambda Calculus

While every interpolation equation $\left\langle\mathbf{X} L_{1} \ldots L_{m}, R\right\rangle$ has a trivial solution $\left[\lambda x_{1} \ldots x_{m} . R / \mathbf{X}\right]\left(x_{i} \notin \operatorname{Fv}(R)\right)$ in the (general) lambda calculus, the interpolation problem in the linear lambda calculus is not trivial. That the interpolation problem in the linear lambda calculus is in NP is an immediate corollary of the following theorem.

Theorem 6.7 (de Groote [14]). The matching problem in the linear lambda calculus is NP-complete.

Theorem 6.8 (Salvati and de Groote [46]). Second-order matching in the linear lambda calculus is NP-complete even if every unknown appears exactly once.

Salvati and de Groote [46, Proposition 3] have claimed that third-order interpolation in the linear lambda calculus is NP-complete by a reduction from the 1 N -satisfiability problem (Theorem 6.3). For a given $1 \mathrm{~N}-\mathrm{CNF}$ $\mathcal{F}=\left\{\mathcal{C}_{1}, \ldots, \mathcal{C}_{m}\right\}$ on $\mathcal{V}=\left\{v_{1}, \ldots, v_{l}\right\}$, they define a third-order interpolation equation $\langle L, R\rangle$ on $\Sigma=\langle\mathscr{A}, \mathscr{C}, \tau\rangle$ as follows: ${ }^{2}$

$$
\begin{aligned}
& \mathscr{A}=\{o\}, \mathscr{C}=\left\{\mathrm{a}, \mathrm{c}_{j}, \mathrm{~g} \mid 1 \leq j \leq m\right\}, \\
& \quad \tau(\mathrm{a})=o, \tau\left(\mathrm{c}_{j}\right)=o^{m} \rightarrow o, \tau(\mathrm{~g})=o^{l} \rightarrow o, \\
& L \equiv \mathbf{X}\left(\lambda x_{1}^{o} \ldots x_{m}^{o} \cdot \mathrm{c}_{1} x_{1} \ldots x_{m}\right) \ldots\left(\lambda x_{1}^{o} \ldots x_{m}^{o} \cdot \mathrm{c}_{m} x_{1} \ldots x_{m}\right), \\
& R \equiv \mathrm{~g} V_{1} \ldots V_{l} \text { for } V_{i} \equiv N_{i} P_{i, 1} \ldots P_{i, m} \\
& \quad \text { where } N_{i} \equiv \mathrm{c}_{j} \text { for } \neg v_{i} \in \mathcal{C}_{j} \text { and } P_{i, j} \equiv \begin{cases}\mathrm{c}_{j} \mathrm{a}^{m} & \text { if } v_{i} \in \mathcal{C}_{j}, \\
\mathrm{a} & \text { otherwise. } .\end{cases}
\end{aligned}
$$

[^13]We present a counterexample to their claim that $\mathcal{F}$ is satisfiable iff $\langle L, R\rangle$ has a solution. Consider their reduction from the following $1 \mathrm{~N}-\mathrm{CNF} \mathcal{F}$ :

$$
\begin{aligned}
& \text { Instance } \mathcal{F}: \quad \mathcal{C}_{1}=\left\{v_{1}\right\}, \mathcal{C}_{2}=\left\{\neg v_{1}\right\} \\
& \text { Reduction }\langle L, R\rangle: L \equiv \mathbf{X}\left(\lambda x_{1} x_{2} \cdot \mathrm{c}_{1} x_{1} x_{2}\right)\left(\lambda x_{1} x_{2} . \mathrm{c}_{2} x_{1} x_{2}\right) \\
& R \equiv \mathrm{~g}\left(\mathrm{c}_{2}\left(\mathrm{c}_{1} \text { aa) }\right) \mathrm{a}\right)
\end{aligned}
$$

If their reduction is correct, $\langle L, R\rangle$ has no solution, since $\mathcal{F}$ is not satisfiable. In fact, however, $\left[\lambda y_{1}^{o^{2} \rightarrow o} y_{2}^{o^{2} \rightarrow o} \cdot \mathrm{~g}\left(y_{2}\left(y_{1} \mathrm{aa}\right) \mathrm{a}\right) / \mathbf{X}\right]$ is a solution for $\langle L, R\rangle$.

Actually, the proposition that third-order interpolation in the linear lambda calculus is NP-complete is an easy corollary to the NP-hardness of the string rearrangement problem.

Proposition 6.9. Third-order interpolation modulo $\beta(\beta \eta)$ in the linear lambda calculus is NP-complete even if all the constants, variables (other than the unknown) and arguments of the unknown in the problem instances have types o or $o \rightarrow o$.

Proof. Let the pair of $w$ and $\left\langle w_{1}, \ldots, w_{m}\right\rangle$ on an alphabet be an instance of the string rearrangement problem. Clearly it has a solution iff the interpolation equation

$$
\left\langle\mathbf{X} / \mathrm{w}_{1} / \ldots / \mathrm{w}_{m} /, / \mathrm{w} /\right\rangle
$$

admits a solution.
It is not hard to show that the second-order interpolation problem in the linear lambda calculus is in P. If a given instance $\left\langle\mathbf{X} L_{1} \ldots L_{m}, R\right\rangle$ of the interpolation problem in the linear lambda calculus is second-order, then all $L_{i}$ have atomic types. $\left\langle\mathbf{X} L_{1} \ldots L_{m}, R\right\rangle$ has a solution $\left[\lambda z_{1} \ldots z_{m} . S / \mathbf{X}\right]$ iff $R$ can be represented as $R \equiv S\left[L_{1} / z_{1}, \ldots, L_{m} / z_{m}\right]$, because substituting a term of an atomic type for a variable produces no new $\beta$-redex. Thus, the following procedure determines whether $\left\langle\mathbf{X} L_{1} \ldots L_{m}, R\right\rangle$ admits a solution.

```
let \(\mathcal{L}:=\left\{L_{1}, \ldots, L_{m}\right\}\) and \(S:=R\);
while \(\mathcal{L} \neq \varnothing\) do
    take \(L_{i} \in \mathcal{L}\) such that \(L_{i}\) is not a proper subterm of any elements of \(\mathcal{L}\);
    let \(\mathcal{L}:=\mathcal{L}-\left\{L_{i}\right\}\);
    if there is \(S^{\prime}\) such that \(S=S^{\prime}\left[L_{i} / z_{i}\right]\) and \(z_{i}\) occurs exactly once in \(S^{\prime}\)
        then let \(S:=S^{\prime}\)
        else output "No" and halt;
    end if
end while
output "YES" \(\left(\lambda z_{1} \ldots z_{m} . S\right.\) is a solution);
```

In the next section, we will show that interpolation in the linear lambdacalculus is still NP-hard even in the absence of constants. For this purpose, we present another proof for the NP-hardness of third-order interpolation in the linear lambda calculus by a reduction from the polarized 1 N -satisfiability problem in the remainder of this section. The reduction is an elaboration of the one by Salvati and de Groote. We then eliminate constants from the new reduction with a certain technique, which does not work for the simple reduction in the proof for Proposition 6.9.

Definition 6.10. Let $\mathcal{F}=\left\{\mathcal{C}_{1}, \ldots, \mathcal{C}_{m}, \mathcal{D}_{1}, \ldots, \mathcal{D}_{n}\right\}$ be a polarized 1N-CNF on $\mathcal{V}=\left\{v_{1}, \ldots, v_{l}\right\}$ where each $\mathcal{C}_{j}$ is a positive clause and each $\mathcal{D}_{k}$ is a negative clause. First we introduce the following two functions $\mu:\{1, \ldots, l\} \times$ $\{1, \ldots, m\} \rightarrow\{0,1, \ldots, m\}$ and $\nu:\{1, \ldots, l\} \rightarrow\{1, \ldots, n\}$ which represent the positive and negative occurrences of each Boolean variable respectively:

$$
\begin{aligned}
\mu_{\mathcal{F}}(i, j) & = \begin{cases}j & \text { if } v_{i} \in \mathcal{C}_{j}, \\
0 & \text { otherwise },\end{cases} \\
\nu_{\mathcal{F}}(i) & =k \text { for } \neg v_{i} \in \mathcal{D}_{k} .
\end{aligned}
$$

We then define an interpolation equation $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ on a higher-order signature $\Sigma=\langle\mathscr{A}, \mathscr{C}, \tau\rangle$, where $\mathscr{A}=\{o\}, \mathscr{C}=\left\{\mathrm{c}_{j}, \mathrm{~d}_{k}, \mathrm{f}, \mathrm{g} \mid 0 \leq j \leq m\right.$ and $1 \leq k \leq n\}$, and $\tau\left(\mathrm{c}_{j}\right)=o, \tau\left(\mathrm{~d}_{k}\right)=o^{m} \rightarrow o, \tau(\mathrm{f})=o \rightarrow o, \tau(\mathrm{~g})=o^{l} \rightarrow o$. Let

$$
\begin{aligned}
& L_{\mathcal{F}} \equiv \mathbf{X} C_{1} \ldots C_{m} D_{1} \ldots D_{n} \\
& \quad \text { where }\left\{\begin{array}{l}
C_{j} \equiv \mathrm{fc}_{j}, \\
D_{k} \equiv \lambda x_{1}^{o} \ldots x_{m}^{o} \cdot \mathrm{~d}_{k}\left(\mathrm{f} x_{1}\right) \ldots\left(\mathrm{f} x_{m}\right), \\
\tau(\mathbf{X})=o^{m} \rightarrow\left(o^{m} \rightarrow o\right)^{n} \rightarrow o \text { and } \operatorname{ord}(\mathbf{X})=3,
\end{array}\right. \\
& R_{\mathcal{F}} \equiv \mathrm{g} V_{1} \ldots V_{l} \\
& \text { where } V_{i} \equiv \mathrm{~d}_{\nu(i)}\left(\mathrm{fc}_{\mu(i, 1)}\right) \ldots\left(\mathrm{fc}_{\mu(i, m)}\right) .
\end{aligned}
$$

The intuition behind the reduction is the following. Each $C_{j}$ in $L_{\mathcal{F}}$ represents the positive clause $\mathcal{C}_{j}$ and each $D_{k}$ in $L_{\mathcal{F}}$ represents the negative clause $\mathcal{D}_{k}$. Each $V_{i}$ in $R_{\mathcal{F}}$ represents the occurrences of the Boolean variable $v_{i} \in \mathcal{V}$ in the clauses of $\mathcal{F} . V_{i}$ contains $\mathrm{c}_{j}$ for $j \neq 0\left(\right.$ respectively $\left.\mathrm{d}_{k}\right)$ iff $v_{i}$ appears in $\mathcal{C}_{j}\left(\right.$ resp. $\neg v_{i}$ appears in $\left.\mathcal{D}_{k}\right)$. If $\mathcal{F}$ is satisfied by a valuation $\psi$, then for each $\mathcal{C}_{j}\left(\right.$ resp. $\left.\mathcal{D}_{k}\right)$, there is $v_{i} \in \mathcal{C}_{j}\left(\right.$ resp. $\left.\neg v_{i} \in \mathcal{C}_{k}\right)$ such that $\psi\left(v_{i}\right)=1\left(\right.$ resp. $\left.\psi\left(\neg v_{i}\right)=1\right)$. In this case, we can construct a solution $\left[\lambda y_{1} \ldots y_{m} z_{1} \ldots z_{n} . \mathrm{g} U_{1} \ldots U_{l} / \mathbf{X}\right]$ which puts the argument $C_{j}\left(\right.$ resp. $\left.D_{k}\right)$ of $\mathbf{X}$ into $U_{i}$ via $y_{j}\left(\right.$ resp. $z_{k}$ ) and makes $U_{i}$ equivalent to $V_{i}$ by $\beta$-reduction (see

Example 6.11). Conversely if $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ has a solution $[S / \mathbf{X}]$, then $S$ must be of the form $\lambda y_{1} \ldots y_{m} z_{1} \ldots z_{n}$. $\mathbf{g} U_{1} \ldots U_{l}$ and must put each argument $C_{j}$ (resp. $D_{k}$ ) of $\mathbf{X}$ into $U_{i}$ for some $i$ via $y_{j}\left(r e s p . z_{k}\right)$ by the linearity. Then, one can find a valuation $\psi$ such that if $S$ puts the argument $C_{j}$ (resp. $D_{k}$ ) into $U_{i}$, then $\psi$ satisfies $\mathcal{C}_{j}$ via $v_{i}$ (resp. $\mathcal{D}_{k}$ via $\neg v_{i}$ ). The presence of the constant f is the essential difference between Salvati and de Groote's reduction [46] and ours. Due to the number of occurrences of f in $V_{i}$ for each $i, C_{j}$ and $D_{k}$ cannot simultaneously be put into the same $U_{i}$ for any $j$ and $k$. This corresponds to the fact that any valuation $\psi$ on $\mathcal{V}$ cannot simultaneously satisfy $\mathcal{C}_{j}$ via $v_{i}$ and $\mathcal{D}_{k}$ via $\neg v_{i}$ (see Example 6.12).

Example 6.11. Instance $\mathcal{F}: \mathcal{C}_{1}=\left\{v_{1}\right\}, \mathcal{C}_{2}=\left\{v_{1}, v_{2}\right\}, \mathcal{D}_{1}=\left\{\neg v_{1}, \neg v_{2}\right\}$

$$
\begin{array}{ll}
\text { Reduction }\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle: & L_{\mathcal{F}} \equiv \mathbf{X}\left(\mathrm{fc}_{1}\right)\left(\mathrm{fc}_{2}\right)\left(\lambda x_{1} x_{2} \cdot \mathrm{~d}_{1}\left(\mathrm{f} x_{1}\right)\left(\mathrm{f} x_{2}\right)\right) \\
& R_{\mathcal{F}} \equiv \mathrm{g}\left(\mathrm{~d}_{1}\left(\mathrm{fc}_{1}\right)\left(\mathrm{fc}_{2}\right)\right)\left(\mathrm{d}_{1}\left(\mathrm{fc}_{0}\right)\left(\mathrm{fc}_{2}\right)\right)
\end{array}
$$

Let $\psi$ be defined as $\psi\left(v_{1}\right)=1, \psi\left(v_{2}\right)=0$. Corresponding to the fact that $\psi$ satisfies $\mathcal{C}_{1}$ via $v_{1}, \mathcal{C}_{2}$ via $v_{1}$, and $\mathcal{D}_{1}$ via $\neg v_{2}$, we give a solution [ $\left.\lambda y_{1} y_{2} z_{1} \cdot \mathrm{~g} U_{1} U_{2} / \mathbf{X}\right]$ which puts the argument $C_{1}$ of $\mathbf{X}$ into $U_{1}, C_{2}$ into $U_{1}$, and $D_{1}$ into $U_{2}$. That is, we give a solution $[S / \mathbf{X}]$ where

$$
S \equiv \lambda y_{1} y_{2} z_{1} \cdot \mathrm{~g}\left(\mathrm{~d}_{1} y_{1} y_{2}\right)\left(z_{1} \mathrm{c}_{0} \mathrm{c}_{2}\right)
$$

Indeed, we obtain

$$
L_{\mathcal{F}}[S / \mathbf{X}] \equiv S C_{1} C_{2} D_{1} \rightarrow_{\beta} \mathrm{g}\left(\mathrm{~d}_{1} C_{1} C_{2}\right)\left(D_{1} \mathrm{c}_{0} \mathrm{c}_{2}\right) \rightarrow_{\beta} R_{\mathcal{F}} .
$$

Example 6.12. Instance $\mathcal{F}: \mathcal{C}_{1}=\left\{v_{1}\right\}, \mathcal{C}_{2}=\left\{v_{2}\right\}, \mathcal{D}_{1}=\left\{\neg v_{1}, \neg v_{2}\right\}$

$$
\begin{array}{ll}
\text { Reduction }\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle: & L_{\mathcal{F}} \equiv \mathbf{X}\left(\mathrm{fc}_{1}\right)\left(\mathrm{fc}_{2}\right)\left(\lambda x_{1} x_{2} \cdot \mathrm{~d}_{1}\left(\mathrm{f} x_{1}\right)\left(\mathrm{f} x_{2}\right)\right) \\
& R_{\mathcal{F}} \equiv \mathrm{g}\left(\mathrm{~d}_{1}\left(\mathrm{fc}_{1}\right)\left(\mathrm{fc}_{0}\right)\right)\left(\mathrm{d}_{1}\left(\mathrm{fc}_{0}\right)\left(\mathrm{fc}_{2}\right)\right)
\end{array}
$$

$\mathcal{F}$ is not satisfiable and $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ has no solution. e.g.,

$$
\begin{aligned}
& L_{\mathcal{F}}\left[\lambda y_{1} y_{2} z_{1} \cdot \mathrm{~g}\left(\mathrm{~d}_{1} y_{1}\left(\mathrm{fc}_{0}\right)\right)\left(z_{1} \mathrm{c}_{0} y_{2}\right) / \mathrm{X}\right] \\
& \rightarrow{ }_{\beta} \mathrm{g}\left(\mathrm{~d}_{1}\left(\mathrm{fc}_{1}\right)\left(\mathrm{fc}_{0}\right)\right)\left(\mathrm{d}_{1}\left(\mathrm{fc}_{0}\right)\left(\mathrm{f}\left(\mathrm{fc}_{2}\right)\right)\right) \\
& \boldsymbol{F}_{\beta} R_{\mathcal{F}} .
\end{aligned}
$$

Lemma 6.13. $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ admits a solution whenever $\mathcal{F}$ is satisfiable.
Proof. Suppose that $\mathcal{F}$ is satisfied by a valuation $\psi$. Then, for each clause of $\mathcal{F}$, one can choose a literal via which $\psi$ satisfies the clause. Let a function
$\phi$ from $\mathcal{F}$ to $\mathcal{V} \cup \neg \mathcal{V}$ be such a choice. That is, $\phi\left(\mathcal{C}_{j}\right) \in \mathcal{C}_{j}, \psi\left(\phi\left(\mathcal{C}_{j}\right)\right)=1$, $\phi\left(\mathcal{D}_{k}\right) \in \mathcal{D}_{k}$, and $\psi\left(\phi\left(\mathcal{D}_{k}\right)\right)=1$. Define $S$ by

$$
\begin{aligned}
& S \equiv \lambda y_{1} \ldots y_{m} z_{1} \ldots z_{n} \cdot \mathrm{~g} U_{1} \ldots U_{l} \\
& U_{i} \equiv \begin{cases}z_{\nu(i)} \mathrm{c}_{\mu(i, 1)} \ldots \mathrm{c}_{\mu(i, m)} & \text { if } \phi\left(\mathcal{D}_{\nu(i)}\right)=\neg v_{i}, \\
\mathrm{~d}_{\nu(i)} U_{i, 1} \ldots U_{i, m} & \text { otherwise },\end{cases} \\
& U_{i, j} \equiv \begin{cases}y_{j} & \text { if } \phi\left(\mathcal{C}_{j}\right)=v_{i}, \\
\mathrm{fc}_{\mu(i, j)} & \text { otherwise. }\end{cases}
\end{aligned}
$$

First we confirm the linearity of $S$. Each $z_{k}$ indeed occurs exactly once in $S$, since $z_{k}$ appears in $U_{i}$ iff $\phi\left(\mathcal{D}_{k}\right)=\neg v_{i}$. For $y_{j}$, let $i$ be such that $\phi\left(\mathcal{C}_{j}\right)=v_{i}$. Then, $\psi\left(v_{i}\right)=1$ and thus $\phi\left(\mathcal{D}_{\nu(i)}\right) \neq \neg v_{i}$ because $\psi\left(\phi\left(\mathcal{D}_{\nu(i)}\right)\right)=1$. Hence, $U_{i} \equiv \mathrm{~d}_{\nu(i)} U_{i, 1} \ldots U_{i, m}$ and the only occurrence of $y_{j}$ in $S$ is in $U_{i, j}$. Therefore, $S$ is a linear $\lambda$-term.

In order to see that $L_{\mathcal{F}}[S / \mathbf{X}] \rightarrow_{\beta} R_{\mathcal{F}}$, it is enough to check that $U_{i} \sigma \rightarrow_{\beta} V_{i}$ for the substitution $\sigma=\left[C_{1} / y_{1}, \ldots, C_{m} / y_{m}, D_{1} / z_{1}, \ldots, D_{n} / z_{n}\right]$. If $\phi\left(\mathcal{D}_{\nu(i)}\right)=$ $\neg v_{i}$, then

$$
\begin{aligned}
U_{i} \sigma & \equiv D_{\nu(i)} \mathrm{c}_{\mu(i, 1)} \ldots \mathrm{c}_{\mu(i, m)} \\
& \equiv\left(\lambda x_{1} \ldots x_{m} \cdot \mathrm{~d}_{\nu(i)}\left(\mathrm{f}_{1}\right) \ldots\left(\mathrm{f} x_{m}\right)\right) \mathrm{c}_{\mu(i, 1)} \ldots \mathrm{c}_{\mu(i, m)} \\
& \oiint_{\beta} \mathrm{d}_{\nu(i)}\left(\mathrm{fc}_{\mu(i, 1)}\right) \ldots\left(\mathrm{f}_{\mu(i, m)}\right) \\
& \equiv V_{i} .
\end{aligned}
$$

If $\phi\left(\mathcal{D}_{\nu(i)}\right) \neq \neg v_{i}$, then it is easy to see that $U_{i, j} \sigma \equiv \mathrm{fc}_{\mu(i, j)}$. If $\phi\left(\mathcal{C}_{j}\right) \neq v_{i}$, $U_{i, j} \sigma \equiv \mathrm{fc}_{\mu(i, j)} \sigma \equiv \mathrm{fc}_{\mu(i, j)}$. Otherwise, $\phi\left(\mathcal{C}_{j}\right)=v_{i}$ implies $\mu(i, j)=j$ and $U_{i, j} \sigma \equiv y_{j} \sigma \equiv C_{j} \equiv \mathrm{fc}_{j} \equiv \mathrm{fc}_{\mu(i, j)}$. Thus,

$$
U_{i} \sigma \equiv \mathrm{~d}_{\nu(i)} U_{i, 1} \ldots U_{i, m} \sigma \rightarrow_{\beta} \mathrm{d}_{\nu(i)}\left(\mathrm{fc}_{\mu(i, 1)}\right) \ldots\left(\mathrm{fc}_{\mu(i, m)}\right) \equiv V_{i} .
$$

Lemma 6.14. $\mathcal{F}$ is satisfiable whenever $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ admits a solution.
Proof. Suppose that $[S / \mathbf{X}]$ is a solution for $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$. We can assume that $S$ is in long normal form and $S \equiv \lambda y_{1} \ldots y_{m} z_{1} \ldots z_{n}$. $S^{\prime}$. Let $\sigma$ denote the substitution $[\vec{C} / \vec{y}, \vec{D} / \vec{z}]$. Since $S^{\prime} \sigma \rightarrow_{\beta} R_{\mathcal{F}}, S^{\prime}$ must be equal to $\mathrm{g} U_{1} \ldots U_{l}$ for some $\lambda$-terms $U_{i}$ such that $U_{i} \sigma \rightarrow_{\beta} V_{i}$. It is obvious that the head of each $U_{i}$ is either $z_{\nu(i)}$ or $\mathrm{d}_{\nu(i)}$. We show that $\mathcal{F}$ is satisfied by the valuation $\psi$ defined as follows:

$$
\psi\left(v_{i}\right)= \begin{cases}0 & \text { if the head of } U_{i} \text { is } z_{\nu(i)} \\ 1 & \text { otherwise }\end{cases}
$$

We show that each positive clause $\mathcal{C}_{j} \in \mathcal{F}$ is satisfied by $\psi$. Suppose that $y_{j}$ appears in $U_{i}$. Then, the head of $U_{i}$ cannot be $z_{k}$ for any $k$, because if both $y_{j}$ and $z_{k}$ are in $U_{i}, U_{i} \sigma$ contains at least $(m+1)$ occurrences of f , so $U_{i} \sigma$ never $\beta$-reduces to $V_{i}$, which contains exactly $m$ occurrences of f . Thus, the head of $U_{i}$ is $\mathrm{d}_{\nu(i)}$ and $\psi\left(v_{i}\right)=1$. Since $y_{j} \sigma \equiv \mathrm{fc}_{j}, U_{i} \sigma$ and $V_{i}$ must contain $\mathrm{c}_{j}$ and this implies $\mu(i, j)=j$, i.e., $v_{i} \in \mathcal{C}_{j}$. $\psi$ satisfies $\mathcal{C}_{j}$ via $v_{i}$.

We show that each negative clause $\mathcal{D}_{k} \in \mathcal{F}$ is satisfied by $\psi$. Suppose that $z_{k}$ appears in $U_{i}$. Since $z_{k} \sigma$ contains $\mathrm{d}_{k}, k=\nu(i)$ and the head of $U_{i}$ is $z_{k}$. Therefore, $\psi\left(v_{i}\right)=0$ and $\neg v_{i} \in \mathcal{D}_{k}$. $\psi$ satisfies $\mathcal{D}_{k}$ via $\neg v_{i}$.

Proposition 6.15. Third-order interpolation modulo $\beta(\beta \eta)$ in the linear lambda calculus is NP-complete.

Proof. By Theorem 6.7 and Lemmas 6.13 and 6.14.

### 6.5 Elimination of Constants

In this section, we show that the interpolation problem in the linear lambda calculus is NP-hard even if there are no constants, by eliminating constants in $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ in Definition 6.10.

For general unification problem, it is easy to construct a constant-free $\left\langle P^{*}, Q^{*}\right\rangle$ from a unification equation $\langle P, Q\rangle$, such that $\langle P, Q\rangle$ has a solution iff $\left\langle P^{*}, Q^{*}\right\rangle$ has a solution. We obtain $P^{*}$ and $Q^{*}$ by successive transformations performed on $P$ and $Q$ : first we replace the constants a by fresh variables $\vec{a}$, then we replace each unknown $\mathbf{X}_{i}$ with $\mathbf{Y}_{i} \vec{a}$ and finally we abstract the free variables. If $\left[S_{i} / \mathbf{X}_{i}\right]$ is a solution for $\langle P, Q\rangle$, then $\left[\lambda \vec{a} . S_{i}^{*} / \mathbf{Y}_{i}\right]$ is a solution for $\left\langle P^{*}, Q^{*}\right\rangle$, where $S^{*}$ is obtained by replacing each constant a by the variable $a$. If $\left[S_{i}^{*} / \mathbf{Y}_{i}\right]$ is a solution for $\left\langle P^{*}, Q^{*}\right\rangle$, then $\left[S_{i}^{*} \overrightarrow{\mathrm{a}} / \mathbf{X}_{i}\right]$ is a solution for $\langle P, Q\rangle$.

However, such a transformation does not work for the unification problem in the linear lambda calculus, because free variables can occur at most once in a linear $\lambda$-term, while constants can occur any number of times. To construct a linear interpolation equation $\left\langle L_{\mathcal{F}}^{*} \equiv \mathrm{Y} C_{1}^{*} \ldots C_{m}^{*} D_{1}^{*} \ldots D_{n}^{*}, R_{\mathcal{F}}^{*}\right\rangle$ by eliminating constants from $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ defined in Definition 6.10 , we adopt the following strategy:

- Let $T$ be among $C_{1}, \ldots, C_{m}, D_{1}, \ldots, D_{n}, R_{\mathcal{F}}$.
- Let $T^{\prime}$ be the result of replacing occurrences of each constant in $T$ with suitable free variables or $\lambda$-terms constructed from free variables.
- Let $T^{*} \equiv \lambda \vec{x} \cdot T^{\prime}$ for a sequence $\vec{x}$ of the elements of $\operatorname{Fv}\left(T^{\prime}\right)$.

The main issue is what variable or $\lambda$-term each occurrence of a constant should be replaced with. The formal definition and an example are given in Definition 6.16 and Example 6.17. We give the basic strategy of our transformation here. First, for each constant a, we provide an atomic type $p_{\mathrm{a}}$. Second we replace the $i$-th occurrence of a in $T$ with the application $x_{\mathrm{a}, i} y_{\mathrm{a}, i}$ of variables $x_{\mathrm{a}, i}$ of type $p_{\mathrm{a}} \rightarrow \tau(\mathrm{a})$ and $y_{\mathrm{a}, i}$ of type $p_{\mathrm{a}}$. Finally, we close $T$ by $\lambda \vec{x} \vec{y}$. This way, we can avoid identifying $\lambda$-terms which are surrogates for distinct constants of the same type, while preserving the well-typedness and the linearity of the interpolation equation. If $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ has a solution $[S / \mathbf{X}]$, then $\left\langle L_{\mathcal{F}}^{*}, R_{\mathcal{F}}^{*}\right\rangle$ also has a solution $\left[S^{*} / \mathbf{Y}\right]$ which does not essentially differ from $[S / \mathbf{X}]$. Each occurrence of a constant a in $S$ is replaced with the application $x_{\mathrm{a}, i} y_{\mathrm{a}, i}$ of two bound variables of $R_{\mathcal{F}}^{*}$ for the appropriate $i$. Moreover, $\left[S^{*} / \mathbf{Y}\right]$ lets the arguments $T^{*}$ of the unknown $\mathbf{Y}$ be applied to bound variables of $R_{\mathcal{F}}^{*}$ which constitute the surrogates for constants which appear in $T$.

A major problem is that one may construct a solution which causes a substitution of a complex $\lambda$-term for the bound variable $x_{\mathrm{a}, i}$ of an argument $T^{*}$ of the unknown $\mathbf{Y}$, since a $\lambda$-term which has the type $p_{\mathrm{a}} \rightarrow \tau(\mathrm{a})$ is not necessarily a bound variable $x_{\mathrm{a}, j}$ in $R_{\mathcal{F}}^{*}$ for some $j$. For instance, consider the interpolation equation $\langle\mathbf{X} / \mathrm{ac} / / \mathrm{b} /$, $/ \mathrm{abc} /\rangle$ where $\tau(\mathrm{a})=\tau(\mathrm{b})=\tau(\mathrm{c})=$ str. Clearly it has no solution. By applying the above conversion to $\langle\mathbf{X} / \mathrm{ac} / / \mathrm{b} /$, /abc/ $\rangle$, we obtain the following equation:

$$
\begin{aligned}
&\left\langle\mathbf{Y}\left(\lambda x_{\mathrm{a}}^{p_{\mathrm{a}} \rightarrow s t r} y_{\mathrm{a}}^{p_{\mathrm{a}}} x_{\mathrm{c}}^{p_{\mathrm{c}} \rightarrow s t r} y_{\mathrm{c}}^{p_{\mathrm{c}}} z^{o} \cdot x_{\mathrm{a}} y_{\mathrm{a}}\left(x_{\mathrm{c}} y_{\mathrm{c}} z\right)\right)\left(\lambda x_{\mathrm{b}}^{p_{\mathrm{b}} \rightarrow s t r} y_{\mathrm{b}}^{p_{\mathrm{b}}} z^{o} \cdot x_{\mathrm{b}} y_{\mathrm{b}} z\right),\right. \\
&\left.\lambda x_{\mathrm{a}}^{p_{\mathrm{a}} \rightarrow \operatorname{str}} y_{\mathrm{a}}^{p_{\mathrm{a}}} x_{\mathrm{b}}^{p_{\mathrm{b}} \rightarrow \operatorname{str}} y_{\mathrm{b}}^{p_{\mathrm{b}}} x_{\mathrm{c}}^{p_{\mathrm{c}} \rightarrow s t r} y_{\mathrm{c}}^{p_{\mathrm{c}}} z^{o} \cdot x_{\mathrm{a}} y_{\mathrm{a}}\left(x_{\mathrm{b}} y_{\mathrm{b}}\left(x_{\mathrm{c}} y_{\mathrm{c}} z\right)\right)\right\rangle,
\end{aligned}
$$

but, this has a solution

$$
\begin{aligned}
& {\left[\lambda w_{1}^{\left(p_{\mathrm{a}} \rightarrow s t r\right) \rightarrow p_{\mathrm{a}} \rightarrow\left(p_{\mathrm{c}} \rightarrow s t r\right) \rightarrow p_{\mathrm{c}} \rightarrow s t r} w_{2}^{\left(p_{\mathrm{b}} \rightarrow s t r\right) \rightarrow p_{\mathrm{b}} \rightarrow s t r} .\right.} \\
& \left.\lambda x_{\mathrm{a}}^{p_{\mathrm{a}} \rightarrow s t r} y_{\mathrm{a}}^{p_{\mathrm{a}}} x_{\mathrm{b}}^{p_{\mathrm{b}} \rightarrow s t r} y_{\mathrm{b}}^{p_{\mathrm{b}}} x_{\mathrm{c}}^{p_{\mathrm{c}} \rightarrow s t r} y_{\mathrm{c}}^{p_{\mathrm{c}}} \cdot w_{1} x_{\mathrm{a}} y_{\mathrm{a}}\left(\lambda v_{\mathrm{c}}^{p_{\mathrm{c}}} z^{o} \cdot w_{2} x_{\mathrm{b}} y_{\mathrm{b}}\left(x_{\mathrm{c}} v_{\mathrm{c}} z\right)\right) y_{\mathrm{c}} / \mathbf{Y}\right] .
\end{aligned}
$$

This is the reason why we do not employ the reduction from the string rearrangement problem as in the proof of Proposition 6.9 for constant-free cases. Fortunately, we can tightly restrict $\lambda$-terms that can be substituted for bound variables in the arguments $T^{*}$ of $\mathbf{Y}$, when we employ the reduction in Definition 6.10 as a basis for the new reduction. Recall that in $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ in Definition 6.10, $\mathrm{c}_{j}$ is always an argument of $\mathrm{f}, \mathrm{fc}_{j}$ is of $\mathrm{d}_{k}$ for some $k$, and a $\lambda$-term whose head is $\mathrm{d}_{k}$ is of g . Hence we can let $\lambda$-terms in $\left\langle L_{\mathcal{F}}^{*}, R_{\mathcal{F}}^{*}\right\rangle$ which are surrogates for $\mathrm{c}_{j}$ have type $q$, surrogates for f have type $q \rightarrow r$, surrogates for $\mathrm{d}_{k}$ have type $r^{m} \rightarrow t$, and surrogates for g have type $t^{l} \rightarrow u$.

Now, we give a formal definition of the linear interpolation equation $\left\langle L_{\mathcal{F}}^{*}, R_{\mathcal{F}}^{*}\right\rangle$ for a given polarized $1 \mathrm{~N}-\mathrm{CNF} \mathcal{F}$.

Definition 6.16. Let $\mathcal{F}=\left\{\mathcal{C}_{1}, \ldots, \mathcal{C}_{m}, \mathcal{D}_{1}, \ldots, \mathcal{D}_{n}\right\}$ be a polarized 1 N-CNF on $\mathcal{V}=\left\{v_{1}, \ldots, v_{l}\right\}$ where $\mathcal{C}_{1}, \ldots, \mathcal{C}_{m}$ are positive clauses and $\mathcal{D}_{1}, \ldots, \mathcal{D}_{n}$ are negative clauses. Two functions $\mu$ and $\nu$ are defined as in Definition 6.10. Let $\mathscr{A}=\left\{p_{j}, q, r, s_{k}, t, u \mid 0 \leq j \leq m, 1 \leq k \leq n\right\}$. Let $\vec{w}_{\mathcal{F}}$ be a sequence of the variables in the set

$$
\left\{c_{i, j}^{p_{\mu(i, j)}}, \bar{c}_{i, j}^{p_{\mu(i, j)} \rightarrow q}, f_{i, j}^{q \rightarrow r}, d_{i, \nu(i)}^{s_{\nu}(i)}, \bar{d}_{i, \nu(i)}^{s_{\nu(i)} \rightarrow r^{m} \rightarrow t}, g^{t^{t^{\prime} \rightarrow u}} \mid 1 \leq i \leq l \text { and } 1 \leq j \leq m\right\} .
$$

$\left\langle L_{\mathcal{F}}^{*}, R_{\mathcal{F}}^{*}\right\rangle$ is defined by

$$
\begin{aligned}
& L_{\mathcal{F}}^{*} \equiv \mathbf{Y} C_{1}^{*} \ldots C_{m}^{*} D_{1}^{*} \ldots D_{n}^{*} \text { where } \\
& \qquad\left\{\begin{array}{l}
C_{j}^{*} \equiv \lambda f^{q \rightarrow r} \bar{c}_{j}^{p_{j} \rightarrow q} c_{j}^{p_{j}} . f\left(\bar{c}_{j} c_{j}\right) \\
D_{k}^{*} \equiv \lambda \bar{d}_{k}^{s_{k} \rightarrow r^{m} \rightarrow t} d_{k}^{s,} f_{1}^{q \rightarrow r} \ldots f_{m}^{q \rightarrow r} x_{1}^{q} \ldots x_{m}^{q} . \bar{d}_{k} d_{k}\left(f_{1} x_{1}\right) \ldots\left(f_{m} x_{m}\right) \\
\tau(\mathbf{Y})=\tau\left(C_{1}^{*}\right) \rightarrow \cdots \rightarrow \tau\left(C_{m}^{*}\right) \rightarrow \tau\left(D_{1}^{*}\right) \rightarrow \cdots \rightarrow \tau\left(D_{n}^{*}\right) \rightarrow \tau\left(R_{\mathcal{F}}^{*}\right)
\end{array}\right. \\
& R_{\mathcal{F}}^{*} \equiv \lambda \vec{w}_{\mathcal{F} .} \cdot g^{t^{t} \rightarrow u} V_{1}^{*} \ldots V_{l}^{*} \quad \text { where } \\
& \quad V_{i}^{*} \equiv \bar{d}_{i, \nu(i)}^{s_{\nu(i)} \rightarrow r^{m} \rightarrow t} d_{i, \nu(i)}^{s_{\nu(i)}}\left(f_{i, 1}^{q \rightarrow r}\left(\bar{c}_{i, 1}^{p_{\mu(i, 1)} \rightarrow q} c_{i, 1}^{p_{\mu(i, 1)}}\right)\right) \ldots\left(f_{i, m}^{q \rightarrow r}\left(\bar{c}_{i, m}^{p_{\mu(i, m)} \rightarrow q} c_{i, m}^{p_{\mu(i, m)}}\right)\right) .
\end{aligned}
$$

Indeed the above $\lambda$-terms are well-typed.

$$
\begin{aligned}
\tau\left(C_{j}^{*}\right) & =(q \rightarrow r) \rightarrow\left(p_{j} \rightarrow q\right) \rightarrow p_{j} \rightarrow r \\
\tau\left(D_{k}^{*}\right) & =\left(s_{k} \rightarrow r^{m} \rightarrow t\right) \rightarrow s_{k} \rightarrow(q \rightarrow r)^{m} \rightarrow q^{m} \rightarrow t \\
\operatorname{ord}(\mathbf{Y}) & =4
\end{aligned}
$$

Throughout this section, we use the following type assignment to variables for $1 \leq i \leq l, 1 \leq j \leq m$ and $1 \leq k \leq n$, which is consistent with the above definition:

$$
\begin{gathered}
\tau\left(c_{j}\right)=p_{j}, \quad \tau\left(c_{i, j}\right)=p_{\mu(i, j)}, \quad \tau\left(\bar{c}_{j}\right)=p_{j} \rightarrow q, \quad \tau\left(\bar{c}_{i, j}\right)=p_{\mu(i, j)} \rightarrow q, \\
\tau\left(d_{k}\right)=\tau\left(d_{i, k}\right)=s_{k}, \quad \tau\left(\bar{d}_{k}\right)=\tau\left(\bar{d}_{i, k}\right)=s_{k} \rightarrow r^{m} \rightarrow t \\
\tau(f)=\tau\left(f_{j}\right)=\tau\left(f_{i, j}\right)=q \rightarrow r, \quad \tau(g)=t^{l} \rightarrow u \\
\tau\left(x_{j}\right)=q, \quad \tau\left(y_{j}\right)=\tau\left(C_{j}^{*}\right), \quad \tau\left(z_{k}\right)=\tau\left(D_{k}^{*}\right) .
\end{gathered}
$$

Example 6.17. Instance $\mathcal{F}: \mathcal{C}_{1}=\left\{v_{1}\right\}, \mathcal{C}_{2}=\left\{v_{2}\right\}, \mathcal{D}_{1}=\left\{\neg v_{1}, \neg v_{2}\right\}$
Reduction with Constants $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ :

$$
\begin{aligned}
L_{\mathcal{F}} & \equiv \mathbf{X}\left(\mathrm{fc}_{1}\right)\left(\mathrm{fc}_{2}\right)\left(\lambda x_{1} x_{2} \cdot \mathrm{~d}_{1}\left(\mathrm{f} x_{1}\right)\left(\mathrm{f} x_{2}\right)\right) \\
R_{\mathcal{F}} & \equiv \mathrm{g}\left(\mathrm{~d}_{1}\left(\mathrm{fc}_{1}\right)\left(\mathrm{fc}_{0}\right)\right)\left(\mathrm{d}_{1}\left(\mathrm{fc}_{0}\right)\left(\mathrm{fc}_{2}\right)\right)
\end{aligned}
$$

Reduction without Constants $\left\langle L_{\mathcal{F}}^{*}, R_{\mathcal{F}}^{*}\right\rangle$ :

$$
\begin{aligned}
& L_{\mathcal{F}}^{*} \equiv \mathbf{Y}\left(\lambda f \bar{c}_{1} c_{1} \cdot f\left(\bar{c}_{1} c_{1}\right)\right)\left(\lambda f \bar{c}_{2} c_{2} \cdot f\left(\bar{c}_{2} c_{2}\right)\right) \\
& \quad\left(\lambda \bar{d}_{1} d_{1} f_{1} f_{2} x_{1} x_{2} \cdot \bar{d}_{1} d_{1}\left(f_{1} x_{1}\right)\left(f_{2} x_{2}\right)\right) \\
& R_{\mathcal{F}}^{*} \equiv \lambda \vec{w}_{\mathcal{F}} \cdot g\left(\bar{d}_{1,1} d_{1,1}\left(f_{1,1}\left(\bar{c}_{1,1} c_{1,1}\right)\right)\left(f_{1,2}\left(\bar{c}_{1,2} c_{1,2}\right)\right)\right) \\
& \quad\left(\bar{d}_{2,1} d_{2,1}\left(f_{2,1}\left(\bar{c}_{2,1} c_{2,1}\right)\right)\left(f_{2,2}\left(\bar{c}_{2,2} c_{2,2}\right)\right)\right)
\end{aligned}
$$

where the type of each variable is as Definition 6.16. $\mathcal{F}$ is not satisfiable and $\left\langle L_{\mathcal{F}}, R_{\mathcal{F}}\right\rangle$ and $\left\langle L_{\mathcal{F}}^{*}, R_{\mathcal{F}}^{*}\right\rangle$ have no solution. Note that

$$
\begin{aligned}
{\left[\lambda y_{1} y_{2} z_{1} \vec{w}_{\mathcal{F} .} g\right.} & \left(z_{1} \bar{d}_{1,1} d_{1,1} f_{1,1} f_{1,2}\left(\bar{c}_{1,1} c_{1,1}\right)\left(\bar{c}_{1,2} c_{1,2}\right)\right) \\
& \left.\left(\bar{d}_{2,1} d_{2,1}\left(y_{1} f_{2,1} \bar{c}_{2,1} c_{2,1}\right)\left(y_{2} f_{2,2} \bar{c}_{2,2} c_{2,2}\right)\right) / \mathbf{Y}\right]
\end{aligned}
$$

is not a solution for $\left\langle L_{\mathcal{F}}^{*}, R_{\mathcal{F}}^{*}\right\rangle$. $\left(y_{1} f_{2,1} \bar{c}_{2,1} c_{2,1}\right)$ is not a $\lambda$-term of the simply typed lambda calculus, for $\tau\left(y_{1}\right)=(q \rightarrow r) \rightarrow\left(p_{1} \rightarrow q\right) \rightarrow p_{1} \rightarrow r$ but $\tau\left(\bar{c}_{2,1}\right)=p_{\mu(2,1)} \rightarrow q=p_{0} \rightarrow q$ and $\tau\left(c_{2,1}\right)=p_{\mu(2,1)}=p_{0}$.

Lemma 6.18. $\left\langle L_{\mathcal{F}}^{*}, R_{\mathcal{F}}^{*}\right\rangle$ admits a solution whenever $\mathcal{F}$ is satisfiable.
Proof. Suppose that $\psi$ is a valuation which satisfies $\mathcal{F}$. As in the proof of Lemma 6.13 , we can find $\phi: \mathcal{F} \rightarrow(\mathcal{V} \cup \neg \mathcal{V})$ which indicates a literal via which each clause is satisfied. We show that a solution is given by $[S / \mathbf{Y}]$ where

$$
S \equiv \lambda y_{1} \ldots y_{m} z_{1} \ldots z_{n} \vec{w}_{\mathcal{F}} . g U_{1} \ldots U_{l} \quad \text { where }
$$

$$
\begin{aligned}
U_{i} \equiv \begin{cases}z_{\nu(i)} \bar{d}_{i, \nu(i)} d_{i, \nu(i)} f_{i, 1} \ldots f_{i, m}\left(\bar{c}_{i, 1} c_{i, 1}\right) \ldots\left(\bar{c}_{i, m} c_{i, m}\right) & \text { if } \phi\left(\mathcal{D}_{\nu(i)}\right)=\neg v_{i} \\
\bar{d}_{i, \nu(i)} d_{i, \nu(i)} U_{i, 1} \ldots U_{i, m} & \text { otherwise },\end{cases} \\
U_{i, j} \equiv \begin{cases}y_{j} f_{i, j} \bar{c}_{i, j} c_{i, j} & \text { if } \phi\left(\mathcal{C}_{j}\right)=v_{i}, \\
f_{i, j}\left(\bar{c}_{i, j} c_{i, j}\right) & \text { otherwise }\end{cases}
\end{aligned}
$$

Indeed $S$ is a well-typed closed linear $\lambda$-term. The linearity of $S$ can be checked as in the proof of Lemma 6.13. We check the well-typedness of $S$. Recall that $\tau\left(z_{\nu(i)}\right)=\left(s_{\nu(i)} \rightarrow r^{m} \rightarrow t\right) \rightarrow s_{\nu(i)} \rightarrow(q \rightarrow r)^{m} \rightarrow q^{m} \rightarrow t$, $\tau\left(\bar{d}_{i, \nu(i)}\right)=s_{\nu(i)} \rightarrow r^{m} \rightarrow t, \tau\left(d_{i, \nu(i)}\right)=s_{\nu(i)}, \tau\left(f_{i, j}\right)=q \rightarrow r$ and $\tau\left(\bar{c}_{i, j} c_{i, j}\right)=$
$q$. Hence, $U_{i}$ is well-typed and has the type $t$ if $\phi\left(\mathcal{D}_{\nu(i)}\right)=\neg v_{i}$. If $\phi\left(\mathcal{D}_{\nu(i)}\right) \neq$ $\neg v_{i}$, it is enough to check that each $U_{i, j}$ has type $r$. If $\phi\left(\mathcal{C}_{j}\right)=v_{i}$ and $U_{i, j} \equiv y_{j} f_{i, j} \bar{c}_{i, j} c_{i, j}$, then $v_{i} \in \mathcal{C}_{j}, \tau\left(\bar{c}_{i, j}\right)=p_{\mu(i, j)} \rightarrow q=p_{j} \rightarrow q$ and $\tau\left(c_{i, j}\right)=$ $p_{\mu(i, j)}=p_{j}$. Recall that $\tau\left(y_{j}\right)=(q \rightarrow r) \rightarrow\left(p_{j} \rightarrow q\right) \rightarrow p_{j} \rightarrow r$ and $\tau\left(f_{i, j}\right)=q \rightarrow r$. Hence, $U_{i, j}$ is well-typed and has the type $r$. If $\phi\left(\mathcal{C}_{j}\right) \neq v_{i}$, it is clear that $U_{i, j}$ is well-typed and has the type $r$. Therefore, $S$ is well-typed.

Second we show that $L_{\mathcal{F}}^{*}[S / \mathbf{Y}] \rightarrow_{\beta} R_{\mathcal{F}}^{*}$. By the definition, $L_{\mathcal{F}}^{*}[S / \mathbf{Y}] \rightarrow_{\beta}$ $\lambda \vec{w}_{\mathcal{F}} . g U_{1} \ldots U_{l} \sigma$ for $\sigma=\left[\vec{C}^{*} / \vec{y}, \vec{D}^{*} / \vec{y}\right]$. It is enough to show that $U_{i} \sigma \rightarrow{ }_{\beta} V_{i}^{*}$. If $\phi\left(\mathcal{D}_{\nu(i)}\right)=\neg v_{i}$, then

$$
\begin{aligned}
U_{i} \sigma & \equiv D_{\nu(i)}^{*} \bar{d}_{i, \nu(i)} d_{i, \nu(i)} f_{i, 1} \ldots f_{i, m}\left(\bar{c}_{i, 1} c_{i, 1}\right) \ldots\left(\bar{c}_{i, m} c_{i, m}\right) \\
& \rightarrow_{\beta} \bar{d}_{i, \nu(i)} d_{i, \nu(i)}\left(f_{i, 1}\left(\bar{c}_{i, 1} c_{i, 1}\right)\right) \ldots\left(f_{i, m}\left(\bar{c}_{i, m} c_{i, m}\right)\right) \\
& \equiv V_{i}^{*} .
\end{aligned}
$$

Otherwise, $U_{i} \equiv \bar{d}_{i, \nu(i)} d_{i, \nu(i)} U_{i, 1} \ldots U_{i, m}$. It is obvious that $U_{i, j} \sigma \rightarrow_{\beta} f_{i, j}\left(\bar{c}_{i, j} c_{i, j}\right)$ for all $j$, since $C_{j}^{*} f_{i, j} \bar{c}_{i, j} c_{i, j} \rightarrow{ }_{\beta} f_{i, j}\left(\bar{c}_{i, j} c_{i, j}\right)$.

Lemma 6.19. Suppose that a linear $\lambda$-term $M$ contains no free variables other than the elements of $\vec{w}_{\mathcal{F}}$ and $y_{1}, \ldots, y_{m}, z_{1}, \ldots, z_{n}$.

If $M$ has an atomic type $\delta$ and contains a subterm $N$ of type $\gamma_{1} \rightarrow \cdots \rightarrow$ $\gamma_{i} \rightarrow \gamma$ with $\gamma \in \mathscr{A}$, then $\gamma \leq \delta$ where $\leq$ is the partial order on $\mathscr{A}$ such that $p_{j} \lesseqgtr q \leq r \leq t \leq u$ for $0 \leq j \leq m$ and $s_{k} \lesseqgtr t \lesseqgtr u$ for $1 \leq k \leq n$. If $\tau(M)=q \rightarrow r$, then $M={ }_{\eta} f_{i, j}$ for some $i$ and $j$.

Lemma 6.20. Suppose that a subterm of a linear $\lambda$-term $M$ has an atomic type $p$. Then, for every $N$ such that $M={ }_{\beta} N, N$ has a subterm of type $p$.

Lemma 6.21. $\mathcal{F}$ is satisfiable whenever $\left\langle L_{\mathcal{F}}^{*}, R_{\mathcal{F}}^{*}\right\rangle$ admits a solution.
Proof. Suppose that $[S / \mathbf{Y}]$ is a solution. We can assume that $S$ is $\beta$-normal and $S \equiv \lambda y_{1} \ldots y_{m} z_{1} \ldots z_{n} \vec{w}_{\mathcal{F}} \cdot S^{\prime}$. Because of the type of the variable $g$ in $R_{\mathcal{F}}^{*}$, the head of $S^{\prime}$ is neither $y_{j}$ nor $z_{k}$. $S^{\prime}$ must be $g U_{1} \ldots U_{l}$ for some $U_{1}, \ldots, U_{l}$ of type $t$. Let $\sigma=\left[\overrightarrow{C^{*}} / \vec{y}, \vec{D}^{*} / \vec{z}\right]$. Since $S C_{1}^{*} \ldots C_{m}^{*} D_{1}^{*} \ldots D_{n}^{*} \rightarrow_{\beta}$ $\lambda \vec{w}_{\mathcal{F}} . g U_{1} \ldots U_{l} \sigma \rightarrow_{\beta} \lambda \vec{w}_{\mathcal{F}} . g V_{1}^{*} \ldots V_{l}^{*}$, we have $U_{i} \sigma \rightarrow_{\beta} V_{i}^{*}$. Note that
${ }^{(*)} U_{i}$ contains no free variables other than the elements of

$$
\begin{aligned}
& \operatorname{Fv}\left(V_{i}^{*}\right) \cup\left\{y_{j}, z_{k} \mid 1 \leq j \leq m, 1 \leq k \leq n\right\} \\
= & \left\{\bar{d}_{i, \nu(i)}, d_{i, \nu(i)}, f_{i, j}, \bar{c}_{i, j}, c_{i, j}, y_{j}, z_{k} \mid 1 \leq j \leq m, 1 \leq k \leq n\right\} .
\end{aligned}
$$

Since $U_{i}$ has type $t$, the head of $U_{i}$ must be $\bar{d}_{i, \nu(i)}$ or $z_{k}$ for some $k$.
First, we show that if $U_{i}$ contains $z_{k}$, then $z_{k}$ is the head of $U_{i}$ and $U_{i}$ contains neither $y_{j}$ nor $z_{k^{\prime}}$ for any $j$ and $k^{\prime} \neq k$. Since $z_{k} \sigma \equiv D_{k}^{*}$ contains a
variable of type $s_{k}, U_{i} \sigma$ and its $\beta$-normal form $V_{i}^{*}$ contain a subterm of type $s_{k}$ by Lemma 6.20. This implies that $k=\nu(i)$ and $V_{i}^{*}$ contains no subterm of type $s_{k^{\prime}}$ for any $k^{\prime} \neq k$ and thus $U_{i} \sigma$ does so. Therefore, $U_{i}$ does not contain $z_{k^{\prime}}$ for any $k^{\prime} \neq k$. If $U_{i} \equiv \bar{d}_{i, \nu(i)} M_{\nu(i)} R_{1} \ldots R_{m}$ for $\tau\left(M_{\nu(i)}\right)=s_{\nu(i)}$ and $\tau\left(R_{h}\right)=r$ for $1 \leq h \leq m$, then $z_{k}$ cannot occur in $U_{i}$ by Lemma 6.19. Therefore, if $U_{i}$ contains $z_{k}$, then $U_{i}$ is of the form

$$
U_{i} \equiv z_{k} N_{k} M_{k} F_{1} \ldots F_{m} Q_{1} \ldots Q_{m}
$$

where $\tau\left(N_{k}\right)=s_{k} \rightarrow r^{m} \rightarrow t, \tau\left(M_{k}\right)=s_{k}, \tau\left(F_{h}\right)=q \rightarrow r$, and $\tau\left(Q_{h}\right)=q$ for $1 \leq h \leq m$. is proved. We check that $y_{j}$ cannot occur in $N_{k}, M_{k}, Q_{h}$ or $F_{h}$ for any $h$. It is obvious that $y_{j}$ is not in $M_{k}, Q_{h}$ or $F_{h}$ by $\left(^{*}\right)$ and Lemma 6.19. Since $N_{k}$ does not contain $z_{k^{\prime}}$ for any $k^{\prime}, N_{k}$ can be written as

$$
N_{k} \equiv \lambda d_{k}^{s_{k}} w_{1}^{r} \ldots w_{m}^{r} \cdot \bar{d}_{i, \nu(i)} d_{k} R_{1} \ldots R_{m}
$$

where $\tau\left(R_{h}\right)=r$ for all $h$ (provided that $\nu(i)=k$ ). It is easy to see that if $w_{h}$ appears in some $R_{h^{\prime}}$, then $R_{h^{\prime}}=w_{h}$. Thus the sequence $R_{1} \ldots R_{m}$ is just a permutation of $w_{1} \ldots w_{m} . y_{j}$ does not appear in $N_{k}$. Therefore, $y_{j}$ cannot occur in $U_{i}$ unless the head of $U_{i}$ is $\bar{d}_{i, \nu(i)}$.

Now, we show that $\mathcal{F}$ is satisfied by the valuation $\psi$ defined as follows:

$$
\psi\left(v_{i}\right)= \begin{cases}0 & \text { if the head of } U_{i} \text { is a variable } z_{k} \text { for some } k \\ 1 & \text { otherwise }\end{cases}
$$

By the linearity of $S$, each variable $y_{1}, \ldots, y_{m}, z_{1}, \ldots, z_{n}$ appears in $U_{i}$ for some $i$. To show that each positive clause $\mathcal{C}_{j}$ is satisfied by $\psi$, suppose that $y_{j}$ occurs in $U_{i}$. The above discussion claims that the head of $U_{i}$ is $\bar{d}_{i, \nu(i)}$ and thus $\psi\left(v_{i}\right)=1$. Since $y_{j} \sigma \equiv C_{j}^{*}$ contains a variable of type $p_{j}, U_{i} \sigma$ and its $\beta$-normal form $V_{i}^{*}$ contain a subterm of type $p_{j}$ by Lemma 6.20. Therefore, $\tau\left(c_{i, j}\right)=p_{\mu(i, j)}=p_{j}$ and $v_{i} \in \mathcal{C}_{j}$. Thus $\psi$ satisfies $\mathcal{C}_{j}$ via $v_{i}$.

To show that each negative clause $\mathcal{D}_{k}$ is satisfied by $\psi$, suppose that $z_{k}$ occurs in $U_{i}$. The above discussion claims that $z_{k}$ is the head of $U_{i}$ and $k=\nu(i)$. Therefore, $\psi$ satisfies $\mathcal{D}_{k}$ via $\neg v_{i}$.

Theorem 6.22. Fourth-order interpolation modulo $\beta$ ( $\beta \eta$ ) in the linear lambda calculus in the absence of constants is NP-complete.

Proof. By Theorem 6.7 and Lemmas 6.18 and 6.21.
The order of the equation in Definition 6.16 is higher than the one in Definition 6.10 by one, though the maximum order of the types of the variables appearing in the former equation is the same as the one of the constants in
the latter. The reason of the increase of the order is due to the presence of $\lambda$-abstraction.

It is easy to see that Theorem 6.22 does not hold for the third-order case unless $\mathrm{P}=\mathrm{NP}$. For an interpolation equation $\left\langle\mathbf{X} L_{1} \ldots L_{m}, R\right\rangle$, if $L_{i}$ is a constant-free closed linear $\lambda$-term whose type is at most second-order, then $L_{i} \equiv \lambda x_{i}^{p_{i}} \cdot x_{i}$ for some atomic type $p_{i} \in \mathscr{A}$. Hence $\left\langle\mathbf{X} L_{1} \ldots L_{m}, R\right\rangle$ has a linear solution modulo $\beta$ (resp. $\beta \eta$ ) iff (resp. the $\eta$-long form of) $R$ has a subterm of type $p_{i}$ for every $i$ by Lemma 6.20.

## Chapter 7

## Conclusions

This thesis has investigated the mathematical properties of some extensions and restrictions of ACGs. In particular, the aspect of ACGs as a generalization of well-established grammar formalisms is investigated.

In Chapter 3, we have shown that allowing deleting operations do not enrich the expressive power of ACGs. The result entails the equivalence between linear CFTGs and non-duplicating CFTGs, and between LCFRSs and MCFGs, as corollaries. It is future work whether our procedure for elimination of vacuous $\lambda$-abstraction can be generalized so that Fisher's result $[8,9]$ is covered.

In Chapter 4, we have proposed a lexicalization method of semilexicalized ACGs which preserves the order of the abstract vocabulary. Our result entails that LCFRSs admit lexicalization. Moreover, our lexicalization method for second-order ACGs has a close connection with Schabes et al.'s preceding research $[47,48]$ that converts finitely ambiguous CFGs into lexicalized TAGs. There remains an open question, whether or not second and thirdorder semilexicalized ACGs can be lexicalized preserving the orders of the lexicons. If this open problem is solved in the affirmative, it would give a solution for the corresponding problems for mildly context-sensitive formalisms. To answer the question, it seems inevitable to analyze the possible forms of object terms assigned to abstract constants by the lexicon.

In Chapter 5, we have presented that several well-known tree transducer formalisms are encodable in two-dimensional ACGs. In particular, we have seen the equivalence between linear macro tree transducers and ACGs belonging to $\mathbf{G}_{\text {tree }}(2,1(\mathrm{sr}), 2)$. Therefore, the ACG formalism has the rich potential as a general framework in which other existing two-dimensional formalisms, as well as one-dimensional formalisms, may be encoded. The hierarchy of second-order two-dimensional ACGs, however, contains many subclasses whose generative capacity is not known yet, while the hierar-
chy of second-order one-dimensional ACGs has been characterized almost completely. More detailed characterization of the hierarchy of second-order two-dimensional ACGs is future work.

Those results show that the ACG formalism can be thought of as a generalization of existing grammar formalisms in various meanings. ACGs encode not only individual grammars, but also operations on them, i.e., linearization and lexicalization. Moreover, ACGs generalize two-dimensional formalisms as well as one-dimensional ones. This thesis demonstrates that investigating ACGs is to study various mathematical properties of many other well-established grammar formalisms including, but not limited to, mildly context-sensitive grammars.

On the other hand, in Chapter 6, we have proved that third-order interpolation in the linear lambda calculus is NP-complete, and moreover, fourthorder interpolation in the linear lambda calculus is NP-complete even when we exclude constants from the problem instances. Those issues are not about the ACG formalism itself, though they (in particular the former) are motivated by a topic on ACGs.

Since the ACG is a simple formalism based on simply typed linear lambda calculus, investigation of ACGs sometimes offers motivations and topics to the fields of lambda-calculus and linear logic. For instance, Salvati [43] introduces an extended type system for the linear lambda calculus, called syntactic descriptions, for parsing second-order ACGs, and he also proposes an algorithm for solving the linear matching problem in the linear lambda calculus that utilizes syntactic descriptions [44]. Kanazawa [25] presents a new proof for the Interpolation Theorem for the implicational fragment of intuitionistic logic and its substructural subsystems, which would be used for an algorithm that learns two-dimensional ACGs generating pairs of a string and its meaning [24]. Conversely, since the simply typed lambda calculus has been studied very well so far, we have plenty of useful tools for investigating the mathematical properties of ACGs. Indeed, the basic mathematical properties of (lexicalized) ACGs shown in Chapter 2 and 5 are directly derived from existing theorems on the intuitionistic linear logic, and moreover, other theorems presented in this thesis rely on existing theorems and techniques in the intuitionistic linear logic implicitly or explicitly. This thesis strengthens the concept that the ACG formalism might be a linkage between mathematical linguistics and logic.

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[^0]:    ${ }^{1} \mathscr{L}_{2}(\mathrm{~A})$ and $\mathscr{L}_{2}(\mathrm{E})$ are not represented as simply typed linear $\lambda$-terms here. Nevertheless, by introducing constants Some and Every of type $(e \rightarrow t)^{2} \rightarrow t$, we can represent them by linear $\lambda$-terms as $\mathscr{L}_{2}(\mathbf{A})=\lambda w_{1} w_{2} . \operatorname{Some} w_{1} w_{2}$ and $\mathscr{L}_{2}(\mathbb{E})=\lambda w_{1} w_{2}$.Every $w_{1} w_{2}$.

[^1]:    ${ }^{1}$ Yoshinaka and Kanazawa [61] have presented an example of a fourth-order lexicalized ACG generating an NP-complete language independently from Salvati.

[^2]:    ${ }^{2}$ The idea presented in this subsection is based on a discussion among speakers, Ph. de Groote, S. Salvati, R. Muskens, M. Kanazawa, and Yoshinaka, of the First Workshop on Lambda Calculus and Formal Grammar held in 2005, Tokyo.

[^3]:    ${ }^{3}$ Recall that a lexical entry of an ACG is a triple consisting of an abstract constant, its type, and the assigned object term.

[^4]:    ${ }^{4}$ For instance，we replace $\mathrm{D}_{\mathrm{ac} q^{\prime}}$ with two new constants that have different types $\vec{t} \rightarrow$ $\left[p^{\prime} \rightarrow p\right] \rightarrow\left[q^{\prime} \rightarrow q\right] \rightarrow q^{\prime} \rightarrow r$ and $\vec{t} \rightarrow\left[p^{\prime} \rightarrow p\right] \rightarrow\left[q^{\prime} \rightarrow q\right] \rightarrow\left[q^{\prime} \rightarrow r\right]$.

[^5]:    ${ }^{5}$ Substitute $K_{k}\left[\vec{M}_{k} / \vec{x}_{k}\right]$ for $M$ in Case $3 . K_{k}$ in Case 4 and $M_{0}$ in Case 3 have the same form.

[^6]:    ${ }^{6}$ Our actual method generates a lexicalized ACG whose lexical entries are a little more complicated.

[^7]:    ${ }^{7}$ If $N_{i, 0}$ is in long normal form, a subterm of that form is found. We can assume it without loss of generality.
    ${ }^{8}$ A permutation of the arguments of $x_{i, j}$ can be performed with the same permutation on the order of the abstraction in $M_{i, j}$.

[^8]:    ${ }^{9}$ The author could not find a reference on the complexity of a fixed Petri-net reachability set.

[^9]:    ${ }^{1} \mathrm{~A} \lambda \mathrm{~K}$-lexicon can map a constant to non-linear $\lambda$-terms. See Section 3.2.

[^10]:    ${ }^{2}$ Precisely, $\vec{q}_{h}$ is the sequence consisting of $q_{i, j, g}$ with $d_{i, j}=h$ such that if $\vec{q}_{h}=$ $\vec{r}_{1} q_{i_{1}, j_{1}, g_{1}} q_{i_{2}, j_{2}, g_{2}} \vec{r}_{2}$, then either $i_{1}<i_{2}$ or $i_{1}=i_{2} \wedge j_{1}<j_{2}$ or $i_{1}=i_{2} \wedge j_{1}=j_{2} \wedge g_{1}<g_{2}$ holds. In the remainder of this section, if a sequence is defined without specifying the order of the elements, they are ordered in this manner according to their indexes.

[^11]:    ${ }^{3}$ Precisely speaking, Salvati gives a slight modification on the given ACG $\mathscr{G}_{0}$ before constructing a DTWT. Nevertheless, it is not an obstacle to our discussion at all, since the same modification can be done on 2D-ACGs.

[^12]:    ${ }^{1}$ The polarity is not a mandatory requirement for this proof.

[^13]:    ${ }^{2}$ An inessential change is made to the original reduction to facilitate comparison with our reduction.

