# An Optimal Algorithm for Constructing the Reduced Gröbner Basis of Binomial Ideals 

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#### Abstract

In this paper, we present an optimal, exponential space algorithm for generating the reduced Gröbner basis of binomial ideals. We make use of the close relationship between commutative semigroups and pure difference binomial ideals. Based on an optimal algorithm for the uniform word problem in commutative semigroups, we first derive an exponential space algorithm for constructing the reduced Gröbner basis of pure difference binomial ideals. In addition to some applications to finitely presented commutative semigroups, this algorithm is then extended to an exponential space algorithm for generating the reduced Gröbner basis of binomial ideals over $\mathbb{Q}$ in general.


## 1. Introduction

One of the most active areas of research in computer algebra is the design and analysis of algorithms for computational problems in commutative algebra. In particular, computational problems for polynomial ideals occur, as mathematical subproblems, in many areas of mathematics, and they also have a number of applications in various areas of computer science, such as language generating and term rewriting systems, tiling problems, algebraic manifolds, motion planning, and several models for parallel systems.

Using Gröbner bases (see Buchberger, 1965, 1976, 1985; also Hironaka, 1964) many of these problems become easily expressible and algorithmically solvable. For practical applications, in particular, the implementation in computer algebra systems, it is important to establish upper complexity bounds for the normal form algorithms which transform a given polynomial ideal basis into a Gröbner basis. First steps were obtained in David and Bayer (1982) and Möller and Mora (1984) where upper bounds for the degrees in a minimal Gröbner basis were derived. In Dubé (1990), Dubé obtained an improved degree bound of $2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$ (with $d$, the maximum degree of the input basis and $k$, the number of indeterminates) for the degree of polynomials in a reduced Gröbner basis employing only combinatorial arguments. By transforming a representation of the normal form of a polynomial into a system of linear equations, Kühnle and Mayr (1996) exhibited an exponential space computation of Gröbner bases. Their algorithm, however, is based on rather complex parallel computations like parallel rank computations of matrices, and, above that, makes extensive use of the parallel computation thesis of Fortune and Wyllie (1978).
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In this paper, we exploit the close relationship between commutative semigroups and pure difference binomial ideals (for an investigation of the algebraic structure of general binomial ideals see Eisenbud and Sturmfels (1996); concerning toric ideals, a special subset of pure difference binomial ideals, some results based on the connections between toric ideals and integer programming can be found in e.g. Di Biase and Urbanke (1995); Hoşten and Sturmfels (1995); Thomas (1998); Hoşten and Thomas (1998)). Based on the algorithm in Mayr and Meyer (1982) for the uniform word problem in commutative semigroups, we derive an exponential space algorithm for constructing the reduced Gröbner basis of a general binomial ideal over $\mathbb{Q}$. This algorithm can be implemented, in the case of pure difference binomial ideals, without any difficult parallel rank computations of matrices, or any other complex parallel computations. By the results in Mayr and Meyer (1982) and Huynh (1986), which give a doubly exponential lower bound (in the size of the problem instance) for the maximal degree of the elements of Gröbner bases of pure difference binomial ideals as well as for the cardinality of such bases, our algorithm is space optimal.

Thus, our algorithm and the complexity bounds reported in this paper completely characterize the (asymptotic) computational complexity of Gröbner basis computations for general binomial ideals by basically making use of the close relationship between commutative semigroups and binomial ideals. We do not consider other techniques commonly used for computing Gröbner bases of ideals and modules, such as critical pairs and completion, because their actual computational complexity is much more complex to investigate. And, for most of these algorithms, the space complexity is doubly exponential, one exponential worse than our algorithm.
The remainder of this paper is organized as follows. In Section 2, we briefly introduce the basic notations and fundamental concepts. In Section 3, we derive an exponential space algorithm for constructing the reduced Gröbner basis of pure difference binomial ideals, and we give some applications to finitely presented commutative semigroups. Then, in Section 4, this algorithm is extended to an exponential space algorithm for generating the reduced Gröbner basis of binomial ideals in general.

## 2. Preliminaries

### 2.1. BASIC DEFINITIONS AND NOTATIONS

The polynomial ideals which we obtain by using the relationship of finitely presented commutative semigroups and polynomial ideals are pure difference binomial ideals, i.e. ideals that have a basis consisting only of differences of two terms. By looking at Buchberger's algorithm (Buchberger, 1965), it is not hard to see that the reduced Gröbner basis of a pure difference binomial ideal still consists only of pure difference binomials.
Let $X$ denote the finite set $\left\{x_{1}, \ldots, x_{k}\right\}$ and ${ }^{\dagger} \mathbb{Q}[X]$ the (commutative) ring of polynomials with indeterminates $x_{1}, \ldots, x_{k}$ and rational coefficients. A term $t$ in $x_{1}, \ldots, x_{k}$ is a product of the form

$$
t=x_{1}^{e_{1}} \cdot x_{2}^{e_{2}} \cdots x_{k}^{e_{k}},
$$

with $\left(e_{1}, e_{2}, \ldots, e_{k}\right) \in \mathbb{N}^{k}$ the degree vector of $t$. By the degree $\operatorname{deg}(t)$ of a term $t$, we shall mean the integer $e_{1}+e_{2}+\cdots+e_{k}$ (which is $\geq 0$ ).

[^0]Each polynomial $f\left(x_{1}, \ldots, x_{k}\right) \in \mathbb{Q}[X]$ is a finite sum

$$
f\left(x_{1}, \ldots, x_{k}\right)=\sum_{i=1}^{n} a_{i} \cdot t_{i}
$$

with $a_{i} \in \mathbb{Q} \backslash\{0\}$ the coefficient of the $i$ th term $t_{i}$ of $f$. The product $m_{i}=a_{i} \cdot t_{i}$ is called the $i$ th monomial of the polynomial $f$. The degree of a polynomial is the maximum of the degrees of its terms.
For $f_{1}, \ldots, f_{h} \in \mathbb{Q}[X],\left\langle f_{1}, \ldots, f_{h}\right\rangle \subseteq \mathbb{Q}[X]$ denotes the ideal generated by $\left\{f_{1}, \ldots, f_{h}\right\}$ that is ${ }^{\dagger}$

$$
\left\langle f_{1}, \ldots, f_{h}\right\rangle:=\left\{\sum_{i=1}^{h} p_{i} f_{i} ; p_{i} \in \mathbb{Q}[X] \text { for } i \in I_{h}\right\} .
$$

Whenever $I=\left\langle f_{1}, \ldots, f_{h}\right\rangle,\left\{f_{1}, \ldots, f_{h}\right\}$ is called a basis of $I$.
An admissible term ordering $\succeq$ is given by any admissible ordering on $\mathbb{N}^{k}$, i.e. any total ordering $\geq$ on $\mathbb{N}^{k}$ satisfying the following two conditions:

$$
\begin{equation*}
e \geq(0, \ldots, 0) \text { for all } e \in \mathbb{N}^{k} \tag{T1}
\end{equation*}
$$

(T2) $\quad a>b \quad \Rightarrow \quad a+c>b+c$ for all $a, b, c \in \mathbb{N}^{k}$.
If $\left(d_{1}, \ldots, d_{k}\right)>\left(e_{1}, \ldots, e_{k}\right)$, we say that the term $x_{1}^{d_{1}} \cdots x_{k}^{d_{k}}$ is greater in the term ordering than the term $x_{1}^{e_{1}} \cdots x_{k}^{e_{k}}$ (written $x_{1}^{d_{1}} \cdots x_{k}^{d_{k}} \succ x_{1}^{e_{1}} \cdots x_{k}^{e_{k}}$ ).
For a polynomial $f\left(x_{1}, \ldots, x_{k}\right)=\sum_{i=1}^{n} a_{i} \cdot t_{i}$ we always assume that $t_{1} \succ t_{2} \succ \cdots \succ t_{n}$. For any such non-zero polynomial $f \in \mathbb{Q}[X]$, we define the leading term $L T(f):=t_{1}$.
For the sake of constructiveness, we assume that the term ordering is given as part of the input by a $k \times k$ integer matrix $T$ such that $x_{1}^{d_{1}} \cdots x_{k}^{d_{k}} \succ x_{1}^{e_{1}} \cdots x_{k}^{e_{k}}$ iff, for the corresponding degree vectors $d$ and $e, T d$ is lexicographically greater than $T e$ (see Robbiano, 1985; Weispfenning, 1987).

Let $I$ be an ideal in $\mathbb{Q}[X]$, and let some admissible term ordering $\succeq$ be given. A finite subset $\left\{g_{1}, \ldots, g_{r}\right\} \subseteq I$ of polynomials in $I$ is called a Gröbner basis of $I$ (w.r.t. $\succeq$ ), if
(G) $\quad\left\{L T\left(g_{1}\right), \ldots, L T\left(g_{r}\right)\right\}$ is a basis of the leading term ideal $L T(I)$ of $I$, which is the smallest ideal containing the leading terms of all $f \in I$, or equivalently: if $f \in I$, then $L T(f) \in\left\langle L T\left(g_{1}\right), \ldots, L T\left(g_{r}\right)\right\rangle$.
A Gröbner basis is called reduced if no monomial in any one of its polynomials is divisible by the leading term of any other polynomial in the basis.
For a finite alphabet $X=\left\{x_{1}, \ldots, x_{k}\right\}$, let $X^{*}$ denote the free commutative monoid generated by $X$. An element $u$ of $X^{*}$ is called a (commutative) word. For a word the order of the symbols is immaterial, and we shall in the sequel use an exponent notation: $u=x_{1}^{e_{1}} \ldots x_{k}^{e_{k}}$, where ${ }^{\ddagger} e_{i}=\Phi\left(u, x_{i}\right) \in \mathbb{N}$ for $i=1, \ldots, k$. We identify any $u \in X^{*}$ (resp., the corresponding vector $u=\left(\Phi\left(u, x_{1}\right), \ldots, \Phi\left(u, x_{k}\right)\right) \in \mathbb{N}^{k}$ ) with the term $u=$ $x_{1}^{\Phi\left(u, x_{1}\right)} \cdot x_{2}^{\Phi\left(u, x_{2}\right)} \cdots x_{k}^{\Phi\left(u, x_{k}\right)}$ and vice versa.

Let $\mathcal{P}=\left\{l_{i} \equiv r_{i} ; i \in I_{h}\right\}$ be some (finite) commutative semigroup presentation with $l_{i}, r_{i} \in X^{*}$ for $i \in I_{h}$. We say that a word $v \in X^{*}$ is derived in one step from $u \in X^{*}$ (written $u \rightarrow v(\mathcal{P})$ ) by application of the congruence $\left(l_{i} \equiv r_{i}\right) \in \mathcal{P}$ iff, for some $w \in X^{*}$, we have $u=w l_{i}$ and $v=w r_{i}$, or $u=w r_{i}$ and $v=w l_{i}$ (note, since " $\equiv "$ is symmetric, " $\rightarrow$ " is symmetric, i.e. $u \rightarrow v(\mathcal{P}) \Leftrightarrow v \rightarrow u(\mathcal{P})$ ). The word $u$ derives $v$, written $u \equiv v \bmod \mathcal{P}$,

[^1]iff $u \xrightarrow{*} v(\mathcal{P})$, where $\xrightarrow{*}$ is the reflexive transitive closure of $\rightarrow$. More precisely, we write $u \xrightarrow{+} v(\mathcal{P})$, where $\xrightarrow{+}$ is the transitive closure of $\rightarrow$, if $u \xrightarrow{*} v(\mathcal{P})$ and $u \neq v$. A sequence $\left(u_{0}, \ldots, u_{n}\right)$ of words $u_{i} \in X^{*}$ with $u_{i} \rightarrow u_{i+1}(\mathcal{P})$ for $i=0, \ldots, n-1$, is called a derivation (of length $n$ ) of $u_{n}$ from $u_{0}$ in $\mathcal{P}$. The congruence class of $u \in X^{*}$ modulo $\mathcal{P}$ is the set $[u]_{\mathcal{P}}=\left\{v \in X^{*} ; u \equiv v \bmod \mathcal{P}\right\}$.

By $I(\mathcal{P})$, we denote the $\mathbb{Q}[X]$-ideal generated by $\left\{l_{1}-r_{1}, \ldots, l_{h}-r_{h}\right\}$, i.e.

$$
I(\mathcal{P}):=\left\{\sum_{i=1}^{h} p_{i}\left(l_{i}-r_{i}\right) ; p_{i} \in \mathbb{Q}[X] \text { for } i \in I_{h}\right\}
$$

### 2.2. THE UNIFORM WORD PROBLEM AND THE CORRESPONDING PURE DIFFERENCE BINOMIAL IDEAL MEMBERSHIP PROBLEM

The uniform word problem for commutative semigroups is the problem of deciding for a commutative semigroup presentation $\mathcal{P}$ over some alphabet $X$ and two words $u, v \in X^{*}$ whether $u \equiv v \bmod \mathcal{P}$. The polynomial ideal membership problem is the problem of deciding for given polynomials $f, f_{1}, \ldots, f_{h} \in \mathbb{Q}[X]$ whether $f \in\left\langle f_{1}, \ldots, f_{h}\right\rangle$.

Proposition 2.1. (Mayr and Meyer, 1982) Let $X=\left\{x_{1}, \ldots, x_{k}\right\}$ be some finite alphabet, $\mathcal{P}=\left\{l_{i} \equiv r_{i} ; i \in I_{h}\right\}$ a finite commutative semigroup presentation over $X$, and $u$, $v$ two words in $X^{*}$ with $u \neq v$. Then from $u \equiv v \bmod \mathcal{P}$ it follows that $u-v \in I(\mathcal{P})$, and vice versa, i.e. if there exist $p_{1}, \ldots, p_{h} \in \mathbb{Q}[X]$ such that $u-v=\sum_{i=1}^{h} p_{i}\left(l_{i}-r_{i}\right)$, then there is a derivation $u=\gamma_{0} \rightarrow \gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n}=v(\mathcal{P})$ of $v$ from $u$ in $\mathcal{P}$ such that, for $j \in\{0,1, \ldots, n\}$,

$$
\operatorname{deg}\left(\gamma_{j}\right) \leq \max \left\{\operatorname{deg}\left(l_{i} p_{i}\right), \operatorname{deg}\left(r_{i} p_{i}\right) ; i \in I_{h}\right\}
$$

In the fundamental paper (Hermann, 1926), Hermann gave a doubly exponential degree bound for the polynomial ideal membership problem:

Proposition 2.2. (Hermann, 1926) Let $X=\left\{x_{1}, \ldots, x_{k}\right\}, f, f_{1}, \ldots, f_{h} \in \mathbb{Q}[X]$, and $d=\max \left\{\operatorname{deg}\left(f_{i}\right) ; i \in I_{h}\right\}$. If $f \in\left\langle f_{1}, \ldots, f_{h}\right\rangle$, then there exist $p_{1}, \ldots, p_{h} \in \mathbb{Q}[X]$ such that:
(i) $f=\sum_{i=1}^{h} p_{i} f_{i}$, and
(ii) $\left.\operatorname{deg}\left(p_{i}\right) \leq \operatorname{deg}(f)+(h d)^{2^{k}}\right]$ for all $i \in I_{h}$.

By size $(\cdot)$ we shall denote the number of bits needed to encode the argument in some standard way (using radix representation for numbers).

Then the two propositions yield an exponential space upper bound for the uniform word problem for commutative semigroups:

Proposition 2.3. (Mayr and Meyer, 1982) Let $X=\left\{x_{1}, \ldots, x_{k}\right\}$ be some finite alphabet, and $\mathcal{P}=\left\{l_{i} \equiv r_{i} ; i \in I_{h}\right\}$ a finite commutative semigroup presentation over $X$. Then there is a (deterministic) Turing machine $M$ and some constant $c>0$ independent of $\mathcal{P}$, such that $M$ decides for any two words $u, v \in X^{*}$ whether $u \equiv v \bmod \mathcal{P}$ using at most space $(\operatorname{size}(u, v, \mathcal{P}))^{2} \cdot 2^{c \cdot k}$.

## 3. Constructing the Reduced Gröbner Basis of a Pure Difference Binomial Ideal in Exponential Space

In this section, we derive an exponential space algorithm for generating the reduced Gröbner basis of a pure difference binomial ideal. For this purpose, we first analyze the elements of the reduced Gröbner basis.

### 3.1. THE REDUCED GRÖBNER BASIS OF PURE DIFFERENCE BINOMIAL IDEALS

Let $\mathcal{P}$ be a commutative semigroup presentation over some alphabet $X$. If $C$ is some congruence class of $\mathcal{P}$, and $G$ is a Gröbner basis of the pure difference binomial ideal $I(\mathcal{P})$ w.r.t. some admissible term ordering $\succeq$, then the minimal element $m_{C}$ of $C$ w.r.t. $\succ$ is not reducible modulo $G$.

Proposition 3.1. (Huynh, 1986) Let $X=\left\{x_{1}, \ldots, x_{k}\right\}, \mathcal{P}=\left\{l_{i} \equiv r_{i} ; i \in I_{h}\right\}$ with $l_{i}, r_{i} \in X^{*}$ for all $i \in I_{h}$, and let $G=\left\{h_{1}-m_{1}, \ldots, h_{r}-m_{r}\right\}$ be the reduced Gröbner basis of the ideal $I(\mathcal{P})$ w.r.t. some admissible term ordering $\succeq\left(h_{i} \succ m_{i}\right.$ for all $\left.i \in I_{r}\right)$. Then:
(i) $m_{i}$ is the minimal element (w.r.t. $\succ$ ) of the congruence class $\left[h_{i}\right]_{\mathcal{P}}, i \in I_{r}$.
(ii) $L T(I(\mathcal{P})$ ) (the set of the leading terms of $I(\mathcal{P})$ ) is the set of all terms with nontrivial congruence class which are not the minimal element in their congruence class w.r.t. $\succ . H=\left\{h_{1}, \ldots, h_{r}\right\}$ is the set of the minimal elements of $L T(I(\mathcal{P}))$ w.r.t. divisibility.

If $s \in X^{*}$ is the minimal element of its congruence class $[s]_{\mathcal{P}}$ w.r.t. $\succ$, then every subword $s^{\prime}$ of $s$ is also the minimal element of its congruence class $\left[s^{\prime}\right]_{\mathcal{P}}$ w.r.t. $\succ$.

### 3.2. THE ALGORITHM

In this section, we give an exponential space algorithm for generating the reduced Gröbner basis of a pure difference binomial ideal. To show the correctness and the complexity of the algorithm, we need the results of the previous sections and the following upper bound for the total degree of polynomials in a Gröbner basis, obtained by Dubé (1990). Note that we use exponential notation in representing words over $X$.

Proposition 3.2. (Dubé, 1990) Let $F=\left\{f_{1}, \ldots, f_{h}\right\} \subset \mathbb{Q}[X]=\mathbb{Q}\left[x_{1}, \ldots, x_{k}\right], I=$ $\left\langle f_{1}, \ldots, f_{h}\right\rangle$ the ideal of $\mathbb{Q}[X]$ generated by $F$, and let $d$ be the maximum degree of any $f \in F$. Then for any admissible term ordering $\succeq$, the degree of polynomials required in a Gröbner basis for I w.r.t. $\succeq$ is bounded by

$$
2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}
$$

Let $X=\left\{x_{1}, \ldots, x_{k}\right\}, \succeq$ an admissible term ordering on $X^{*}$, and $\mathcal{P}=\left\{l_{i} \equiv r_{i} ; i \in I_{h}\right\}$ where, for all $i \in I_{h}, l_{i}, r_{i} \in X^{*}$ and w.l.o.g. $l_{i} \succ r_{i}$. We shall give an exponential space algorithm for generating the reduced Gröbner basis of the pure difference binomial ideal $I(\mathcal{P})$ w.r.t. $\succeq$. Let $H$ denote the set $\left\{h_{1}, \ldots, h_{r}\right\}$ of the minimal elements of $L T(I(\mathcal{P}))$
w.r.t. divisibility, and $m_{i}$ the minimal element of $\left[h_{i}\right]_{\mathcal{P}}$ w.r.t. $\succ$, for $i \in I_{r}$. From Proposition 3.1, we know that the set $G=\left\{h_{1}-m_{1}, \ldots, h_{r}-m_{r}\right\}$ is the reduced Gröbner basis of $I(\mathcal{P})$ w.r.t. $\succeq$.
Proposition 3.1 shows that $L T(I(\mathcal{P})) \supseteq\left\{l_{1}, \ldots, l_{h}\right\}$ and that $L T(I(\mathcal{P})) \subseteq\left\langle l_{1}, \ldots, l_{h}\right.$, $\left.r_{1}, \ldots, r_{h}\right\rangle$. Let $L$ (resp., $R$ ) be the subset of $\left\{l_{1}, \ldots, l_{h}\right\}$ (resp., $\left\{r_{1}, \ldots, r_{h}\right\}$ ) containing all those elements which are also minimal (w.r.t. divisibility) in $\left\{l_{1}, \ldots, l_{h}, r_{1}, \ldots, r_{h}\right\}$.

Then $H \supseteq L$, and we still have to determine the elements in $H \backslash L$, as well as the minimal element $m_{i}$ (w.r.t. $\succ$ ) of the congruence class of each $h_{i} \in H$. From Proposition 3.2, we know that the degrees $\operatorname{deg}\left(h_{i}\right)$ and $\operatorname{deg}\left(m_{i}\right)$ are bounded by $2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$, where $d$ is the maximum degree of the $l_{i}-r_{i},\left(l_{i} \equiv r_{i}\right) \in \mathcal{P}$. Since $H \backslash L \subseteq L T(\langle L, R\rangle) \backslash L$, we consider the terms in $L T(\langle L, R\rangle) \backslash L$ with degree $\leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$.

Lemma 3.1. For a term $u \in X^{*}$ with non-trivial congruence class, the minimal element w.r.t. $\succ$ of $[u]_{\mathcal{P}}$ is of the form $t \cdot r_{i}$ with $r_{i} \in R, t \in X^{*}$.

Proof. W.l.o.g. assume that $u$ is not the minimal element $m_{u}$ of $[u]_{\mathcal{P}}$ w.r.t. $\succ$. Then there is a derivation in $\mathcal{P}$ leading from $u$ to $m_{u} \prec u$, i.e. $u \xrightarrow{+} m_{u}(\mathcal{P})$, where $m_{u}=t \cdot r_{i}$ for some $r_{i} \in R, t \in X^{*}$ (note that $l_{j} \succ r_{j}$ for all $j \in I_{h}$ ).

For $h=x_{1}^{e_{1}} \cdots x_{k}^{e_{k}} \in X^{*}$ and $i \in I_{k}$ such that $e_{i} \geq 1$, define $h^{(i)}:=x_{1}^{e_{1}} \cdots x_{i}^{e_{i}-1} \cdots x_{k}^{e_{k}}$. Then $h \in H$ iff, for all $i \in I_{k}$ with $e_{i} \geq 1, h^{(i)} \notin L T(I(\mathcal{P}))$, i.e. $h^{(i)}$ is the minimal element of $\left[h^{(i)}\right]_{\mathcal{P}}$ w.r.t. $\succ$. Thus, $H$ consists exactly of those terms $h \in X^{*}$ which have degree $\leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$, which are congruent to some term $t \cdot r_{i} \prec h$ with $r_{i} \in R$, $t \in X^{*}$, and $\operatorname{deg}\left(t \cdot r_{i}\right) \leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$, and for which, for all applicable $i,\left[h^{(i)}\right]_{\mathcal{P}}$ is trivial.
For terms $h$ and $t \cdot r_{i}$ with $\operatorname{deg}(h)$ and $\operatorname{deg}\left(t \cdot r_{i}\right) \leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$, by Proposition 2.3, the decision whether $h \equiv t \cdot r_{i} \bmod \mathcal{P}$ uses at most space $(\operatorname{size}(\mathcal{P}))^{2} \cdot 2^{c \cdot k}$ for some constant $c>0$ independent of $\mathcal{P}$. Hence, the condition regarding the reducibility of $h$ can also be checked in space $(\operatorname{size}(\mathcal{P}))^{2} \cdot 2^{c \cdot k}$. We decide for the words $t \cdot r_{i}$ with $h \succ t \cdot r_{i}, r_{i} \in R$, $t \in X^{*}, \operatorname{deg}\left(t \cdot r_{i}\right) \leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$ in ascending order w.r.t. $\succ$ whether $t \cdot r_{i} \equiv h \bmod \mathcal{P}$ until we find the minimal element $m_{h}$ of $[h]_{\mathcal{P}}$, or there is no more $t \cdot r_{i}$ with $h \succ t \cdot r_{i}$, $r_{i} \in R, t \in X^{*}, \operatorname{deg}\left(t \cdot r_{i}\right) \leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$. In the latter case, $h \notin H$ and we have to consider the next element of $L T(\langle L, R\rangle) \backslash L$ with degree $\leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$. Otherwise, $h \in L T(I(\mathcal{P}))$ and we have to determine whether $h \in H$.
Testing non-reducibility of the $h^{(i)}$ can also be done in exponential space because of Proposition 2.3 and

Lemma 3.2. A term $u \in X^{*}$ with $\operatorname{deg}(u) \leq \bar{d}$ is an element of $\operatorname{LT}(I(\mathcal{P}))$ iff there is some $t \cdot r_{i}$ with $u \succ t \cdot r_{i}, r_{i} \in R, t \in X^{*}$, and $\operatorname{deg}\left(t \cdot r_{i}\right) \leq \bar{d}+2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$ such that $u \xrightarrow{+} t \cdot r_{i}(\mathcal{P})$.

Proof. We only have to prove the degree bound. Note that $u \in L T(I(\mathcal{P}))$ iff either $u \in H$, and thus, $\operatorname{deg}\left(m_{u}\right) \leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$, where $m_{u}$ is the minimal (w.r.t. $\succ$ ) element of $[u]_{\mathcal{P}}$, or there is some $h \in H$ with $u=t_{u} \cdot h$ for some $t_{u} \in X^{*}$. The degree of the minimal (w.r.t. $\succ$ ) element $m_{h}$ of $[h]_{\mathcal{P}}$ is bounded by $2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$. Since $\succeq$ is an admissible term ordering, from $h \succ m_{h}$ we obtain $u \succ t_{u} \cdot m_{h}$ with $\operatorname{deg}\left(t_{u} \cdot m_{h}\right) \leq \bar{d}+2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$.

From this, we derive the exponential space algorithm given in Figure 1.

Putting everything together, we proved the theorem:
Theorem 3.1. Let $X=\left\{x_{1}, \ldots, x_{k}\right\}, \mathcal{P}=\left\{l_{i} \equiv r_{i} ; i \in I_{h}\right\}$ with $l_{i}, r_{i} \in X^{*}$ for all $i \in I_{h}$, and $\succeq$ some admissible term ordering. Then there is an algorithm which generates the reduced Gröbner basis $G=\left\{h_{1}-m_{1}, \ldots, h_{r}-m_{r}\right\}$ of the pure difference binomial ideal $I(\mathcal{P})$ w.r.t. $\succeq$ using at most space $(\operatorname{size}(\mathcal{P}))^{2} \cdot 2^{\bar{c} \cdot k} \leq 2^{c \cdot \operatorname{size}(\mathcal{P})}$, where $\bar{c}, c>0$ are some constants independent of $\mathcal{P}$.

From the results in Huynh (1986), we know that, in the worst case, any Gröbner basis of $I(\mathcal{P})$ has maximal degree at least $2^{2^{c^{\prime} \cdot \text { size }(\mathcal{P})}}$ for some constant $c^{\prime}>0$ independent of $\mathcal{P}$. Hence, any algorithm that computes Gröbner bases of pure difference binomial ideals requires at least exponential space in the worst case.

### 3.3. APPLICATIONS

## TESTING FOR REDUCIBILITY

Let $\mathcal{P}$ be a finite commutative semigroup presentation over some finite alphabet $X$, $u \in X^{*}$, and $\succeq$ some admissible term ordering. Then $u$ is the minimal element of $[u]_{\mathcal{P}}$ w.r.t. $\succ$ iff $u$ is in normal form modulo a Gröbner basis $G$ of $I(\mathcal{P})$ w.r.t. $\succeq$, i.e. $u$ is not reducible modulo $G$. Thus, by Proposition 3.1, $u$ is in normal form modulo $G$ iff $u \notin L T(I(\mathcal{P}))$.

Corollary 3.1. Let $X=\left\{x_{1}, \ldots, x_{k}\right\}$, $\mathcal{P}=\left\{l_{i} \equiv r_{i} ; i \in I_{h}\right\}$ with $l_{i}, r_{i} \in X^{*}$ for all $i \in I_{h}$, and $\succeq$ some admissible term ordering. Then for any $u \in X^{*}$ there is an algorithm which decides whether $u \in L T(I(\mathcal{P}))$, as well as whether $u$ is the minimal element of its congruence class (w.r.t. $\succ$ ), i.e. $u$ is in normal form modulo a Gröbner basis of $I(\mathcal{P})$ w.r.t. $\succeq$, using at most space $\operatorname{size}(u)+(\operatorname{size}(\mathcal{P}))^{2} \cdot 2^{\bar{c} \cdot k} \leq \operatorname{size}(u)+2^{\text {c.size }(\mathcal{P})}$, where $\bar{c}$, $c>0$ are some constants independent of $u$ and $\mathcal{P}$.

Proof. Let $G=\left\{h_{1}-m_{1}, \ldots, h_{r}-m_{r}\right\}$ be the reduced Gröbner basis of $I(\mathcal{P})$. Then $L T(I(\mathcal{P}))$ is generated by $\left\{h_{1}, \ldots, h_{r}\right\}$, and $u \in L T(I(\mathcal{P}))$ iff there is some $h_{i}, i \in I_{r}$, which divides $u$. Hence, Corollary 3.1 is a direct consequence of Theorem 3.1.

## Constructing the Reduced Gröbner Basis of a Pure Difference Binomial Ideal

```
Input: admissible term ordering \(\succeq\),
            \(\mathcal{P}=\left\{l_{1}-r_{1}, \ldots, l_{h}-r_{h}\right\}\) with \(l_{i}, r_{i} \in X^{*}, l_{i} \succ r_{i} \forall i \in I_{h}\)
Output: the reduced Gröbner basis \(G=\left\{h_{1}-m_{1}, \ldots, h_{r}-m_{r}\right\}\) of \(I(\mathcal{P})\)
\(L:=\) the subset of elements of \(\left\{l_{1}, \ldots, l_{h}\right\}\) that are minimal in \(\left\{l_{1}, \ldots, l_{h}, r_{1}, \ldots, r_{h}\right\}\) w.r.t.
    divisibility
\(R:=\) the subset of elements of \(\left\{r_{1}, \ldots, r_{h}\right\}\) that are minimal in \(\left\{l_{1}, \ldots, l_{h}, r_{1}, \ldots, r_{h}\right\}\) w.r.t.
    divisibility
\(k:=\) number of indeterminates \(; d:=\max \left\{\operatorname{deg}\left(l_{i}\right), \operatorname{deg}\left(r_{i}\right) ; i \in I_{h}\right\} ; G:=\emptyset\)
for each \(h=x_{1}^{e_{1}} \cdots x_{k}^{e_{k}} \in L T(\langle L, R\rangle) \backslash L\) with degree \(\leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}\) do
    \(g b:=\) false
    if there exists \(t \cdot r_{i}\) with \(h \succ t \cdot r_{i}, r_{i} \in R, t \in X^{*}, \operatorname{deg}\left(t \cdot r_{i}\right) \leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}\)
        which is \(\equiv h \bmod \mathcal{P}\) then \(\quad /^{*} h \in L T(I(\mathcal{P})) \quad * /\)
        \(m:=\) the minimal (w.r.t. \(\succ\) ) among these terms
        \(g b:=\) true
    end_if
    if \(g b\) then \(\quad /^{*} h \in L T(I(\mathcal{P}))^{* /}\)
        \(\bar{d}:=\operatorname{deg}(h)\)
        for each \(i \in I_{k}\) with \(e_{i} \geq 1\) while \(g b\) do
            \(h^{\prime}:=x_{1}^{e_{1}} \cdots x_{i}^{e_{i}-1} \cdots x_{k}^{e_{k}}\)
            if there exists \(t \cdot r_{j}\) with \(h^{\prime} \succ t \cdot r_{j}, r_{j} \in R, t \in X^{*}, \operatorname{deg}\left(t \cdot r_{j}\right) \leq(\bar{d}-1)+2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}\)
                which is \(\equiv h^{\prime} \bmod \mathcal{P}\) then \(\quad /^{*} h^{\prime} \in L T(I(\mathcal{P})) \Rightarrow h \notin H \quad * /\)
                \(g b:=\) false
            end_if
        end_for
    end_if
    if \(g b\) then \(/^{*} h \in H^{* /}\)
        \(G:=G \cup\{h-m\}\)
    end_if
end_for
for each \(l_{i} \in L\) do
    \(m:=\) the minimal (w.r.t. \(\succ\) ) among the terms \(t \cdot r_{j}\) with \(l_{i} \succ t \cdot r_{j}, r_{j} \in R, t \in X^{*}\),
        \(\operatorname{deg}\left(t \cdot r_{j}\right) \leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}\) which are \(\equiv l_{i} \bmod \mathcal{P}\)
    \(G:=G \cup\left\{l_{i}-m\right\}\)
end_for
```

Figure 1. Algorithm for constructing the reduced Gröbner basis of a pure difference binomial ideal.

## FINDING THE MINIMAL ELEMENT AND THE NORMAL FORM

The next corollary shows that the minimal element of a congruence class, or equivalently, the normal form of a term can be found in exponential space.

Corollary 3.2. Let $X=\left\{x_{1}, \ldots, x_{k}\right\}$, $\mathcal{P}=\left\{l_{i} \equiv r_{i} ; i \in I_{h}\right\}$ with $l_{i}, r_{i} \in X^{*}$ for all $i \in I_{h}$, and $\succeq$ some admissible term ordering. Then there is an algorithm which determines for any word $u \in X^{*}$ the minimal element of its congruence class (w.r.t. $\succ$ ), or equivalently, which determines for any term $u \in X^{*}$ the normal form of $u$ modulo
a Gröbner basis of $I(\mathcal{P})$ w.r.t. $\succeq$, using at most space $(\operatorname{size}(u, \mathcal{P}))^{2} \cdot 2^{\bar{c} \cdot k} \leq 2^{c \cdot \operatorname{size}(u, \mathcal{P})}$, where $\bar{c}, c>0$ are some constants independent of $u$ and $\mathcal{P}$.

Proof. In addition to $x_{1}, \ldots, x_{k}$ we introduce a new variable $s$, and we add to $\mathcal{P}$ the congruence $s \equiv u$, where $u$ is the word in $X^{*}$ for whose congruence class we like to determine the minimal element $m_{u}$ (w.r.t. $\succ$ ). Let $X_{s}=X \cup\{s\}, \mathcal{P}_{s}=\mathcal{P} \cup\{s \equiv u\}$, and let $\succeq_{s}$ be the admissible term ordering on $X_{s}^{*}$ which results from $\succeq$ by adding $s \succ w$ for all $w \in X^{*}$. Then, by Proposition 3.1, $L T\left(I\left(\mathcal{P}_{s}\right)\right)=L T(I(\mathcal{P})) \cup\left\{s \cdot t ; t \in X_{s}^{*}\right\}$, in particular, $s \in L T\left(I\left(\mathcal{P}_{s}\right)\right)$, and, since $s$ is minimal in $L T\left(I\left(\mathcal{P}_{s}\right)\right)$ w.r.t. divisibility, $H_{s}=H \cup\{s\}$, where $H$ (resp., $H_{s}$ ) is the set of the minimal elements of $L T(I(\mathcal{P}))$ (resp., $L T\left(I\left(\mathcal{P}_{s}\right)\right)$ ) w.r.t. divisibility. Because $s \succ w$ for all $w \in X^{*}$, the minimal element of some congruence class $[v]_{\mathcal{P}_{s}}, v \in X^{*}$, w.r.t. $\succ_{s}$ is the same as the minimal element of $[v]_{\mathcal{P}}$ w.r.t. $\succ$. Thus, because of Proposition 3.1,s-mu is an element of the reduced Gröbner basis of $I(\mathcal{P})$ w.r.t. $\succeq$, and, by Theorem 3.1, we can determine the minimal element $m_{u}$ of $[u]_{\mathcal{P}}$ (w.r.t. $\succ)$ in space $(\operatorname{size}(u, \mathcal{P}))^{2} \cdot 2^{\bar{c} \cdot k}$ for some constant $\bar{c}>0$ independent of $u$ and $\mathcal{P}$.

## 4. Constructing the Reduced Gröbner Basis of a Binomial Ideal in Exponential Space

The algorithm of Theorem 3.1 generates the reduced Gröbner basis of pure difference binomial ideals. In this section, we will be concerned with constructing the reduced Gröbner basis of binomial ideals in general.

### 4.1. BASICS

Let $m=a \cdot t$ be a monomial in $\mathbb{Q}[X]$ with $a \in \mathbb{Q}$, and $t$ a term in $X^{*}$. Then we write $C(m)$ for the coefficient $a$, and $T(m)$ for the term $t$ of the monomial $m$. By $M[X]$ we denote the set of all monomials in $\mathbb{Q}[X]$, including 0 .

By a binomial in $\mathbb{Q}[X]$ we mean a polynomial with at most two monomials, say $l-r$. For a finite set $\mathcal{B}=\left\{l_{i}-r_{i} ; i \in I_{h}\right\}$ with $l_{i}, r_{i} \in M[X]$ for all $i \in I_{h}, I(\mathcal{B})$ denotes the binomial $\mathbb{Q}[X]$-ideal generated by $\mathcal{B}$, i.e.

$$
I(\mathcal{B}):=\left\{\sum_{i=1}^{h} p_{i}\left(l_{i}-r_{i}\right) ; p_{i} \in \mathbb{Q}[X] \text { for } i \in I_{h}\right\}
$$

W.l.o.g. we assume that there are no $i, j \in I_{h}, i \neq j$, with $l_{i}-r_{i}=c \cdot\left(l_{j}-r_{j}\right)$ for some $c \in \mathbb{Q} \backslash\{0\}$. (Otherwise we could remove one of the two binomials.)
As in the case of pure difference binomial ideals, we see from Buchberger's algorithm that the reduced Gröbner basis of a binomial ideal still consists only of binomials.

In the following, we generalize the algorithm of Theorem 3.1 from pure difference binomial ideals to binomial ideals. First, we establish some technical details.

For $X=\left\{x_{1}, \ldots, x_{k}\right\}$, and $\mathcal{B}=\left\{l_{i}-r_{i} ; i \in I_{h}\right\}$ a set of binomials in $\mathbb{Q}[X]$ with $l_{i} \in X^{*}$, i.e. $C\left(l_{i}\right)=1, T\left(l_{i}\right)=l_{i}$, and $r_{i} \in M[X]$ for all $i \in I_{h}$, we define the corresponding commutative semigroup presentation

$$
\mathcal{P}(\mathcal{B})=\left\{l_{i} \equiv T\left(r_{i}\right) ;\left(l_{i}-r_{i}\right) \in \mathcal{B}\right\}
$$

where we set

$$
T(0)=x_{1}^{-\infty} \cdot x_{2}^{-\infty} \cdots x_{k}^{-\infty}
$$

Let $\succeq$ be some admissible term ordering on $X^{*}$. Then we extend $\succeq$ to an admissible term ordering on $X_{0}^{*}=X^{*} \cup\left\{x_{1}^{-\infty} \cdots x_{k}^{-\infty}\right\}$ by setting, for all $t \in X^{*}$,

$$
t \succ x_{1}^{-\infty} \cdot x_{2}^{-\infty} \cdots x_{k}^{-\infty}
$$

If we agree that $-\infty+n=n+(-\infty)=-\infty$ for any integer $n$, and $-\infty+(-\infty)=-\infty$, then the whole formalism for commutative semigroups introduced in Section 2 still works for $\mathcal{P}(\mathcal{B})$. The only difference is that, in addition to the words in $X^{*}$, we have the word $x_{1}^{-\infty} \cdot x_{2}^{-\infty} \cdots x_{k}^{-\infty}$ which corresponds to 0 when we consider polynomials. In particular, we still have, for $u, v \in X_{0}^{*}$,

$$
u-v \in I(\mathcal{P}(\mathcal{B})) \quad \Longleftrightarrow \quad u \equiv v \bmod \mathcal{P}(\mathcal{B})
$$

W.l.o.g. we assume that, for all $i \in I_{h}$,

$$
l_{i} \succ r_{i},
$$

and that there are no $i, j \in I_{h}, i \neq j$, with $\left(l_{i}=l_{j}\right) \wedge\left(T\left(r_{i}\right)=T\left(r_{j}\right)\right)$. (Otherwise, since there is no $c \in \mathbb{Q}$ with $l_{i}-r_{i}=c \cdot\left(l_{j}-r_{j}\right)$, we know that $l_{i} \in I(\mathcal{B})$ and $r_{i} \in I(\mathcal{B})$, and we replace the two binomials in $\mathcal{B}$ by $l_{i}$ and $T\left(r_{i}\right)$.)

Let $u, v \in X_{0}^{*}$, and let $D$ be a derivation of length $n$ in $\mathcal{P}(\mathcal{B})$ leading from $u$ to $v$. Then there are terms $w_{i} \in X_{0}^{*}$ such that $D$ has the form $u=T\left(a_{1}\right) \cdot w_{1} \rightarrow T\left(b_{1}\right) \cdot w_{1}=$ $T\left(a_{2}\right) \cdot w_{2} \rightarrow T\left(b_{2}\right) \cdot w_{2} \rightarrow \cdots \rightarrow T\left(b_{n}\right) \cdot w_{n}=v$, where $a_{i}=l_{j_{i}}$ and $b_{i}=r_{j_{i}}$, or $a_{i}=r_{j_{i}}$ and $b_{i}=l_{j_{i}}, j_{i} \in I_{h}, i \in I_{n}$.
Attach to each $l_{i} \rightarrow T\left(r_{i}\right)(\mathcal{P}(\mathcal{B})), i \in I_{h}$, the multiplicative factor $C\left(r_{i}\right)$ if $r_{i} \neq 0$ (resp., 1 if $r_{i}=0$ ), and to each $T\left(r_{i}\right) \rightarrow l_{i}(\mathcal{P}(\mathcal{B})), i \in I_{h}$, the multiplicative factor $\frac{1}{C\left(r_{i}\right)}$ if $r_{i} \neq 0$ (resp., 1 if $r_{i}=0$ ). Taking these factors into account, we obtain from $D$ a derivation in which the $i$ th step has the form

$$
c \cdot l_{j_{i}} \cdot w_{i} \rightarrow c \cdot c_{i} \cdot T\left(r_{j_{i}}\right) \cdot w_{i}
$$

with $c_{i}=C\left(r_{j_{i}}\right)$ resp., $c_{i}=1$, or

$$
c \cdot T\left(r_{j_{i}}\right) \cdot w_{i} \rightarrow c \cdot c_{i} \cdot l_{j_{i}} \cdot w_{i}
$$

with $c_{i}=\frac{1}{C\left(r_{j_{i}}\right)}$ resp., $c_{i}=1$ for some constant $c \in \mathbb{Q} \backslash\{0\}$ resulting from the first $(i-1)$ steps of $D$.
Thus, we define the multiplicative factor $\mathcal{C}(D)$ of $D$ as

$$
\mathcal{C}(D)=c_{1} \cdot c_{2} \cdots c_{n} .
$$

Then, for any derivation $D$ in $\mathcal{P}(\mathcal{B})$ leading from $u$ to $v, u, v \in X_{0}^{*}$, we have

$$
\sum_{i=1}^{n} d_{i} \cdot\left(l_{j_{i}}-r_{j_{i}}\right) \cdot w_{i}=u-\mathcal{C}(D) \cdot v
$$

where $d_{1}=1$ if $u=l_{j_{1}} \cdot w_{1}$ resp., $d_{1}=-c_{1}$ if $u=T\left(r_{j_{1}}\right) \cdot w_{1}$, and, for $i>1, d_{i}=c_{1} \cdots c_{i-1}$ if the $i$ th step of $D$ uses $l_{i} \rightarrow T\left(r_{i}\right)(\mathcal{P}(\mathcal{B}))$ resp., $d_{i}=-c_{1} \cdots c_{i}$ if the $i$ th step of $D$ uses $T\left(r_{i}\right) \rightarrow l_{i}(\mathcal{P}(\mathcal{B}))$. Therefore, $u-\mathcal{C}(D) \cdot v \in I(\mathcal{B})$. Note that $u$ and $v$ are elements of $I(\mathcal{B})$ if $x_{1}^{-\infty} \cdots x_{k}^{-\infty}$ occurs in $D$.

By Propositions 2.1 and 2.2, we conclude the following:
Theorem 4.1. Let $X=\left\{x_{1}, \ldots, x_{k}\right\}, \mathcal{B}=\left\{l_{i}-r_{i} ; i \in I_{h}\right\}$ with $l_{i} \in X^{*}, r_{i} \in M[X]$ for all $i \in I_{h}$, and let $u, v$ be two monomials in $M[X] \backslash\{0\}$ with $T(u) \neq T(v)$. Then the following are equivalent.
(i) There exists $d \in \mathbb{Q} \backslash\{0\}$ such that $u-d \cdot v \in I(\mathcal{B})$.
(ii) There is a repetition-free derivation $D: T(u)=\gamma_{0} \rightarrow \gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n}=T(v)$ in $\mathcal{P}(\mathcal{B})$ leading from $T(u)$ to $T(v)$ such that, for $j \in\{0,1, \ldots, n\}$,

$$
\operatorname{size}\left(\gamma_{j}\right) \leq \operatorname{size}(u, v, \mathcal{B}) \cdot 2^{c \cdot k}
$$

where $c>0$ is some constant independent of $u$, $v$, and $\mathcal{B}$.
By Proposition 2.3, we have:
Corollary 4.1. Let $X=\left\{x_{1}, \ldots, x_{k}\right\}$, and $\mathcal{B}=\left\{l_{i}-r_{i} ; i \in I_{h}\right\}$ with $l_{i} \in X^{*}, r_{i} \in$ $M[X]$ for all $i \in I_{h}$. Then there is a (deterministic) Turing machine TM and some constant $c>0$ independent of $\mathcal{B}$ such that TM decides for any two monomials $u, v \in$ $M[X] \backslash\{0\}, T(u) \neq T(v)$, whether there exists $d \in \mathbb{Q} \backslash\{0\}$ such that $u-d \cdot v \in I(\mathcal{B})$, using at most space $(\operatorname{size}(u, v, \mathcal{B}))^{2} \cdot 2^{c \cdot k}$.

To obtain similar results concerning the membership of a single monomial in $I(\mathcal{B})$, we need a further detail.

Theorem 4.2. Let $X=\left\{x_{1}, \ldots, x_{k}\right\}, \mathcal{B}=\left\{l_{i}-r_{i} ; i \in I_{h}\right\}$ with $l_{i} \in X^{*}, r_{i} \in M[X]$ for all $i \in I_{h}$, and $u \neq 0$ a monomial in $M[X]$. Then the following are equivalent.
(i) $u \in I(\mathcal{B})$.
(ii) There is a derivation D: $T(u)=\gamma_{0} \rightarrow \gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n}$ of length $n$ in $\mathcal{P}(\mathcal{B})$ leading from $T(u)$ to $x_{1}^{-\infty} \cdots x_{k}^{-\infty}$, or from $T(u)$ to $T(u)$ with $\mathcal{C}(D) \neq 1$ such that, for $j \in\{0,1, \ldots, n\}$,

$$
\operatorname{size}\left(\gamma_{j}\right) \leq \operatorname{size}(u, \mathcal{B}) \cdot 2^{c_{1} \cdot k}
$$

and

$$
n \leq 2^{\operatorname{size}(u, \mathcal{B}) \cdot 2^{c_{2} \cdot k}}
$$

where $c_{1}, c_{2}>0$ are some constants independent of $u$ and $\mathcal{B}$.
Proof. By the above considerations, we already know that if there is a derivation $D$ in $\mathcal{P}(\mathcal{B})$ leading from $T(u)$ to $x_{1}^{-\infty} \cdots x_{k}^{-\infty}$, then $C(u) \cdot\left(T(u)-\mathcal{C}(D) \cdot x_{1}^{-\infty} \cdots x_{k}^{-\infty}\right)=u \in$ $I(\mathcal{B})$. Furthermore, we know that if there is a derivation $D$ in $\mathcal{P}(\mathcal{B})$ leading from $T(u)$ to $T(u)$ with $\mathcal{C}(D) \neq 1$, then $T(u)-\mathcal{C}(D) \cdot T(u)=(1-\mathcal{C}(D)) \cdot T(u) \in I(\mathcal{B})$, and thus, $u \in I(\mathcal{B})$. Hence, it suffices to show that (i) implies (ii).
W.l.o.g. we assume that $C(u)=1$. If $u \in I(\mathcal{B})$. Then, by Proposition 2.2, there exist $p_{1}^{\prime}, \ldots, p_{h}^{\prime} \in \mathbb{Q}[X]$ such that

$$
u=\sum_{i=1}^{h} p_{i}^{\prime}\left(l_{i}-r_{i}\right)
$$

and $\operatorname{size}\left(p_{i}^{\prime}\right) \leq \operatorname{size}(u, \mathcal{B}) \cdot 2^{c_{1} \cdot k}$, where $c_{1}>0$ is some constant independent of $u$ and $\mathcal{B}$. We may assume that $u=\sum_{j=1}^{m} p_{j}\left(l_{i_{j}}-r_{i_{j}}\right)$, for some $m \geq 1$, where $p_{j} \in M[X] \backslash\{0\}$, $\operatorname{deg}\left(p_{j}\right) \leq \operatorname{deg}\left(p_{i_{j}}^{\prime}\right), j \in I_{m}, i_{j} \in I_{h}$, and there are no $j_{1}, j_{2} \in I_{m}, j_{1} \neq j_{2}$, with $\left(T\left(p_{j_{1}}\right)=\right.$ $\left.T\left(p_{j_{2}}\right)\right) \wedge\left(l_{i_{j_{1}}}=l_{i_{j_{2}}}\right) \wedge\left(T\left(r_{i_{j_{1}}}\right)=T\left(r_{i_{j_{2}}}\right)\right)$. In the following, we construct from this polynomial identity a replacement tree from which we then extract a derivation $D$ either leading from $u$ to $x_{1}^{-\infty} \cdots x_{k}^{-\infty}$, or from $u$ to $u$ with $\mathcal{C}(D) \neq 1$. First, the notion of a replacement tree is defined.

A replacement tree w.r.t. $/ \mathcal{B}$ is a pair $(V, E)$, where $V$ is a subset of the set of terms $X_{0}^{*}$, and $E$ is a subset of the set of ordered 4 -tuples $(V \times V \times \mathbb{Q} \backslash\{0\} \times \mathbb{Q})$. The elements of $E$ are called arcs. An arc $\left(v_{1}, v_{2}, c, d\right) \in E$ is directed from the term $v_{1}$ to the term $v_{2}$. Its meaning is that $v_{2}$ is derived in one step from $v_{1}$ by application of a congruence in $\mathcal{P}(\mathcal{B})$. The third and fourth components $c, d$ of the arc are called the coefficients of the arc. The rational number $c \neq 0$ is the multiplicative factor of the production $l_{i} \rightarrow T\left(r_{i}\right)(\mathcal{P}(\mathcal{B}))$ resp., $T\left(r_{i}\right) \rightarrow l_{i}(\mathcal{P}(\mathcal{B})), i \in I_{h}$, by which $v_{2}$ is derived from $v_{1}$, and the rational number $d$ shows "how much $v_{2}$ " is derived from "how much $v_{1}$ ", i.e. $\frac{d}{c} \cdot v_{1} \rightarrow d \cdot v_{2}$.

The in-degree of a term $v \in V, \operatorname{deg}_{\text {in }}(v)$, is the number of arcs directed into $v$, and the out-degree $\operatorname{deg}_{\text {out }}(v)$ of $v$ is the number of arcs directed out of $v$. In a replacement tree, exactly one term in $V$ has in-degree zero. This term is the root of the replacement tree. The terms in $V$ with out-degree zero are called leaves.
A replacement tree is divided up into levels. A term of a replacement tree belongs to level $i, i \in \mathbb{N}$, if the length of the shortest derivation contained in the replacement tree leading from the root to that term is $i$. A replacement tree has the form shown in Figure 2. (The coefficients of the arcs do not appear in the picture.)

For each term $v \in V$ in the replacement tree which is not the root, the sum of the coefficients $d_{i}^{\text {in }}$ of the incoming $\operatorname{arcs}\left(., v, ., d_{i}^{\text {in }}\right) \in E, i \in I_{\operatorname{deg}_{\text {in }}(v)}$, equals the sum of the quotients $\frac{d_{j}^{\text {out }}}{c_{j}^{j u t}}$ of the coefficients $d_{j}^{\text {out }}, c_{j}^{\text {out }}$ of the outgoing $\operatorname{arcs}\left(v, ., c_{j}^{\text {out }}, d_{j}^{\text {out }}\right) \in E$, $j \in I_{\operatorname{deg}_{\text {out }}(v)}$, i.e.

$$
\begin{equation*}
d_{1}^{\text {in }}+\cdots+d_{\operatorname{deg}_{\text {in }}(v)}^{\text {in }}=\frac{d_{1}^{\text {out }}}{c_{1}^{\text {out }}}+\cdots+\frac{d_{\operatorname{deg}_{\text {out }}(v)}^{\text {out }}}{c_{\operatorname{deg}_{\text {out }}(v)}^{\text {out }}} . \tag{4.1}
\end{equation*}
$$

Note that the leaves in a replacement tree have out-degree zero and thus, for leaves, the right-hand side of this equation is zero. The quotients $\frac{d_{j}^{\text {out }}}{c_{j}^{\text {out }}}$ of the coefficients $d_{j}^{\text {out }}, c_{j}^{\text {out }}$ of the $\operatorname{arcs}\left(u, ., c_{j}^{\text {out }}, d_{j}^{\text {out }}\right) \in E, j \in I_{\operatorname{deg}_{\text {out }}(u)}$, directed out of the root $u$ satisfy

$$
\begin{equation*}
1=\frac{d_{1}^{\text {out }}}{c_{1}^{\text {out }}}+\cdots+\frac{d_{\mathrm{deg}_{\text {out }}(u)}^{\text {out }}}{c_{\mathrm{deg}_{\text {out }}(u)}^{\text {out }}} \tag{4.2}
\end{equation*}
$$

The root of the replacement tree $(V, E)$ to be constructed from the polynomial identity

$$
\begin{equation*}
u=\sum_{j=1}^{m} p_{j}\left(l_{i_{j}}-r_{i_{j}}\right) \tag{4.3}
\end{equation*}
$$

is the term $u$. We start with $V=\{u\}$ and $E=\emptyset$. As $u$ appears as a term on the left-hand side of (4.3), the sum of the monomials $p_{j} a_{j}, j \in I_{m}$, on the right-hand side of (4.3) with $T\left(p_{j} a_{j}\right)=u, a_{j}=l_{i_{j}}$, or $a_{j}=-r_{i_{j}}$ yields $u$, i.e. for $J_{u}=\left\{j \in I_{m} ; T\left(p_{j} a_{j}\right)=u, a_{j}=\right.$ $l_{i_{j}}$, or $\left.a_{j}=-r_{i_{j}}\right\}$, we have

$$
\sum_{j \in J_{u}} p_{j} a_{j}=u
$$

implying

$$
\sum_{j \in J_{u}} p_{j} b_{j}=\sum_{j \in I_{m} \backslash J_{u}} p_{j}\left(l_{i_{j}}-r_{i_{j}}\right),
$$

where $b_{j}=r_{i_{j}}$ if $a_{j}=l_{i_{j}}$ resp., $b_{j}=-l_{j_{i}}$ if $a_{j}=-r_{i_{j}}$.
This elimination of all the monomials in (4.3) with power product part $u$ can be inter-
level 0
level 1
level 2
level 3


level $l-2$

level $l-1$
level $l$




Figure 2. Replacement tree.
preted as one-step derivations $C\left(p_{j} a_{j}\right) \cdot u \rightarrow C\left(p_{j} b_{j}\right) \cdot T\left(p_{j} b_{j}\right), j \in J_{u}$. Add $\left\{T\left(p_{j} b_{j}\right) ; j \in\right.$ $\left.J_{u}\right\}$ to $V$. Then these one-step derivations correspond to the arcs

$$
\left(u, T\left(p_{j} b_{j}\right), c_{j}, C\left(p_{j} b_{j}\right)\right), j \in J_{u}
$$

where the $c_{j}$ are the multiplicative factors of the productions $T\left(a_{j}\right) \rightarrow T\left(b_{j}\right)(\mathcal{P}(\mathcal{B}))$, i.e. $V:=V \cup\left\{T\left(p_{j} b_{j}\right) ; j \in J_{u}\right\}, E:=E \cup\left\{\left(u, T\left(p_{j} b_{j}\right), c_{j}, C\left(p_{j} b_{j}\right)\right) ; j \in J_{u}\right\}$.
The polynomial identity (4.3) can now be written as

$$
\begin{equation*}
\sum_{i=1}^{t} e_{i} v_{i}=\sum_{j \in I_{m} \backslash J_{u}} p_{j}\left(l_{i_{j}}-r_{i_{j}}\right) \tag{4.4}
\end{equation*}
$$

with $v_{i} \in V \backslash\left\{x_{1}^{-\infty} \cdots x_{k}^{-\infty}\right\}$ (if $x_{1}^{-\infty} \cdots x_{k}^{-\infty} \in V$, then we remember that $x_{1}^{-\infty} \cdots x_{k}^{-\infty}$
corresponds to 0 ) and $e_{i} \in \mathbb{Q} \backslash\{0\}, i \in I_{t}$. The terms in $\left\{v_{1}, \ldots, v_{t}\right\}$ are assumed to be pairwise different, and, for each $i \in I_{t}, e_{i}$ is the resulting coefficient when summing up all coefficients $C\left(p_{j} b_{j}\right), j \in J_{u}$, with $T\left(p_{j} b_{j}\right)=v_{i}$, i.e.

$$
e_{i}=\sum_{j \in J_{u} ; T\left(p_{j} b_{j}\right)=v_{i}} C\left(p_{j} b_{j}\right) .
$$

The next step in the construction of the replacement tree is to repeat the described procedure for $e_{1} v_{1}$ on the left-hand side of (4.4). In general, such an elimination step works as follows. At the beginning, we have a polynomial identity

$$
\begin{equation*}
\sum_{i=1}^{t} e_{i} v_{i}=\sum_{j \in I_{m} \backslash J_{\mathrm{el}}} p_{j}\left(l_{i_{j}}-r_{i_{j}}\right) \tag{4.5}
\end{equation*}
$$

where, for $i \in I_{t}, v_{i} \in V \backslash\left\{x_{1}^{-\infty} \cdots x_{k}^{-\infty}\right\}, v_{i} \neq v_{i^{\prime}}$ for all $i^{\prime} \in I_{t} \backslash\{i\}, e_{i} \in \mathbb{Q} \backslash\{0\}$, and $J_{\mathrm{el}} \subset I_{m}$ contains the indices of already eliminated monomials. Choose a term $v_{l}$, $l \in I_{t}$, which, for instance, belongs to the lowest level among all $v_{i}, i \in I_{t}$. The monomial $e_{l} v_{l}$ on the left-hand side of (4.5) equals the sum of the monomials $p_{j} a_{j}, j \in I_{m} \backslash J_{\mathrm{el}}$, on the right-hand side of (4.5) with $T\left(p_{j} a_{j}\right)=v_{l}, a_{j}=l_{i_{j}}$, or $a_{j}=-r_{i_{j}}$, i.e. for $J_{v_{l}}=\left\{j \in I_{m} \backslash J_{\mathrm{el}} ; T\left(p_{j} a_{j}\right)=v_{l}, a_{j}=l_{i_{j}}\right.$, or $\left.a_{j}=-r_{i_{j}}\right\}$, we have

$$
\sum_{j \in J_{v_{l}}} p_{j} a_{j}=e_{l} v_{l}
$$

which implies

$$
\sum_{j \in J_{v_{l}}} p_{j} b_{j}+\sum_{i \in I_{t} \backslash\{l\}} e_{i} v_{i}=\sum_{j \in\left(I_{m} \backslash J_{\mathrm{el}}\right) \backslash J_{v_{l}}} p_{j}\left(l_{i_{j}}-r_{i_{j}}\right),
$$

where $b_{j}=r_{i_{j}}$ if $a_{j}=l_{i_{j}}$ resp., $b_{j}=-l_{j_{i}}$ if $a_{j}=-r_{i_{j}}$.
Let $V:=V \cup\left\{T\left(p_{j} b_{j}\right) ; j \in J_{v_{l}}\right\}$, and $E:=E \cup\left\{\left(v_{l}, T\left(p_{j} b_{j}\right), c_{j}, C\left(p_{j} b_{j}\right)\right) ; j \in J_{v_{l}}\right\}$, where $c_{j}$ is the multiplicative factor of $T\left(a_{j}\right) \rightarrow T\left(b_{j}\right)(\mathcal{P}(\mathcal{B}))$. From (4.5) we obtain as a new polynomial identity

$$
\sum_{i=1}^{\bar{t}} \bar{e}_{i} \bar{v}_{i}=\sum_{j \in\left(I_{m} \backslash J_{\text {el }}\right) \backslash J_{v_{l}}} p_{j}\left(l_{i_{j}}-r_{i_{j}}\right),
$$

where, for $i \in I_{\bar{t}}, \bar{v}_{i} \in V \backslash\left\{x_{1}^{-\infty} \cdots x_{k}^{-\infty}\right\}, \bar{v}_{i} \neq \bar{v}_{i^{\prime}}$ for all $i^{\prime} \in I_{\bar{t}} \backslash\{i\}$, and, if in (4.5) there is some $i_{1} \in I_{t}$ with $v_{i_{1}}=\bar{v}_{i}$, then

$$
\bar{e}_{i}=e_{i_{1}}+\sum_{j \in J_{v_{l}} ; T\left(p_{j} b_{j}\right)=\bar{v}_{i}} C\left(p_{j} b_{j}\right),
$$

else

$$
\bar{e}_{i}=\sum_{j \in J_{v_{l}} ; T\left(p_{j} b_{j}\right)=\bar{v}_{i}} C\left(p_{j} b_{j}\right)
$$

The construction is finished if $\bar{e}_{i}=0$ for all $i \in I_{\bar{t}}$. Then the pair $(V, E)$ is a replacement tree, because, by construction, for all $v \in V \backslash\{u\}$, the coefficients of the incoming and outgoing arcs satisfy equation (4.1), and the coefficients of the arcs directed out of the root $u$ satisfy equation (4.2).

If in the replacement tree $x_{1}^{-\infty} \cdots x_{k}^{-\infty} \in V$, then there is a derivation $u=\gamma_{0} \rightarrow$


Figure 3. $D_{1}, D_{2}$ : two disjoint repetition-free derivations leading from $a$ to $b$.
$\gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n}=x_{1}^{-\infty} \cdots x_{k}^{-\infty}$ in $\mathcal{P}(\mathcal{B})$ such that, for $j \in\{0,1, \ldots, n\}, \operatorname{size}\left(\gamma_{j}\right) \leq$ $\operatorname{size}(u, \mathcal{B}) \cdot 2^{c_{1} \cdot k}$, and $n \leq 2^{\operatorname{size}(u, \mathcal{B}) \cdot 2^{c_{2} \cdot k}}$, where $c_{1}, c_{2}>0$ are some constants independent of $u$ and $\mathcal{B}$, and there is nothing left to prove.

In the following, we assume that $x_{1}^{-\infty} \cdots x_{k}^{-\infty} \notin V$. We show how to extract from the constructed replacement tree $(V, E)$ a derivation $D$ in $\mathcal{P}(\mathcal{B})$ leading from $u$ to $u$ with $\mathcal{C}(D) \neq 1$.

Since the leaves of a replacement tree have out-degree zero, the coefficients $d_{i}$ of the $\operatorname{arcs}\left(., b, ., d_{i}\right) \in E, i \in I_{\operatorname{deg}_{\text {in }}(b)}$, directed into a leaf $b \in V$ satisfy $d_{1}+\cdots+d_{\operatorname{deg}_{\text {in }}(b)}=0$. Because $b \neq x_{1}^{-\infty} \cdots x_{k}^{-\infty}$, it follows from the construction of the replacement tree $(V, E)$ that, for all $i \in I_{\operatorname{deg}_{\text {in }}(b)}, d_{i} \neq 0$, and thus $\operatorname{deg}_{\text {in }}(b) \geq 2$.

Take an arbitrary leaf $b \in V$, and select in the replacement tree ( $V, E$ ) two repetitionfree derivations

$$
D_{1}: a=\gamma_{0}^{1} \rightarrow \gamma_{1}^{1} \rightarrow \cdots \rightarrow \gamma_{n_{1}}^{1}=b
$$

and

$$
D_{2}: a=\gamma_{0}^{2} \rightarrow \gamma_{1}^{2} \rightarrow \cdots \rightarrow \gamma_{n_{2}}^{2}=b
$$

leading from some term $a \in V$ to $b$ with $\gamma_{i}^{1} \notin\left\{\gamma_{1}^{2}, \ldots, \gamma_{n_{2}-1}^{2}\right\}, i \in I_{n_{1}-1}$, and $\gamma_{j}^{2} \notin$ $\left\{\gamma_{1}^{1}, \ldots, \gamma_{n_{1}-1}^{1}\right\}, j \in I_{n_{2}-1}$. Let $\left(\gamma_{i}^{1}, \gamma_{i+1}^{1}, c_{i+1}^{1}, d_{i+1}^{1}\right), i \in\left\{0, \ldots, n_{1}-1\right\}$ resp., $\left(\gamma_{j}^{2}, \gamma_{j+1}^{2}\right.$, $\left.c_{j+1}^{2}, d_{j+1}^{2}\right), j \in\left\{0, \ldots, n_{2}-1\right\}$ denote the corresponding arcs in $E$ (see Figure 3). Then the multiplicative factors of the derivations $D_{1}, D_{2}$ are $\mathcal{C}\left(D_{1}\right)=c_{1}^{1} \cdots c_{n_{1}}^{1}$ and $\mathcal{C}\left(D_{2}\right)=$ $c_{1}^{2} \cdots c_{n_{2}}^{2}$.

If $\mathcal{C}\left(D_{1}\right) \neq \mathcal{C}\left(D_{2}\right)$, we are finished because, by reversing the direction of each step in some derivation $D: v_{1}=\gamma_{0} \rightarrow \gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n}=v_{2}$ in $\mathcal{P}(\mathcal{B})$, we obtain the reverse derivation $D_{\mathrm{r}}: v_{2}=\gamma_{n} \rightarrow \gamma_{n-1} \rightarrow \cdots \rightarrow \gamma_{0}=v_{1}(\mathcal{P}(\mathcal{B}))$ with $\mathcal{C}\left(D_{\mathrm{r}}\right)=\frac{1}{\mathcal{C}(D)}$. Furthermore, from $(V, E)$ we obtain a derivation $D_{u}: u \xrightarrow{*} a(\mathcal{P}(\mathcal{B}))$. Thus, we have a derivation

$$
D: u \xrightarrow{*} a \xrightarrow{+} b \xrightarrow{+} a \xrightarrow{*} u
$$

in $\mathcal{P}(\mathcal{B})$ with

$$
\mathcal{C}(D)=\mathcal{C}\left(D_{u}\right) \cdot \mathcal{C}\left(D_{1}\right) \cdot \frac{1}{\mathcal{C}\left(D_{2}\right)} \cdot \frac{1}{\mathcal{C}\left(D_{u}\right)} \neq 1
$$

In the case $\mathcal{C}\left(D_{1}\right)=\mathcal{C}\left(D_{2}\right)$, we eliminate the arc

$$
\left(\gamma_{n_{1}-1}^{1}, b, c_{n_{1}}^{1}, d_{n_{1}}^{1}\right)
$$

from the replacement tree $(V, E)$. Since $\mathcal{C}\left(D_{1}\right)=\mathcal{C}\left(D_{2}\right), d_{n_{1}}^{1} b$ can be derived from $\frac{d_{n_{1}}^{1}}{\mathcal{C}\left(D_{1}\right)} a=\frac{d_{n_{1}}^{1}}{\mathcal{C}\left(D_{2}\right)} a$ not only by derivation $D_{1}$, but also by derivation $D_{2}$. The goal is to derive $\left(d_{n_{1}}^{1}+d_{n_{2}}^{2}\right) b$ from $\frac{d_{n_{1}}^{1}+d_{n_{2}}^{2}}{\mathcal{C}\left(D_{2}\right)} a$ by $D_{2}$, and to derive no $b$ from no $a$ by $D_{1}$.

To obtain this result, we replace in $E$

$$
\begin{aligned}
& \left(\gamma_{i-1}^{1}, \gamma_{i}^{1}, c_{i}^{1}, d_{i}^{1}\right) \text { by }\left(\gamma_{i-1}^{1}, \gamma_{i}^{1}, c_{i}^{1}, d_{i}^{1}-\frac{d_{n_{1}}^{1}}{c_{i+1}^{1} \cdots c_{n_{1}}^{1}}\right), \text { for each } i \in\left\{1, \ldots, n_{1}-1\right\}, \\
& \left(\gamma_{j-1}^{2}, \gamma_{j}^{2}, c_{j}^{2}, d_{j}^{2}\right) \text { by }\left(\gamma_{j-1}^{2}, \gamma_{j}^{2}, c_{j}^{2}, d_{j}^{2}+\frac{d_{n_{1}}^{1}}{c_{j+1}^{2} \cdots c_{n_{2}}^{2}}\right), \text { for each } j \in\left\{1, \ldots, n_{2}-1\right\}, \\
& \left(\gamma_{n_{2}-1}^{2}, \gamma_{n_{2}}^{2}, c_{n_{2}}^{2}, d_{n_{2}}^{2}\right) \text { by }\left(\gamma_{n_{2}-1}^{2}, \gamma_{n_{2}}^{2}, c_{n_{2}}^{2}, d_{n_{1}}^{1}+d_{n_{2}}^{2}\right), \text { and we remove }\left(\gamma_{n_{1}-1}^{1}, \gamma_{n_{1}}^{1}, c_{n_{1}}^{1},\right.
\end{aligned}
$$

Since

$$
\begin{aligned}
\frac{d_{1}^{1}-\frac{d_{n_{1}}^{1}}{c_{2}^{1} \cdots c_{n_{1}}^{1}}}{c_{1}^{1}}+\frac{d_{1}^{2}+\frac{d_{n_{1}}^{1}}{c_{2}^{2} \cdots c_{n_{2}}^{2}}}{c_{1}^{2}} & =\frac{d_{1}^{1}}{c_{1}^{1}}-\frac{d_{n_{1}}^{1}}{\mathcal{C}\left(D_{1}\right)}+\frac{d_{1}^{2}}{c_{1}^{2}}+\frac{d_{n_{1}}^{1}}{\mathcal{C}\left(D_{2}\right)}=\frac{d_{1}^{1}}{c_{1}^{1}}+\frac{d_{1}^{2}}{c_{1}^{2}} \\
d_{i}^{1}-\left(d_{i}^{1}-\frac{d_{n_{1}}^{1}}{c_{i+1}^{1} \cdots c_{n_{1}}^{1}}\right) & =\frac{d_{n_{1}}^{1}}{c_{i+1}^{1} \cdots c_{n_{1}}^{1}}=\frac{d_{i+1}^{1}}{c_{i+1}^{1}}-\frac{d_{i+1}^{1}-\frac{d_{n_{1}}^{1}}{c_{i+2}^{1} \cdots c_{n_{1}}^{1}}}{c_{i+1}^{1}}, i \in I_{n_{1}-2}, \\
d_{n_{1}-1}^{1}-\left(d_{n_{1}-1}^{1}-\frac{d_{n_{1}}^{1}}{c_{n_{1}}^{1}}\right) & =\frac{d_{n_{1}}^{1}}{c_{n_{1}}^{1}}=\frac{d_{n_{1}}^{1}}{c_{n_{1}}^{1}}-0, \\
d_{j}^{2}-\left(d_{j}^{2}+\frac{d_{n_{1}}^{1}}{c_{j+1}^{2} \cdots c_{n_{2}}^{2}}\right) & =-\frac{d_{n_{1}}^{1}}{c_{j+1}^{2} \cdots c_{n_{2}}^{2}}=\frac{d_{j+1}^{2}}{c_{j+1}^{2}}-\frac{d_{j+1}^{2}+\frac{d_{n_{1}}^{1}}{c_{j+2}^{2} \cdots c_{n_{2}}^{2}}}{c_{j+1}^{2}}, j \in I_{n_{2}-2}, \\
d_{n_{2}-1}^{2}-\left(d_{n_{2}-1}^{2}+\frac{d_{n_{1}}^{1}}{c_{n_{2}}^{2}}\right) & =-\frac{d_{n_{1}}^{1}}{c_{n_{2}}^{2}}=\frac{d_{n_{2}}^{2}}{c_{n_{2}}^{2}}-\frac{d_{n_{1}}^{1}+d_{n_{2}}^{2}}{c_{n_{2}}^{2}},
\end{aligned}
$$

for each $v \in V$, the coefficients of the incoming and outgoing arcs still satisfy equation (4.1) (resp., equation (4.2)). Also, a subsequent removal of all new $\operatorname{arcs}\left(v_{1}, v_{2}, c, d\right) \in$ $E$ with $d=0$ from $E$ does not change this fact. Hence, after removing all terms $v \in V \backslash\{u\}$ with $\operatorname{deg}_{\text {in }}(v)=0\left(=\operatorname{deg}_{\text {out }}(v)\right)$ from $V$, the pair $(V, E)$ is still a replacement tree.

In the polynomial identity (4.3), the procedure just described corresponds to a reduction of the number $m$ of products $p_{j}\left(l_{i_{j}}-r_{i_{j}}\right)$ in the sum on the right-hand side of (4.3) by at least one. Each arc in the replacement tree constructed from the polynomial identity (4.3) corresponds to such a product $p_{j}\left(l_{i_{j}}-r_{i_{j}}\right), i \in I_{m}$. Thus, the above elimination of an arc in the replacement tree corresponds to an elimination of some $p_{j}\left(l_{i_{j}}-r_{i_{j}}\right)$ in (4.3). The resulting modifications of the coefficients of the arcs in the two derivations correspond to appropriate modifications of the coefficients $C\left(p_{j}\right)$ in the respective products $p_{j}\left(l_{i_{j}}-r_{i_{j}}\right)$.

We may repeat the above argument for any leaf in $V$, and by induction obtain a derivation $D: u \xrightarrow{+} u$ with $\mathcal{C}(D) \neq 1$. If, during the induction, no such derivation is found until the replacement tree $(V, E)$ only consists of the root $u$, one leaf $b$, and two disjoint repetition-free derivations $D_{1}$ and $D_{2}$ leading from $u$ to $b$, then, at least now, we have $\mathcal{C}\left(D_{1}\right) \neq \mathcal{C}\left(D_{2}\right)$. Let $D_{1}, D_{2}$, and the corresponding arcs in $E$ be as above. Assume that $\mathcal{C}\left(D_{1}\right)=\mathcal{C}\left(D_{2}\right)$, then from equation (4.1), we obtain

$$
\begin{aligned}
& d_{n_{1}}^{1}=-d_{n_{2}}^{2} \\
& d_{n_{1}-1}^{1}=\frac{d_{n_{1}}^{1}}{c_{n_{1}}^{1}}, \quad d_{n_{1}-2}^{1}=\frac{d_{n_{1}-1}^{1}}{c_{n_{1}-1}^{1}}=\frac{d_{n_{1}}^{1}}{c_{n_{1}-1}^{1} \cdot c_{n_{1}}^{1}}, \quad \ldots, \quad d_{1}^{1}=\frac{d_{n_{1}}^{1}}{c_{2}^{1} \cdots c_{n_{1}}^{1}}
\end{aligned}
$$

and

$$
d_{n_{2}-1}^{2}=\frac{d_{n_{2}}^{2}}{c_{n_{2}}^{2}}, \quad d_{n_{2}-2}^{2}=\frac{d_{n_{2}-1}^{2}}{c_{n_{2}-1}^{2}}=\frac{d_{n_{2}}^{2}}{c_{n_{2}-1}^{2} \cdot c_{n_{2}}^{2}}, \quad \ldots, \quad d_{1}^{2}=\frac{d_{n_{2}}^{2}}{c_{2}^{2} \cdots c_{n_{2}}^{2}}
$$

From equation (4.2), concerning the root we obtain

$$
1=\frac{d_{1}^{1}}{c_{1}^{1}}+\frac{d_{1}^{2}}{c_{1}^{2}}=\frac{d_{n_{1}}^{1}}{c_{1}^{1} \cdots c_{n_{1}}^{1}}+\frac{d_{n_{2}}^{2}}{c_{1}^{2} \cdots c_{n_{2}}^{2}}=\frac{d_{n_{1}}^{1}}{\mathcal{C}\left(D_{1}\right)}+\frac{d_{n_{2}}^{2}}{\mathcal{C}\left(D_{2}\right)}=\frac{d_{n_{1}}^{1}}{\mathcal{C}\left(D_{1}\right)}-\frac{d_{n_{1}}^{1}}{\mathcal{C}\left(D_{2}\right)}
$$

which contradicts the assumption. Thus, there is a derivation $D: u=\gamma_{0} \rightarrow \gamma_{1} \rightarrow$ $\cdots \rightarrow \gamma_{n}=u$ of length $n$ in $\mathcal{P}(\mathcal{B})$ with $\mathcal{C}(D) \neq 1$ such that, for $j \in\{0,1, \ldots, n\}$, $\operatorname{size}\left(\gamma_{j}\right) \leq \operatorname{size}(u, \mathcal{B}) \cdot 2^{c_{1} \cdot k}$, and $n \leq 2^{\operatorname{size}(u, \mathcal{B}) \cdot 2^{c_{2} \cdot k}}$, where $c_{1}, c_{2}>0$ are some constants independent of $u$ and $\mathcal{B}$.

Furthermore, we can show the following:
Theorem 4.3. Let $X=\left\{x_{1}, \ldots, x_{k}\right\}$, and $\mathcal{B}=\left\{l_{i}-r_{i} ; i \in I_{h}\right\}$ with $l_{i} \in X^{*}, r_{i} \in M[X]$ for all $i \in I_{h}$. Then there is a (deterministic) Turing machine TM and some constant $c>0$ independent of $\mathcal{B}$ such that TM decides, for any monomial $u \neq 0$ in $M[X]$, whether $u \in I(\mathcal{B})$ uses at most space $(\operatorname{size}(u, \mathcal{B}))^{2} \cdot 2^{c \cdot k}$.

Proof. By Theorem 4.2, a non-deterministic Turing machine can determine whether $u \in I(\mathcal{B})$ by generating a derivation $D: T(u)=\gamma_{0} \rightarrow \gamma_{1} \rightarrow \cdots \rightarrow \gamma_{n}$ of length $n$ (with $n$ doubly exponentially bounded in the size of the problem instance) in $\mathcal{P}(\mathcal{B})$ leading either from $T(u)$ to $x_{1}^{-\infty} \cdots x_{k}^{-\infty}$, or from $T(u)$ to $T(u)$ with $\mathcal{C}(D) \neq 1$ iff there is such a derivation. If $x_{1}^{-\infty} \cdots x_{k}^{-\infty} \in[T(u)]_{\mathcal{P}(\mathcal{B})}$, then $u \in I(\mathcal{B})$, and, by Proposition 2.3, we are done. In the other case, we may assume w.l.o.g. that the derivation $D$ uses no congruence whose right-hand side is $x_{1}^{-\infty} \cdots x_{k}^{-\infty}$ (note that $l_{i} \succ r_{i}$ for all $i \in I_{h}$ ). The Turing machine guesses $n$ and $2 h$ counters $z_{1}, \ldots, z_{2 h}$-two for each congruence $l_{i} \equiv T\left(r_{i}\right)$ in $\mathcal{P}(\mathcal{B})$-representing how often, and in which direction (i.e. $l_{i} \rightarrow T\left(r_{i}\right)$,
or $T\left(r_{i}\right) \rightarrow l_{i}$ ) each of the congruences is applied in $D$. These counters have to satisfy $z_{1}+z_{2}+\cdots+z_{2 h}=n$, and we obtain

$$
\mathcal{C}(D)=\prod_{i \in I_{h} ; r_{i} \neq 0}\left(\frac{a_{i}}{b_{i}}\right)^{z_{2 i-1}} \cdot\left(\frac{b_{i}}{a_{i}}\right)^{z_{2 i}},
$$

with $a_{i} \in \mathbb{Z} \backslash\{0\}$ the numerator, and $b_{i} \in \mathbb{N} \backslash\{0\}$ the denominator of $C\left(r_{i}\right), i \in I_{h}$, $r_{i} \neq 0$.
Let

$$
Z=\prod_{i \in I_{h} ; r_{i} \neq 0} a_{i}^{z_{2 i-1}} \cdot b_{i}^{z_{2 i}}
$$

and

$$
N=\prod_{i \in I_{h} ; r_{i} \neq 0} b_{i}^{z_{2 i-1}} \cdot a_{i}^{z_{2 i}}
$$

then $\max \{|Z|,|N|\} \leq\left(2^{\operatorname{size}(u, \mathcal{B})}\right)^{2^{\operatorname{size}(u, \mathcal{B}) \cdot 2^{d_{1} \cdot k}}}$ for some constant $d_{1}>0$ independent of $u$ and $\mathcal{B}$. By the Chinese remainder theorem and the prime number theorem (see, e.g. Hardy and Wright, 1985), we know

$$
\begin{aligned}
\mathcal{C}(D)=1 & \Longleftrightarrow Z \\
& \Longleftrightarrow Z \\
& \equiv N \bmod p_{j} \quad \text { for all } 1 \leq j \leq m
\end{aligned}
$$

where $p_{j}, j \in I_{m}$, are the prime numbers satisfying $2 \leq p_{j} \leq d_{2} \cdot \log M$ for any integer $M>2 \cdot \max \{|Z|,|N|\}$, with $d_{2}>0$ some constant independent of $u$ and $\mathcal{B}$. Thus, the products $Z$ and $N$ only have to be computed modulo the prime numbers $p_{j}, j \in I_{m}$, and the decision whether $Z=N$ uses at most space $\operatorname{size}(u, \mathcal{B}) \cdot 2^{d \cdot k}$, with $d>0$ some constant independent of $u$ and $\mathcal{B}$.

The non-deterministic Turing machine can verify that $\mathcal{C}(D) \neq 1$ by guessing a prime $p_{j}$ with $j \in I_{m}$ and computing $Z$ and $N$ modulo this prime. A deterministic Turing machine has to loop through the primes $p_{j}, j \in I_{m}$.

For generating the derivation $D$ in $\mathcal{P}(\mathcal{B})$, the non-deterministic Turing machine has to keep in storage at any time two consecutive words $\gamma_{i-1}$ and $\gamma_{i}$ of $D$ in order to check whether $\gamma_{i-1} \rightarrow \gamma_{i}(\mathcal{P}(\mathcal{B}))$. Therefore, by Theorem 4.2 and the above considerations, there is some constant $\bar{c}>0$ independent of $u$ and $\mathcal{B}$ such that the non-deterministic Turing machine needs at most $\operatorname{size}(u, \mathcal{B}) \cdot 2^{\bar{c} \cdot k}$ tape cells to determine whether $u \in I(\mathcal{B})$.
When simulating the non-deterministic Turing machine by a deterministic one, the standard construction of Savitch (1969) has to be slightly modified, halving the length of the derivation being looked for at every level of the recursion and also guessing (by looping through all possibilities) appropriate values for the tuples of counters.

The deterministic Turing machine calls a recursive Boolean function

$$
\operatorname{reachable}\left(\gamma_{1}, \gamma_{2},\left(z_{1}, \ldots, z_{2 h}\right)\right)
$$

which returns the Boolean value true if there exists a derivation from $\gamma_{1}$ to $\gamma_{2}$ in $\mathcal{P}(\mathcal{B})$ consisting of at most $z_{1}+z_{2}+\cdots+z_{2 h}$ steps, and applying $l_{i} \rightarrow T\left(r_{i}\right)(\mathcal{P}(\mathcal{B}))$ resp., $T\left(r_{i}\right) \rightarrow l_{i}(\mathcal{P}(\mathcal{B})) z_{2 i-1}$ resp., $z_{2 i}$ times, $i \in I_{h}$. The function reachable works by looking for the word $\gamma$ in the middle of the derivation from $\gamma_{1}$ to $\gamma_{2}$, and checking recursively that it is indeed the middle word. For each call we must store the current values of $\gamma, \gamma_{1}$, and $\gamma_{2}$, and the current values of the counters $z_{1}, \ldots, z_{2 h}$. These counters always have to add up to the length of the subderivation, and this length is halved at every level
of the recursion. Thus, the depth of the recursion is the logarithm of the initial value $n$ of $z_{1}+z_{2}+\cdots+z_{2 h}$, and, by Theorem 4.2, there are at most $\operatorname{size}(u, \mathcal{B}) \cdot 2^{c_{1} \cdot k}$ many levels of recursion, each requiring at most $\operatorname{size}(u, \mathcal{B}) \cdot 2^{c_{2} \cdot k}$ space, where $c_{1}, c_{2}>0$ are some constants independent of $u$ and $\mathcal{B}$. Hence, $(\operatorname{size}(u, \mathcal{B}))^{2} \cdot 2^{c \cdot k}$ space suffices for a deterministic Turing machine to decide whether $u \in I(\mathcal{B})$.

### 4.2. THE ALGORITHM

Together with the results of Section 3.2, we are now able to derive an exponential space algorithm for generating the reduced Gröbner basis of the binomial ideal $I(\mathcal{B})$ w.r.t. some admissible term ordering $\succeq$, where $X=\left\{x_{1}, \ldots, x_{k}\right\}$ and $\mathcal{B}=\left\{l_{i}-r_{i} ; i \in I_{h}\right\}$ with $l_{i} \in X^{*}, r_{i} \in M[X]$, and w.l.o.g. $l_{i} \succ r_{i}$ for all $i \in I_{h}$. As in Section 3.1, we first analyze the elements of the reduced Gröbner basis of a binomial ideal. Note that $t \in I(\mathcal{B})$ for all $t \in\left[x_{1}^{-\infty} \cdots x_{k}^{-\infty}\right]_{\mathcal{P}(\mathcal{B})}$.

Lemma 4.1. Let $X=\left\{x_{1}, \ldots, x_{k}\right\}, \mathcal{B}=\left\{l_{i}-r_{i} ; i \in I_{h}\right\}$ with $l_{i} \in X^{*}, r_{i} \in M[X]$ for all $i \in I_{h}$, and let $G=\left\{h_{1}-m_{1}, \ldots, h_{r}-m_{r}\right\}$ be the reduced Gröbner basis of the ideal $I(\mathcal{B})$ w.r.t. some admissible term ordering $\succeq\left(h_{i} \succ m_{i}, C\left(h_{i}\right)=1\right.$ for all $\left.i \in I_{r}\right)$. Then $T\left(m_{i}\right)$ is the minimal element (w.r.t. $\succ$ ) of the congruence class $\left[h_{i}\right]_{\mathcal{P}(\mathcal{B})}, i \in I_{r}$.

Proof. With Theorems 4.1 and 4.2 , this proof follows immediately from the proof of Proposition 3.1.

Lemma 4.2. Let $X=\left\{x_{1}, \ldots, x_{k}\right\}, \mathcal{B}=\left\{l_{i}-r_{i} ; i \in I_{h}\right\}$ with $l_{i} \in X^{*}, r_{i} \in M[X]$ for all $i \in I_{h}$, and let $G=\left\{h_{1}-m_{1}, \ldots, h_{r}-m_{r}\right\}$ be the reduced Gröbner basis of the ideal $I(\mathcal{B})$ w.r.t. some admissible term ordering $\succeq\left(h_{i} \succ m_{i}, C\left(h_{i}\right)=1\right.$ for all $\left.i \in I_{r}\right)$. Then $\operatorname{LT}(I(\mathcal{B}))$ (the set of the leading terms of $I(\mathcal{B})$ ) is the set of all terms $t \neq 0$ with either $t \in I(\mathcal{B})$, or, if $t \notin I(\mathcal{B})$, with non-trivial congruence class in $\mathcal{P}(\mathcal{B})$ such that $t$ is not the minimal (w.r.t. $\succ$ ) element $m_{t}$ of its congruence class (note: if $t \notin I(\mathcal{B})$, then $\left.m_{t} \neq x_{1}^{-\infty} \cdots x_{k}^{-\infty}\right) . H=\left\{h_{1}, \ldots, h_{r}\right\}$ is the set of the minimal elements of $L T(I(\mathcal{B}))$ w.r.t. divisibility.

Proof. With Theorems 4.1 and 4.2, this proof follows immediately from the proof of Proposition 3.1.

For any two terms $t_{1}, t_{2} \in X_{0}^{*}, t_{1} \neq t_{2}$, with $t_{1} \equiv t_{2} \bmod \mathcal{P}(\mathcal{B})$, it follows that $t_{1}-\mathcal{C}(D) \cdot t_{2} \in I(\mathcal{B})$, where $D$ is a derivation from $t_{1}$ to $t_{2}$ in $\mathcal{P}(\mathcal{B})$. By definition,

$$
\mathcal{C}(D)=\prod_{i \in I_{h} ; r_{i} \neq 0} C\left(r_{i}\right)^{z_{2 i-1}} \cdot\left(\frac{1}{C\left(r_{i}\right)}\right)^{z_{2 i}}
$$

where $z_{2 i-1}$ is the number of applications of $l_{i} \rightarrow T\left(r_{i}\right)(\mathcal{P}(\mathcal{B}))$, and $z_{2 i}$ the number of applications of $T\left(r_{i}\right) \rightarrow l_{i}(\mathcal{P}(\mathcal{B}))$ in $D, i \in I_{h}, r_{i} \neq 0$. Since each $z_{i}, i \in I_{2 h}$, is bounded by $2^{\operatorname{size}\left(t_{1}, t_{2}, \mathcal{B}\right) \cdot 2^{c \cdot k}}$ for some constant $c>0$ independent of $t_{1}, t_{2}$, and $\mathcal{B}$, the multiplicative factor $\mathcal{C}(D)$ of $D$ can be triply exponentially large. Its doubly exponentially long representation can be computed in exponential space using $\mathcal{N C}$-circuits for multiplication (Karp and Ramachandran, 1990) and appealing to the parallel computation thesis

## Constructing the Reduced Gröbner Basis of a Binomial Ideal

Input: admissible term ordering $\succeq$ $\mathcal{B}=\left\{l_{1}-r_{1}, \ldots, l_{h}-r_{h}\right\}$ with $l_{i} \in X^{*}, r_{i} \in M[X], l_{i} \succ r_{i} \forall i \in I_{h}$
Output: the reduced Gröbner basis $G=\left\{h_{1}-m_{1}, \ldots, h_{r}-m_{r}\right\}$ of $I(\mathcal{B})$
$L:=\left\{l_{1}, \ldots, l_{h}\right\} \cap \min \left(\left\{l_{1}, \ldots, l_{h}, T\left(r_{1}\right), \ldots, T\left(r_{h}\right)\right\}\right)$
/* $\min ($.$) denotes the minimal elements of the argument w.r.t. divisibility */$
$R:=\left\{T\left(r_{1}\right), \ldots, T\left(r_{h}\right)\right\} \cap \min \left(\left\{l_{1}, \ldots, l_{h}, T\left(r_{1}\right), \ldots, T\left(r_{h}\right)\right\}\right)$
$k:=$ number of indeterminates $; d:=\max \left\{\operatorname{deg}\left(l_{i}\right), \operatorname{deg}\left(r_{i}\right) ; i \in I_{h}\right\} ; G:=\emptyset$
for each $h=x_{1}^{e_{1}} \cdots x_{k}^{e_{k}} \in L T(\langle L, R\rangle) \backslash L$ with degree $\leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$ do
$g b:=$ false
if $h \in I(\mathcal{B})$ then $g b:=$ true
else
if there exists $t \cdot T\left(r_{i}\right)$ with $h \succ t \cdot T\left(r_{i}\right), T\left(r_{i}\right) \in R, t \in X^{*}, \operatorname{deg}\left(t \cdot r_{i}\right) \leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$ which is $\equiv h \bmod \mathcal{P}(\mathcal{B})$
then $m:=$ the minimal (w.r.t. $\succ$ ) among these terms; $g b:=$ true
end_if
end_if
if $g b$ then $\quad /^{*} h \in L T(I(\mathcal{B}))^{*} / \quad \bar{d}:=\operatorname{deg}(h)$
for each $i \in I_{k}$ with $e_{i} \geq 1$ while $g b$ do $h^{\prime}:=x_{1}^{e_{1}} \cdots x_{i}^{e_{i}-1} \cdots x_{k}^{e_{k}}$
if $\left(h^{\prime} \in I(\mathcal{B})\right.$ or there exists $t \cdot T\left(r_{j}\right)$ with $h^{\prime} \succ t \cdot T\left(r_{j}\right), T\left(r_{j}\right) \in R, t \in X^{*}$, $\operatorname{deg}\left(t \cdot r_{j}\right) \leq(\bar{d}-1)+2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}}$ which is $\left.\equiv h^{\prime} \bmod \mathcal{P}(\mathcal{B})\right)$
then $\quad /^{*} \quad h^{\prime} \in L T(I(\mathcal{B})) \Rightarrow h \notin H^{* /} \quad g b:=$ false
end_if
end_for
end_if
if $g b$ then $/^{*} h \in H^{* /}$
if $h \in I(\mathcal{B})$ then $G:=G \cup\{h\}$
else
$\mathcal{C}(D):=$ the multiplicative factor of a derivation $D$ in $\mathcal{P}(\mathcal{B})$ leading from $h$ to $m$ $G:=G \cup\{h-\mathcal{C}(D) \cdot m\}$
end_if
end_if
end_for
for each $l_{i} \in L$ do
if $l_{i} \in I(\mathcal{B})$ then $G:=G \cup\left\{l_{i}\right\}$
else
$m:=$ the minimal (w.r.t. $\succ$ ) among the terms $t \cdot T\left(r_{j}\right)$ with $l_{i} \succ t \cdot T\left(r_{j}\right), T\left(r_{j}\right) \in R$, $t \in X^{*}$,

$$
\operatorname{deg}\left(t \cdot r_{j}\right) \leq 2 \cdot\left(\frac{d^{2}}{2}+d\right)^{2^{k-1}} \text { which are } \equiv l_{i} \bmod \mathcal{P}(\mathcal{B})
$$

$\mathcal{C}(D):=$ the multiplicative factor of a derivation $D$ in $\mathcal{P}(\mathcal{B})$ leading from $l_{i}$ to $m$ $G:=G \cup\left\{l_{i}-\mathcal{C}(D) \cdot m\right\}$
end_if
end_for
Figure 4. Algorithm for constructing the reduced Gröbner basis of a general binomial ideal.
of Fortune and Wyllie (1978). If it suffices to compute a representation of the reduced Gröbner basis of $I(\mathcal{B})$ with the coefficients given as products of negative and positive powers of prime numbers, then an appropriate representation of $\mathcal{C}(D)$ can be computed
from the representations of the $C\left(r_{i}\right)$ used in $D$ directly without any $\mathcal{N C}$-circuits and the parallel computation thesis.
From the algorithm for constructing the reduced Gröbner basis of pure difference binomial ideals in Figure 1, we obtain the rather similar exponential space algorithm for constructing the reduced Gröbner basis of general binomial ideals given in Figure 4. Putting everything together, we have proved the following theorem:

Theorem 4.4. Let $X=\left\{x_{1}, \ldots, x_{k}\right\}, \mathcal{B}=\left\{l_{i}-r_{i} ; i \in I_{h}\right\}$ with $l_{i} \in X^{*}, r_{i} \in M[X]$ for all $i \in I_{h}$, and $\succeq$ some admissible term ordering. Then there is an algorithm which generates the reduced Gröbner basis $G=\left\{h_{1}-m_{1}, \ldots, h_{r}-m_{r}\right\}$ of the binomial ideal $I(\mathcal{B})$ w.r.t. $\succeq$ using at most space $(\operatorname{size}(\mathcal{B}))^{2} \cdot 2^{\bar{c} \cdot k} \leq 2^{c \cdot \operatorname{size}(\mathcal{B})}$, where $\bar{c}, c>0$ are some constants independent of $\mathcal{B}$.

## 5. Conclusion

The results obtained in this paper first give an algorithm for generating the reduced Gröbner basis of a pure difference binomial ideal using at most space $2^{c \cdot n}$, where $n$ is the size of the problem instance, and $c>0$, some constant independent of $n$. The fundamental concept is the algorithm in Mayr and Meyer (1982) for the uniform word problem in commutative semigroups.

Because of the close relationship between commutative semigroups and pure difference binomial ideals, our basis construction algorithm has a number of applications to finitely presented commutative semigroups. Besides those mentioned in Section 3.3, we are able to derive exponential space complete decision procedures for the coverability, the subword, the finite enumeration, the containment, and the equivalence problems for commutative semigroups (see Koppenhagen and Mayr, 1996a, 1997).

Furthermore, as shown in Section 4, we obtain an algorithm for transforming any given basis into the reduced Gröbner basis for binomial ideals in general, also requiring at most space $2^{d \cdot n}$ for some constant $d>0$ independent of the size $n$ of the problem instance. Since, in the worst case, any Gröbner basis of pure difference binomial ideals can have maximal degree doubly exponential in $n$, any algorithm for computing Gröbner bases of binomial ideals requires at least exponential space (see Mayr and Meyer, 1982; Huynh, 1986).

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[^0]:    $\dagger \mathbb{Q}$ denotes the set of rationals, $\mathbb{N}$ the set of non-negative integers, and $\mathbb{Z}$ the set of integers.

[^1]:    ${ }^{\dagger}$ For $n \in \mathbb{N}, I_{n}$ denotes the set $\{1, \ldots, n\}$.
    ${ }^{\ddagger}$ Let $\Phi$ be the Parikh mapping, i.e. $\Phi\left(u, x_{i}\right)$ (also written $\left.(\Phi(u))_{i}\right)$ indicates, for every $u \in X^{*}$ and $i \in\{1, \ldots, k\}$, the number of occurrences of $x_{i} \in X$ in $u$.

