# Formal asymptotics of solutions to certain linear differential equations involving oscillation 

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#### Abstract

Let $\mathcal{P}=\mathcal{P}^{x}$ denote the ring of analytic $2 \pi$-periodic functions in $x$ on the real axis. Let $\mathcal{S}=\mathcal{S}^{x}$ denote the ring of formal Laurent series in $\mathcal{P}\left(\left(e^{-x}\right)\right)$, whose coefficients are defined on a common strip neighbourhood of the real axis. In this paper, we study the linear differential equation


$$
L_{r}(x) h^{(r)}(x)+\cdots+L_{0}(x) h(x)=0
$$

with coefficients $L_{0}, \ldots, L_{r} \neq 0$ in $\mathcal{S}$. We prove that, after a change of variables $x=$ $p(\tilde{x}+\varphi(\tilde{x}))$ with $p \in \mathbb{N}^{*}$ and $\varphi \in \mathcal{P}^{\tilde{x}}$, this equation admits a basis of $r$ formal solutions of the form

$$
h=\left(\varphi_{r-1}(\tilde{x}) \tilde{x}^{r-1}+\cdots+\varphi_{0}(\tilde{x})\right) \exp (\xi \tilde{x}) \exp \left(\psi_{d}(\tilde{x}) e^{d \tilde{x}}+\cdots+\psi_{1}(\tilde{x}) e^{\tilde{x}}\right)
$$

where $\varphi_{0}, \ldots, \varphi_{r-1} \in \mathcal{S}^{\tilde{x}}, \xi \in \mathbb{C}$ and $\psi_{1}, \ldots, \psi_{d} \in \mathcal{P}^{\tilde{x}}$. This generalizes a well known result when $\mathcal{P}$ is replaced by $\mathbb{C}$.

## 1. Introduction

Consider the linear differential equation

$$
\begin{equation*}
L h=L_{r} h^{(r)}+\cdots+L_{0} h=0 \tag{1.1}
\end{equation*}
$$

It is well known, e.g. (Ince, 1926), that if the coefficients $L_{0}, \ldots, L_{r}$ are power series in $\mathbb{C}[[z]]$, then there exists a basis of $r$ formal solutions to (1.1) of the form

$$
h=\left(h_{r-1} \log ^{r-1} z+\cdots+h_{1} \log z+h_{0}\right) z^{\lambda} e^{P(\sqrt[p]{z})},
$$

where $h_{0}, \ldots, h_{r-1}$ are power series in $\mathbb{C}[[\sqrt[p]{x}]], p \in \mathbb{N}^{*}, \lambda \in \mathbb{C}$ and $P=P_{d} x^{-d / p}+$ $\cdots+P_{1} x^{-1 / p}$ a polynomial in $\mathbb{C}\left[\sqrt[p]{z^{-1}}\right]$ without constant term. When replacing $z$ by $e^{-x}$, it follows that, if the coefficients $L_{0}, \ldots, L_{r}$ are in $\mathbb{C}\left[\left[e^{-x}\right]\right]$, then the differential equation (1.1) admits a basis of $r$ formal solutions of the form

$$
h=\left(h_{r-1} x^{r-1}+\cdots+h_{1} x+h_{0}\right) e^{\lambda x} e^{P\left(e^{x / p}\right)}
$$

where $h_{0}, \ldots, h_{r-1} x^{r-1} \in \mathbb{C}\left[\left[e^{-x / p}\right]\right], p \in \mathbb{N}^{*}, \lambda \in \mathbb{C}$ and $P \in \mathbb{C}\left[e^{x / p}\right] e^{x / p}$. This classical result was generalized in (van der Hoeven, 1997) to the case when the coefficients
$\dagger$ The difference between the dates in the title and on the cover are due to the fact that this paper was declared to be "uninteresting" after a long period of refereeing. Nevertheless, several people have asked me for the paper since then, which made me decide to publish this preprint a long time after its time of writing.
$L_{0}, \ldots, L_{r}$ are transseries. This allows for instance to find bases of formal solutions to equations like

$$
e^{\Gamma\left(e^{x}\right)} f^{\prime \prime}+\frac{e^{x}}{e^{e^{x}}+1} f^{\prime}+\operatorname{erf} e^{1998 x} f=0
$$

A major actual drawback of the actual transseries theory (Écalle, 1992; van der Hoeven, 1997) is that it only modelizes "strongly monotonic" asymptotic behaviour, i.e. we do not allow oscillatory behaviour. In this paper, we make a first step towards the formal study of asymptotic linear differential equations which do involve oscillation.

In section 3 , we start by studying the equation (1.1) when $L_{r}=1, L_{r-1}, \ldots, L_{0} \in \mathcal{P}$, where $\mathcal{P}=\mathcal{P}^{x}$ is the set of analytic $2 \pi$-periodic functions on the real axis in $x$. Notice that elements of $\mathcal{P}$ are actually defined on a small strip neighbourhood of the real axis. We show that there exists a basis of solutions to (1.1) of the form $h \in \mathcal{P}[x] e^{\lambda x}$. We next study the inhomogeneous equation

$$
\begin{equation*}
L h=L_{r} h^{(r)}+\cdots+L_{0} h=g \tag{1.2}
\end{equation*}
$$

with $g \in \mathcal{P}[x]$ and show that this equation always admits a solution (and even a very special, so called "distinguished solution") in $\mathcal{P}[x]$. This result persists in the case when $L_{r} \neq 1$, modulo a change of variables of the form $x=\tilde{x}+\varphi(\tilde{x})$, where $\varphi \in \mathcal{P}^{\tilde{x}}$ is an analytic $2 \pi$-periodic function in $\tilde{x}$ on the real axis.

In section 4 , we consider the case when the coefficients $L_{i}$ are in the set $\mathcal{S}$ of Laurent series in $\mathcal{P}\left(\left(e^{-x}\right)\right)$, whose coefficients are defined on a common strip neighbourhood of $\mathbb{R}$. We prove that, modulo a change of variables $x=p(\tilde{x}+\varphi)$, with $p \in \mathbb{N}^{*}$ and $\varphi \in \mathcal{P}^{\tilde{x}}$, there exists a basis of $r$ solutions to (1.1) of the form

$$
h=\left(\varphi_{r-1}(\tilde{x}) \tilde{x}^{r-1}+\cdots+\varphi_{0}(\tilde{x})\right) \exp (\xi \tilde{x}) \exp \left(\psi_{d}(\tilde{x}) e^{d \tilde{x}}+\cdots+\psi_{1}(\tilde{x}) e^{\tilde{x}}\right)
$$

where $\varphi_{0}, \ldots, \varphi_{r-1} \in \mathcal{S}^{\tilde{x}}, \xi \in \mathbb{C}$ and $\psi_{1}, \ldots, \psi_{d} \in \mathcal{P}^{\tilde{x}}$. We will follow a similar proof strategy as in (van der Hoeven, 1997), based on the Newton polygon method and distinguished solutions. Further generalizations of this result will be treated in a forthcoming paper.

## 2. Preliminaries

### 2.1. The coefficients

Let $\mathcal{P}$ be the space of analytic, $2 \pi$-periodic functions on the real axis. Such functions are actually analytic on a strip neighbourhood of the real axis (i.e. a set of the form $\{z \in \mathbb{C}|\varepsilon>|\Im z|\})$. Let $\mathcal{S}$ be the set of Laurent series $f \in \mathcal{P}\left(\left(e^{-x}\right)\right)$, such that the coefficients $f_{\alpha}$ are analytic on a common strip neighbourhood of the real axis. Clearly, $\mathcal{S}$ forms a ring. We will denote by $v_{f}$ the valuation of $f \in \mathcal{S}$ in $e^{-x}$.

When solving algebraic or differential equations with coefficients in $\mathcal{P}$ or $\mathcal{S}$, we will encounter $2 p \pi$-periodic functions with $p \in \mathbb{N}^{*}$, as well as singularities on the real axis, which need be circumvented by passing in the complex plane. For these reasons, we will consider changes of variables

$$
\begin{equation*}
x=p(\tilde{x}+\varphi(\tilde{x}))=p \gamma(\tilde{x}) \tag{2.1}
\end{equation*}
$$

where $p \in \mathbb{N}^{*}, \varphi \in \mathcal{P}$ and the mapping $\gamma: \tilde{x} \mapsto \tilde{x}+\varphi(\tilde{x})$ is bijective in a strip neighbourhood of $\mathbb{R}$. Such a change of variables is called a narrowing and a composition of two
narrowings is again a narrowing. Usually, $x$ and $\tilde{x}$ are bound to certain strip neighbourhoods $U$ resp. $\tilde{U}$ of $\mathbb{R}$ with $\gamma(\tilde{U}) \subseteq U$ and $\gamma$ bijective on $\tilde{U}$. The number $p$ is called the multiplicator of the narrowing.

Since we will sometimes work concurrently with several variables $x, \tilde{x}$, it will be convenient to write $\mathcal{P}^{x}$ instead of $\mathcal{P}$ if we want to emphasize that its elements are $2 \pi$-periodic in $x$ (similarly, we will consider $\mathcal{P}^{\tilde{x}}, \mathcal{S}^{x}$, etc.)

Proposition 2.1. Consider a polynomial equation with coefficients in $\mathcal{P}^{x}$ :

$$
\begin{equation*}
P(f)=P_{d} f^{d}+\cdots+P_{0}=0 \quad\left(P_{d} \neq 0\right) \tag{2.2}
\end{equation*}
$$

Then there exists a narrowing $x=p(\tilde{x}+\varphi(\tilde{x}))$, such that (2.2) admits $d$ solutions in $\mathcal{P}^{\tilde{x}}$, when counted with multiplicities.

Proof. Without loss of generality, we may assume that $P$ is irreducible. Let $\gamma: t \mapsto$ $t+\psi(t)$ be any immersion with $\psi \in \mathcal{P}^{t}$, such that the resultant of $P$ and $P^{\prime}$ does not vanish on $\operatorname{Im} \gamma$. Then each solution $y_{i}$ to

$$
\begin{equation*}
P_{d}(x) y^{d}+\cdots+P_{0}(x)=0 \tag{2.3}
\end{equation*}
$$

in a point $x_{0} \in \operatorname{Im} \gamma$ determines a unique analytic solution $f_{i}$ to (2.2) on $\operatorname{Im} \gamma$ such that $f_{i}\left(x_{0}\right)=y_{i}$. Since $P_{0}, \ldots, P_{d}$ are $2 \pi$-periodic, there exists a permutation $\sigma$ of $\{1, \ldots, d\}$, such that $f_{i}\left(x_{0}+2 \pi\right)=y_{\sigma(i)}$ for all $1 \leqslant i \leqslant d$. By the uniqueness of analytic continuation and induction over $k$, we infer that $f_{i}(x+2 \pi k)=f_{\sigma^{k}(i)}(x)$ for all $x$ and $k \in \mathbb{N}$. Consequently, if $p \in \mathbb{N}^{*}$ is such that $\sigma^{p}=I d$, then $f_{1}, \ldots, f_{d}$ are all $2 \pi p$-periodic and the narrowing $x=p \gamma(\tilde{x})$ satisfies our requirements.

For each ring $R$ and $n \in \mathbb{N}$, let $R[x]_{d}$ be the set of polynomials of degrees at most $d$ in $x$ over $R$. In what follows we shall often consider polynomials in $\mathcal{S}[x]$ and $\mathcal{S}[x]_{d}$ and interpret such polynomials as Laurent series in $\mathcal{S}$ with coefficients in $\mathcal{P}[x]$ resp. $\mathcal{P}[x]_{d}$.

Let $\mathcal{E}_{d, 0}$ denote the set of finite linear combinations $\varphi_{1} e^{\lambda_{1} x}+\cdots+\varphi_{k} e^{\lambda_{k} x}$, with $\varphi_{1}, \ldots, \varphi_{k} \in \mathcal{S}[x]_{d}$ and $\varphi_{1}, \ldots, \varphi_{k} \in \mathbb{C}$. For each polynomial without constant term $P=P_{d} e^{d x}+\cdots+P_{1} e^{x}$ in $\mathcal{P}\left[e^{x}\right]$, we denote $\mathcal{E}_{d, P}=\mathcal{E}_{d, 0} e^{P}$. We define

$$
\mathcal{E}_{d}=\bigoplus_{P} \mathcal{E}_{d, P}
$$

We will search for solutions to (1.1) in $\mathcal{E}_{r-1}$, modulo a suitable narrowing.

### 2.2. Linear differential operators

Let $\partial_{x}=\frac{d}{d x}$ denote the differentiation operator with respect to $x$. Given a linear differential operator

$$
L=L_{r} \partial_{x}^{r}+\cdots+L_{0}
$$

we define the derivative $L^{\prime}$ of $L$ by

$$
L^{\prime}=r \partial_{x}^{r-1}+\cdots+L_{1}
$$

For any $f$ and $g$, we have the product formula

$$
\begin{equation*}
L(f g)=(L f) g+\left(L^{\prime} f\right) g^{\prime}+\cdots+\frac{1}{r!}\left(L^{(r)} f\right) g^{(r)} \tag{2.4}
\end{equation*}
$$

The operator $L$ is said to be monic, if $L_{r}=1$. In that case, $\frac{1}{r} L^{\prime}$ is monic as well. If the $L_{i}$ are in $\mathcal{S}$, then we will denote by $L_{i, \alpha}$ the coefficient of $e^{-\alpha x}$ in $L_{i}$ for each $i, \alpha$.

Given a linear differential operator $L$ and a function $h$, there exists a unique linear differential operator $L_{\times h}$ such that

$$
L_{\times h}(f)=L(h f)
$$

for all $f$. We call $L_{\times h}$ a multiplicative conjugate of $L$. The coefficients of $L_{\times h}$ are given explicitly by

$$
L_{\times h, i}=L^{(i)} h=\sum_{j=i}^{r}\binom{j}{i} L_{j} h^{(j-i)}
$$

We notice that if $L$ has coefficients in $\mathcal{P}$, then $\mathcal{P} e^{\lambda x}$ and $\mathcal{P}[x] e^{\lambda x}$ are stable under $L$ for each $\lambda \in \mathbb{C}$. Consequently, if $h \in \mathcal{P} e^{\lambda x}$, then $e^{-\lambda x} L_{\times h}$ has coefficients in $\mathcal{P}$.

Given a linear differential operator $L$ and a function $\gamma$, we also define $L \circ \gamma$ to be the unique differential operator with

$$
(L \circ \gamma)(f \circ \gamma)=(L f) \circ \gamma
$$

for all $f$. Such operators are encountered when performing a change of variables $x=\gamma(\tilde{x})$. Setting $\tilde{f}=f \circ \gamma, \widetilde{L f}=(L f) \circ \gamma$ and $\tilde{L}=L \circ \gamma$, we then have $f(x)=\tilde{f}(\tilde{x})$ and $\tilde{L} \tilde{f}=\widetilde{L f}$. The coefficients of $L$ are obtained from the relations

$$
\begin{aligned}
f(x) & =\tilde{f}(\tilde{x}) \\
f^{\prime}(x) & =\gamma^{\prime}(\tilde{x})^{-1} \tilde{f}^{\prime}(\tilde{x}) \\
f^{\prime \prime}(x) & =\gamma^{\prime}(\tilde{x})^{-2} \tilde{f}^{\prime \prime}(\tilde{x})-\gamma^{\prime \prime}(\tilde{x}) \gamma^{\prime}(\tilde{x})^{-3} \tilde{f}^{\prime}(\tilde{x}) \\
& \vdots
\end{aligned}
$$

In particular, if $\gamma(\tilde{x})=\tilde{x}+c$ for some constant $c$, then $f^{(j)}(x)=\tilde{f}^{(j)}(\tilde{x})$ and $\tilde{L}_{j}=L_{j} \circ \gamma$ for all $j$.

## 3. Linear differential equations with periodic coefficients

### 3.1. The monic homogeneous case

Consider the homogeneous linear differential equation (1.1), for coefficients $L_{0}, \ldots$, $L_{r} \in \mathcal{P}$ with $L_{r}=1$. Let $\mathcal{H}$ be the space of analytic solutions to (1.1) on the real axis. Since $L_{r}=1$, we have $\operatorname{dim} \mathcal{H}=r$. Let $\mathcal{C}$ be the space of analytic functions on the real axis and consider the mapping $\Phi: \mathcal{C} \rightarrow \mathcal{C}$ defined by

$$
(\Phi f)(x)=f(x+2 \pi)
$$

Since the coefficients of (1.1) are periodic, $\mathcal{H}$ is stable under $\Phi$. From now on, we will only consider the restriction of $\Phi$ to $\mathcal{H}$, which is an isomorphism, since $\Phi$ is invertible and $\mathcal{H}$ finite dimensional. In particular, all eigenvalues of $\Phi$ are non zero; let $e^{2 \pi \lambda}$ be such an eigenvalue. Modulo the change of function $h \rightarrow h / e^{\lambda x}$, we may assume without loss of generality that $\lambda=0$.

By Jordan's theorem, the characteristic space associated to the eigenvalue $e^{2 \pi \lambda}=1$ can be written as a direct sum of invariant subspaces, each on which there exists a basis
$h_{0}, \ldots, h_{\nu-1}$ with respect to which $\Phi$ is represented by the matrix

$$
\left(\begin{array}{cccc}
1 & & & \mathrm{O} \\
1 & 1 & & \\
& \ddots & \ddots & \\
\mathrm{O} & & 1 & 1
\end{array}\right)
$$

On such a subspace, we have in particular $\Phi h_{0}=h_{0}$, whence $h_{0} \in \mathcal{P}$. Next, $\Phi h_{1}=h_{1}+h_{0}$ and setting $\varphi_{1}=h_{1}-h_{0} \frac{x}{2 \pi}$, we observe that $\Phi \varphi_{1}=h_{1}+h_{0}-h_{0} \frac{x+2 \pi}{2 \pi}=\varphi_{1}$. Therefore, $h_{1} \in \mathcal{P}[x]_{1}$. Similarly, for each $1<j<\nu_{0}$, one has $\Phi \varphi_{j}=\varphi_{j}$, where $\varphi_{j}=h_{j}-h_{j-1} \frac{x}{2 \pi}$. By induction on $j$, it follows that $h_{j} \in \mathcal{P}[x]_{j}$.

For each $\lambda \in \mathbb{C}$, let $\nu_{\lambda}$ be the dimension of the characteristic space $\mathcal{H}_{\lambda}$ associated to the eigenvalue $e^{2 \pi \lambda}$. We have just shown that

$$
\mathcal{H}_{\lambda} \subseteq \mathcal{P}[x]_{\nu_{\lambda}-1} e^{\lambda x}
$$

In other words,

Theorem 3.1. Assume that $L_{r}=1$ and $L_{r-1}, \ldots, L_{0} \in \mathcal{P}$. Then the solution space $\mathcal{H}$ to (1.1) admits a basis of elements of the form

$$
h \in \mathcal{P}[x]_{\nu_{\lambda}-1} e^{\lambda x} \quad(\lambda \in \mathbb{C})
$$

where $\nu_{\lambda}=\operatorname{dim} \mathcal{H} \cap \mathcal{P}[x] e^{\lambda x}$ for each $\lambda \in \mathbb{C}$.

### 3.2. Integration

Lemma 3.1. Let $g=\psi e^{\lambda x}$, with $\psi=\psi_{d} x^{d}+\cdots+\psi_{0} \in \mathcal{P}[x]_{d}$.
(a) If $\lambda \notin \mathbb{Z} i$, then there exists a unique primitive $\int g$ of $g$ in $\mathcal{P}[x]_{d} e^{\lambda x}$.
(b) If $\lambda \in \mathbb{Z} i$, there exists a unique primitive $\int g$ of $g$ in $\mathcal{P}[x]_{d+1}$, such that $\left\langle\left(\int g\right)_{0} \mid 1\right\rangle=0$.

Proof. Setting $f=\varphi e^{\lambda x}$, solving $f^{\prime}=g$ in $\mathcal{P}[x] e^{\lambda x}$ is equivalent to solving

$$
\varphi^{\prime}+\lambda \varphi=\psi
$$

in $\mathcal{P}[x]$. We will search for a solution of the form

$$
\varphi=\varphi_{d+1} x^{d+1}+\cdots+\varphi_{0}
$$

Then we have to solve the following system of equations:

$$
\begin{aligned}
\varphi_{d+1}^{\prime}+\lambda \varphi_{d+1} & =0 \\
\varphi_{d}^{\prime}+\lambda \varphi_{d} & =\psi_{d}-(d+1) \varphi_{d+1} ; \\
& \vdots \\
\varphi_{0}^{\prime}+\lambda \varphi_{0} & =\psi_{0}-\varphi_{1}
\end{aligned}
$$

In what follows, we will denote by $a_{j}$ the coefficient of $e^{-\lambda x}$ in the Fourier series of $\psi_{j}$, for each $j$. If $\lambda \notin i \mathbb{Z}$, then $a_{j}=0$.

We take $\varphi_{d+1}=\frac{1}{d+1} a_{d} e^{\lambda x}$, whence $\varphi_{d+1}=0$, if $\lambda \notin \mathbb{Z} i$. The remaining $\varphi_{j}$ are computed by induction over $j=d, \ldots, 0$. We make the induction hypothesis that $\varphi_{j+1} \in \mathcal{P}$ and that the coefficients of $e^{-\lambda x}$ in the Fourier series of $(j+1) \varphi_{j+1}$ and $\psi_{j}$ coincide. Now let

$$
\sum_{k \in \mathbb{Z}} c_{k} e^{i k x}=\psi_{j}-(j+1) \varphi_{j+1}
$$

be the convergent Fourier series of $\psi_{j}-(j+1) \varphi_{j+1}$. Then we take

$$
\varphi_{j}=\frac{a_{j-1}}{j} e^{-\lambda x}+\sum_{k \in \mathbb{Z}, i k+\lambda \neq 0} \frac{c_{k}}{i k+\lambda} e^{i k x}
$$

which is convergent and periodic (in the case $j=0$, we understand $a_{j-1} / j$ to be zero). Since any solution to $\varphi_{j}^{\prime}+\lambda \varphi_{j}=\psi_{j}-(j+1) \varphi_{j+1}$ is analytic, we have $\varphi_{j} \in \mathcal{P}$. The second induction hypothesis is again satisfied at the next stage, by definition of $\varphi_{j}$.

We have thus shown how to compute a primitive $f=\varphi e^{\lambda x}$ of $g$, with $\varphi \in \mathcal{P}[x]_{d+1}$. Moreover, if $\lambda \notin \mathbb{Z} i$, then $\varphi_{d+1}=0$ and $f \in \mathcal{P}[x]_{d} e^{\lambda x}$. Finally, the primitive of $g$ is unique up to a constant factor. If $\lambda \notin \mathbb{Z} i$, this implies that $f$ is unique in $\mathcal{P}[x]_{d} e^{\lambda x}$ with $f^{\prime}=g$. If $\lambda \in \mathbb{Z} i, f$ is unique in $\mathcal{P}[x]_{d+1}$ with the property that the constant term $\left\langle\varphi_{0} e^{\lambda x} \mid 1\right\rangle$ of $f$ vanishes.

The primitive $\int g$ as constructed in the lemma is called the distinguished primitive of $g$. Notice that the mapping $g \mapsto \int g$ is injective and linear on $\mathcal{P}[x] e^{\lambda x}$, for each $\lambda \in \mathbb{C}$ : this is clear if $\lambda \notin \mathbb{Z} i$; otherwise, it follows from the fact that $\langle\varphi+\psi \mid 1\rangle=\langle\varphi \mid 1\rangle+\langle\psi \mid 1\rangle$ for all $\varphi$ and $\psi$. Consequently, the mapping $\int$ may be extended uniquely to a linear, injective mapping from the subvector space of $\mathcal{C}$ generated by the the vector spaces of the form $\mathcal{P}[x] e^{\lambda x}$ into itself.

Let us denote by $\nu_{L}: \mathbb{C} \rightarrow \mathbb{N}$ the mapping which associates $\nu_{\lambda}$ to $\lambda$. Notice that $\nu_{L}$ factors through $\mathbb{C} / \mathbb{Z} i$, since $\nu_{L}(\lambda+i)=\nu_{L}(\lambda)$. We will now study the dependence of $\nu_{L}$ on $L$.

Lemma 3.2. Let $L$ be a monic linear differential operator in $\mathcal{P}\left[\partial_{x}\right]$. Then

$$
\nu_{L \partial_{x}}=\nu_{L}+\nu_{\partial_{x}} .
$$

Proof. Let $\mathcal{I}$ be the solution space to $\left(L \partial_{x}\right) h=0$ and for each $\lambda \in \mathbb{C}$, let $\mathcal{I}_{\lambda} \subseteq \mathcal{P}[x] e^{\lambda x}$ be the characteristic space associated to $e^{2 \pi \lambda}$, for $\Phi$ restricted to $\mathcal{I}$. Then the distinguished primitivation $\int$ maps $\mathcal{H}$ into $\mathcal{I}$ and $\mathcal{H}_{\lambda}$ into $\mathcal{I}_{\lambda}$ for each $\lambda \in \mathbb{C}$, while $\partial_{x}$ maps $\mathcal{I}$ onto $\mathcal{H}$ and $\mathcal{I}_{\lambda}$ onto $\mathcal{H}_{\lambda}$ for each $\lambda \in \mathbb{C}$. For each $\lambda \notin \mathbb{Z} i$, we infer that

$$
\nu_{L \partial_{x}}(\lambda)=\operatorname{dim} \mathcal{I}_{\lambda}=\operatorname{dim} \int \mathcal{H}_{\lambda}=\operatorname{dim} \mathcal{H}_{\lambda}=\nu_{L}(\lambda) .
$$

For $\lambda \in \mathbb{Z} i$, we get

$$
\nu_{L \partial_{x}}(\lambda)=\operatorname{dim} \mathcal{I}_{\lambda}=\operatorname{dim}\left(\int \mathcal{H}_{\lambda} \oplus \mathbb{C}\right)=\operatorname{dim} \mathcal{H}_{\lambda}+1=\nu_{L}(\lambda)+1
$$

This proves the lemma, since $\nu_{\partial_{x}}(\lambda)=1$ if $\lambda \in \mathbb{Z} i$ and $\nu_{\partial_{x}}(\lambda)=0$ otherwise.

### 3.3. The monic inhomogeneous case

Lemma 3.2 may be generalized as follows:

Lemma 3.3. Let $L, K$ be two monic linear differential operators in $\mathcal{P}\left[\partial_{x}\right]$. Then

$$
\nu_{L K}=\nu_{L}+\nu_{K}
$$

Proof. Let us prove the lemma by induction over the order $s$ of $K$. For $s=0$, we have nothing to do. Assume that $s>0$ and let $h$ be a solution to $K h=0$ in $\mathcal{P} e^{\lambda x}$ for some $\lambda$ (such a solutions exists always: see section 3.1). We will first assume that $h^{-1} \in \mathcal{P} e^{-\lambda x}$.

Since each solution of $h^{-1} L_{\times h} f=0$ in $\mathcal{P}[x] e^{\mu x}$ determines a unique solution to $L f=0$ in $\mathcal{P}[x] e^{(\mu+\lambda) x}$ via multiplication by $h$, we have

$$
\begin{equation*}
\nu_{h^{-1} L_{\times h}}(\mu)=\nu_{L}(\mu+\lambda) \tag{3.1}
\end{equation*}
$$

for all $\mu \in \mathbb{C}$. Given $\mu \in \mathbb{C}$, we have in a similar way

$$
\begin{equation*}
\nu_{h^{-1} K_{\times h}}(\mu)=\nu_{K}(\mu+\lambda) \tag{3.2}
\end{equation*}
$$

and

$$
\begin{equation*}
\nu_{h^{-1}(L K)_{\times h}}(\mu)=\nu_{L K}(\mu+\lambda) \tag{3.3}
\end{equation*}
$$

Since $h^{-1} K h=h^{-1} K_{\times h} 1=0$, we can factor $h^{-1} K_{\times h}=\Omega \partial_{x}$. By the induction hypothesis and (3.1), we get

$$
\nu_{h^{-1} L_{\times h} \Omega}(\mu)=\nu_{L}(\mu+\lambda)+\nu_{\Omega}(\mu) .
$$

By lemma 3.2, we therefore have

$$
\nu_{\left(h^{-1} L_{\times h}\right)\left(h^{-1} K_{\times h}\right)}(\mu)=\nu_{L}(\mu+\lambda)+\nu_{\Omega}(\mu)+\nu_{\partial_{x}}(\mu)
$$

Applying the lemma again, we also have

$$
\nu_{h^{-1} K_{\times h}}(\mu)=\nu_{\Omega}(\mu)+\nu_{\partial_{x}}(\mu)
$$

Combining these two equations with (3.2), we obtain

$$
\nu_{\left(h^{-1} L_{\times h}\right)\left(h^{-1} K_{\times h}\right)}(\mu)=\nu_{L}(\mu+\lambda)+\nu_{K}(\mu+\lambda)
$$

But

$$
\left(h^{-1} L_{\times h}\right)\left(h^{-1} K_{\times h}\right)=h^{-1}(L K)_{\times h}
$$

whence the lemma follows from (3.3) in the case when $h^{-1} \in \mathcal{P} e^{-\lambda x}$.
In general, when $e^{-\lambda x} h$ is not invertible in $\mathcal{P}$, we consider a change of variables $x=$ $\tilde{x}+i \varepsilon$, with $\varepsilon \in \mathbb{R}^{*}$ sufficiently small, such that $h$ does not vanish on $i \varepsilon+\mathbb{R}$. Applying the previous argument to the operators $\tilde{L}=L \circ \gamma, \tilde{K}=K \circ \gamma$ and $\widetilde{L K}=(L K) \circ \gamma=\tilde{L} \tilde{K}$, we then find $\nu_{\widetilde{L K}}=\nu_{\tilde{L}}+\nu_{\tilde{K}}$. Moreover, $\nu_{\tilde{L}}=\nu_{L}$, since any solution $f \in \mathcal{P}[x] e^{\mu x}$ to $L f=0$ determines a unique solution $\tilde{f}=f \circ \gamma \in \mathcal{P}^{\tilde{x}}[\tilde{x}] e^{\mu \tilde{x}}$ to $\tilde{L} \tilde{f}=0$. Similarly, $\nu_{\tilde{K}}=\nu_{K}$ and $\nu_{\overparen{L K}}=\nu_{L K}$, whence the lemma.

Theorem 3.2. Assume that $L_{r}=1, L_{r-1}, \ldots, L_{0} \in \mathcal{P}$ and $g \in \mathcal{P}$. Then (1.2) admits at least one solution in $\mathcal{P}[x]_{\nu_{L}(0)}$.

Proof. Assume first that $g$ is invertible in $\mathcal{P}$. Then $\partial_{x}\left(g^{-1} L_{\times g}\right)$ is a monic operator with coefficients in $\mathcal{P}$ and $\left(\partial_{x}\left(g^{-1} L_{\times g}\right)\right)(f / g)=\left(\partial_{x}\left(g^{-1} L\right)\right)(f)=0$, for any solution $f$ to (1.2). Inversely, there exists a solution $h$ to

$$
\begin{equation*}
\partial_{x}\left(g^{-1} L\right)(h)=0 \tag{3.4}
\end{equation*}
$$

such that $g^{-1} L h=1$ : otherwise, $L h$ would vanish for all solutions to (3.4) and the dimension of $\mathcal{H}$ would be at least $r+1$.

Let us now write $h=h_{0} e^{\lambda_{0} x}+\cdots+h_{k} e^{\lambda_{k} x}$, with $h_{0}, \ldots, h_{k} \in \mathcal{P}[x], \lambda_{0}=0$ and pairwise distinct $\lambda_{j}$ modulo $i$. For each $j>0$, we observe that $g^{-1} L\left(h_{j} \lambda_{j}\right) \in \mathcal{P}[x] e^{\lambda_{j} x}$, whence $g^{-1} L\left(h_{j} \lambda_{j}\right)=0$. Hence $f=h_{0} \in \mathcal{P}[x]$ is again a solution to (3.4) and $g^{-1} L f=1$. Now lemma 3.3 implies that

$$
\nu_{\partial_{x}\left(g^{-1} L_{\times g}\right)}(0)=\nu_{g^{-1} L_{\times g}}(0)+1=\nu_{L}(0)+1,
$$

whence $f \in \mathcal{P}[x]_{\nu_{L}(0)}$. This completes the proof in the case when $g$ is invertible in $\mathcal{P}$.
In general, let $c \in \mathbb{R}$ be such that $c>\left|\sup _{x \in \mathbb{R}} g(x)\right|$ and decompose $g=c+\tilde{g}$. Then $c$ and $\tilde{g}$ are both invertible and by what precedes, there exist solutions to $L f_{1}=c$ and $L f_{2}=\tilde{g}$ in $\mathcal{P}[x]_{\nu_{L}(0)}$. Consequently, $f=f_{1}+f_{2}$ is a solution to (1.2) in $\mathcal{P}[x]_{\nu_{L}(0)}$.

Corollary 3.3. Assume that $L_{r}=1, L_{r-1}, \ldots, L_{0} \in \mathcal{P}$ and $g \in \mathcal{P}[x]_{d}$. Then (1.2) has at least one solution in $\mathcal{P}[x]_{d+\nu_{L}(0)}$.

Proof. We prove the corollary by induction over $d$. In the case $d=-1$ we have nothing to do. Assume therefore that $d \geqslant 0$. By theorem 3.2, there exists a $\varphi \in \mathcal{P}[x]_{\nu_{L}(0)}$, with $L \varphi=g_{d}$. Then

$$
L\left(\varphi x^{d}\right)=g_{d} x^{d}+d\left(L^{\prime} \varphi\right) x^{d-1}+\cdots+L^{(d)} \varphi
$$

Consequently, $g-L\left(\varphi x^{d}\right) \in \mathcal{P}[x]_{d-1}$. By the induction hypothesis, there exists a $\psi \in$ $\mathcal{P}[x]_{d+\nu_{L}(0)-1}$, such that $L \psi=g-L\left(\varphi x^{d}\right)$. We conclude that $f=\varphi x^{d}+\psi$ is an element in $\mathcal{P}[x]_{d+\nu_{L}(0)}$ with $L f=g$.

Let us now show how to privilege a particular solution to (1.2) among the solutions in $\mathcal{P}[x]_{d+\nu_{L}(0)}$. This solution will be called the "distinguished primitive" to $L f=g$ and coincides with the distinguished integral if $L=\partial_{x}$. We first recall that $\mathcal{P}$ is a Hilbert space for the Hermitian form defined by

$$
\langle f \mid g\rangle=\frac{1}{2 \pi} \int_{0}^{2 \pi} f(x) \overline{g(x)} d x
$$

For each $j \geqslant 0$, let $H_{j}$ be the vector space of $h_{j} \in \mathcal{P}$, such that there exists a solution $h \in \mathcal{P}[x]$ to $L h=0$ of the form $h=h_{j} x^{j}+\cdots+h_{0}$. For each $f=f_{k} x^{k}+\cdots+f_{0}$, we define $\pi_{L, x^{j}}(f)$ to be the orthogonal projection of $f$ on $H_{j}$. Notice that the operator $\pi_{L, x^{j}}$ is linear.

Theorem 3.4. Assume that $L_{r}=1, L_{r-1}, \ldots, L_{0}$ are in $\mathcal{P}$ and $g \in \mathcal{P}[x]_{d}$. Then there exists a unique solution $f$ in $\mathcal{P}[x]_{d+\nu_{L}(0)}$ to (1.2), such that $\pi_{L, x^{j}}(f)=0$ for all $j$. This solution, which is denoted by $L^{-1} g$, is called the distinguished solution to $L f=g$. The mapping $g \mapsto L^{-1} g$ is linear.

Proof. Let $f$ be a solution to $L f=g$ in $\mathcal{P}[x]_{d+\nu_{L}(0)}$. Let $j$ be maximal such that $\pi_{L, x^{j}}(f) \neq 0$, if such a $j$ exists, and let $h=h_{j} x^{j}+\cdots+h_{0}$ be a solution to $L h=0$ with $\pi_{L, x^{j}}(f)=h_{j}$. Then $\tilde{f}=f-h$ is again a solution to (1.2) in $\mathcal{P}[x]_{d+\nu_{L}(0)}$, but the minimal index $\tilde{\jmath}$ with $\pi_{L, x^{\tilde{\jmath}}}(\tilde{f}) \neq 0$ is strictly smaller than $j$, if such a $\tilde{\jmath}$ exists. Repeating the procedure, we therefore obtain a solution to (1.2) with $\pi_{L, x^{j}}(f)=0$ for all $j$.

Assume that $\tilde{f}$ is a second solution to (1.2) with $\pi_{L, x^{j}}(\tilde{f})=0$ for all $j$. If $\tilde{f} \neq f$, then we would be able to write $h=\tilde{f}-f=h_{j} x^{j}+\cdots+h_{0}$, with $h_{j} \neq 0$ and $0=\pi_{L, x^{j}}(\tilde{f}-f)=$ $\pi_{L, x^{j}}\left(h_{j} x^{j}\right)=h_{j}$, which is impossible. Therefore, $\tilde{f}=f$.

Now consider $g_{1}, g_{2} \in \mathcal{P}[x]$ and let $f_{1}=L^{-1} g_{1}, f_{2}=L^{-1} g_{2}$. We have $L\left(f_{1}+f_{2}\right)=$ $g_{1}+g_{2}$ and $\pi_{L, x^{j}}\left(f_{1}+f_{2}\right)=\pi_{L, x^{j}}\left(f_{1}\right)+\pi_{L, x^{j}}\left(f_{2}\right)=0$ for all $j$. Consequently, $L^{-1}\left(g_{1}+\right.$ $\left.g_{2}\right)=f_{1}+f_{2}$, i.e. $L^{-1}$ is linear.

## 4. Asymptotic linear differential equations

### 4.1. The Newton polygon method

Consider the linear differential equation (1.1), with coefficients $L_{0}, \ldots, L_{r} \in \mathcal{S}$. Each iterated derivative of $h$ may be expressed as $h$ times a differential polynomial $h^{(j)}=$ $R_{j}(f) h$ in the logarithmic derivative $f=h^{\prime} / h$ of $h$. For instance, $R_{0}(f)=1, R_{1}(f)=$ $f, R_{2}(f)=f^{2}+f^{\prime}, R_{3}=f^{3}+3 f^{\prime} f+f^{\prime \prime}$. Hence, solving (1.1) is equivalent to solving the Ricatti equation

$$
L_{r} R_{r}(f)+\cdots+L_{0} R_{0}(f)=0
$$

modulo one integration and one exponentiation: $h=e^{\int f}$. We will use the Newton polygon method in order to solve this equation.

For this purpose, we will actually show how to solve the slightly more general, asymptotic Ricatti equation

$$
\begin{equation*}
R(f)=L_{r} R_{r}(f)+\cdots+L_{0} R_{0}(f)=0 \quad\left(v_{f}>\omega\right) \tag{4.1}
\end{equation*}
$$

with coefficients $L_{0}, \ldots, L_{r} \in \mathcal{S}$ and integer $\omega<0$ or $\omega=-\infty$. We recall that $v_{f} \in \mathbb{Q} \cup$ $\{\infty\}$ denotes the valuation of $f$ in $e^{-x}$. Two main types of solutions can be distinguished: those for which $v_{f} \geqslant 0$ and those for which $v_{f}<0$. Actually, the Newton polygon method will be used in order to reduce the resolution of (4.1) to the case when we only need to find the solutions with $v_{f} \geqslant 0$. In section 4.2 , we will show how to solve this special case using the results from section 3 .

If $v_{f}<0$, then $R_{j}(f)$ and $f^{j}$ coincide up to lower order terms for all $j$, i.e. $v_{R_{j}(f)-f^{j}}>$ $v_{R_{j}(f)}$. Hence, the first term $c e^{-\mu x}$ of a solution to (4.1) with $v_{f}<0$ must also be the first term of a solution to the asymptotic algebraic equation

$$
\begin{equation*}
L_{r} f^{r}+\cdots+L_{0}=0 \quad\left(0>v_{f}>\omega\right) \tag{4.2}
\end{equation*}
$$

The exponent $\mu \in \mathbb{Q}$ of such a first term can be read of from the Newton polygon and the coefficient $c$ is a root of a Newton polynomial (see section 4.3), which is an algebraic equation over $\mathcal{P}$. Furthermore, proposition 2.1 ensures that we may assume without loss of generality that these "potential dominant terms" $c e^{-\mu x}$ of $f$ are in $\mathcal{S}$, modulo a narrowing of $x$.

Assume that we have determined such a potential dominant term $c e^{-\mu x} \in \mathcal{S}$ of a
solution $f$ to 4.1. We then consider the refinement

$$
\begin{equation*}
f=c e^{-\mu x}+\tilde{f} \quad\left(v_{\tilde{f}}>\mu\right) \tag{4.3}
\end{equation*}
$$

i.e. a simultaneous change of functions and the imposition of an asymptotic constraint. Then (4.1) transforms into a new asymptotic Ricatti equation

$$
\begin{equation*}
\tilde{L}_{r} R_{r}(\tilde{f})+\cdots+\tilde{L}_{0} R_{0}(\tilde{f})=0 \quad\left(v_{\tilde{f}}>\mu\right) \tag{4.4}
\end{equation*}
$$

which has again coefficients in $\mathcal{S}$. In section 4.5 , we shall see that the recursive application of this method enables us to find $r$ linearly independent solutions to (1.1) in $\mathcal{E}_{r-1}$.

### 4.2. Distinguished solutions and applications

Assume that $L_{r}=1, L_{r-1}, \ldots, L_{0} \in \mathcal{S}$ and $g \in \mathcal{S}[x]_{d}$. Let $v_{L}$ be the minimum of the valuations of the $L_{i}$ in $e^{-x}$. We define the dominant part $L^{\text {dom }}$ of $L$ to be the linear differential operator with $L_{i}^{\text {dom }}=L_{i, v_{L}}$, where $L_{i, v_{L}}$ denotes the coefficient of $e^{-v_{L} x}$ in $L_{i}$. We notice that $L_{\times e^{\alpha x}}^{\mathrm{dom}}=\left(L_{\times e^{\alpha x}}\right)^{\text {dom }}$ and $L^{\text {dom }}=\left(e^{\alpha x} L\right)^{\text {dom }}$ for all $\alpha \in \mathbb{Z}$. Given $f \in \mathcal{S}, j \in \mathbb{N}$ and $\alpha \in \mathbb{Z}$, we denote $\pi_{L, x^{j} e^{\alpha x}}(f)=\pi_{\left(e^{-\alpha x} L_{x}{ }^{\alpha x}\right)^{\text {dom }}, x^{j}}\left(f_{\alpha}\right)$, where $\pi_{\left(e^{-\alpha x}\right.}^{\left.L_{\times e^{\alpha x}}\right)^{\text {dom }}, x^{j}}$ is as in section 3.3. We also denote $\nu_{L}(\alpha)=\nu_{L^{\text {dom }}}(\alpha)=\nu_{e^{-\alpha x}} L_{\times e^{\alpha x}}^{\text {dom }}(0)$ for all $\alpha$ and $\nu_{L}^{+}(\alpha)=\sum_{\beta \in \mathbb{N}} \nu_{L}(\alpha-\beta)$.

Theorem 4.1. Let $L_{0}, \ldots, L_{r} \in \mathcal{S}, g \in \mathcal{S}[x]_{d}$ and assume that $L^{\mathrm{dom}}$ is monic. Then there exists a unique solution $f$ to (1.2) in $\mathcal{S}_{d+\nu_{L}^{+}\left(v_{L}-v_{g}\right)}$, such that $\pi_{L, x^{j} e^{\alpha x}}(f)=0$ for all $\alpha \in \mathbb{Z}$ and $j \in \mathbb{N}$. We call $f$ the distinguished solution to (1.2) and denote it by $L^{-1} g$. The operator $g \mapsto L^{-1} g$ is linear.

Proof. Without loss of generality, we may assume that $v_{L}=0$, modulo a multiplication of (1.2) by $e^{v_{L} x}$. We first observe that $v_{f} \geqslant v_{g}$. Indeed, otherwise $L_{\times e^{-v_{f} x}}^{\mathrm{dom}} f_{v_{f}}=0$, since $L f=\left(L_{\times e^{-v_{f} x}}^{\mathrm{dom}} f_{v_{f}}+o(1)\right) e^{-v_{f} x}$. Consequently, if $f_{v_{f}, j} x^{j}$ is the leading term of $f_{v_{f}}$, we would have $0=\pi_{L, x^{j} e^{v_{f} x}}(f)=f_{v_{f}, j} \neq 0$.

Let us now show how to compute the coefficients $f_{v_{g}}, f_{v_{g}+1}, \ldots$ of $f$ by induction. Assume that $f_{v_{g}}, \ldots, f_{\alpha-1}$ have been constructed and that $\tilde{g}=g-L\left(f_{v_{g}} e^{-v_{g} x}+\cdots+\right.$ $\left.f_{\alpha-1} e^{-(\alpha-1) x}\right)$ is in $\mathcal{P}[x]_{d+\nu_{-v_{g}}+\cdots+\nu_{1-\alpha}}\left(\left(e^{-x}\right)\right)$, with valuation $v_{\tilde{g}} \geqslant \alpha$. By theorem 3.4,

$$
\begin{equation*}
f_{\alpha}=\left(L_{\times e^{-\alpha x}}^{\text {dom }}\right)^{-1} \tilde{g}_{\alpha} \tag{4.5}
\end{equation*}
$$

is the only solution to the equation $L_{\times e^{-\alpha x}}^{\mathrm{dom}} f_{\alpha}=\tilde{g}_{\alpha}$ in $\mathcal{P}[x]_{d+\nu_{-v_{g}}+\cdots+\nu_{\alpha}}$ with $\pi_{L, x^{j} e^{-\alpha x}}\left(f_{\alpha} e^{-\alpha x}\right)=\pi_{L_{x e^{-\alpha x}}^{\text {dom }}, x^{j}}\left(f_{\alpha}\right)=0$ for all $j$. By construction, the valuation of

$$
g-L\left(f_{v_{g}} e^{-v_{g} x}+\cdots+f_{\alpha} e^{-\alpha x}\right)=\left(\tilde{g}-e^{-\alpha x} L_{\times e^{-\alpha x}}^{\mathrm{dom}} f_{\alpha}\right)-\left(L_{\times e^{-\alpha x}}-e^{-\alpha x} L_{\times e^{-\alpha x}}^{\mathrm{dom}} f_{\alpha}\right)
$$

is at least $\alpha+1$.
We conclude that $f_{v_{g}}, f_{v_{g}+1}, \ldots \in \mathcal{P}[x]_{d+\nu_{L}^{+}\left(-v_{g}\right)}$ are uniquely determined by the conditions that $\pi_{L, x^{j} e^{-\alpha x}}\left(f_{\alpha} e^{-\alpha x}\right)=0$ for all $j, \alpha$ and $g-L\left(f_{v_{g}} e^{-v_{g} x}+\cdots+f_{\alpha} e^{-\alpha x}\right)$ has valuation $>\alpha$ for all $\alpha$. It follows that $f=f_{v_{g}} e^{-v_{g} x}+f_{v_{g}+1} e^{-\left(v_{g}+1\right) x}+\cdots$ is the unique solution in $\mathcal{P}[x]_{d+\nu_{L}^{+}\left(-v_{g}\right)}\left(\left(e^{-x}\right)\right)$ to (1.2), such that $\pi_{L, x^{j} e^{\alpha x}}(f)=0$ for all $\alpha \in \mathbb{Z}$ and $j \in \mathbb{N}$. Since $L^{\text {dom }}$ is monic, the operator $L_{\times e^{-\alpha x}}^{\text {dom }}$ is monic for each $\alpha$. Consequently, the $f_{\alpha}$, which are given by (4.5), are defined on the same common strip neighbourhood of $\mathbb{R}$

Formal asymptotics of solutions to certain linear differential equations involving oscillation
as the coefficients of the $L_{i}$ and $g$. The operator $L^{-1}$ is linear for the same reason as in the proof of theorem 3.4.

Corollary 4.2. Let $L_{0}, \ldots, L_{r} \in \mathcal{S}$ be such that $L^{\text {dom }}$ is monic and let $d_{0}$ be the order of $L^{\mathrm{dom}}$. Then the solutions to (1.1) in $\mathcal{E}_{d_{0}-1,0}$ form a vector space of dimension $d_{0}$.

Proof. By theorem 3.1, the vector space of solutions to $L^{\text {dom }} \varphi=0$ in $\mathcal{E}_{d_{0}-1,0}$ admits a basis $\varphi_{1}, \ldots, \varphi_{d_{0}}$ of solutions of the form $\varphi_{i} \in \mathcal{P}[x]_{\nu_{L}\left(-\lambda_{i}\right)-1} e^{-\lambda_{i} x}$. Each $\varphi_{i}$ determines a solution

$$
h_{i}=e^{-\lambda_{i} x}\left(e^{\lambda_{i} x} L_{\times e^{-\lambda_{i} x}}\right)^{-1}\left(e^{\lambda_{i} x} L_{\times e^{-\lambda_{i} x}}\right)\left(e^{\lambda_{i} x} \varphi_{i}\right)
$$

to (1.1) in $\mathcal{P}[x]_{\nu_{L}\left(-\lambda_{i}\right)-1+\nu_{L}^{+}\left(-\lambda_{i}-1\right)} e^{\lambda_{i} x} \mathcal{S} \subseteq \mathcal{E}_{d_{0}-1,0}$ with dominant term $\varphi_{i}$.
We claim that the $h_{i}$ are linearly independent. Assume for contradiction that $h=$ $c_{1} h_{1}+\cdots+c_{d_{0}} h_{d_{0}}=0$ for certain constants $c_{1}, \ldots, c_{d_{0}}$, not all zero. We may reorder the $h_{1}, \ldots, h_{d_{0}-1}$, such that $c_{1}, \ldots, c_{k}$ are the non zero constants, for which $\Re \lambda_{1}=\cdots=\Re \lambda_{k}$ are minimal. Then the dominant term of $h$ (as a series in $e^{-x}$ whose coefficients are linear combinations of elements in $\mathcal{P}$ times exponentials $e^{-\lambda x}$ with $\Re \lambda=0$ ) is $c_{1} \varphi_{1}+\cdots+c_{k} \varphi_{k}$, which is non zero; contradiction.

On the other hand, the dominant term $\varphi$ of a solution to (1.1) in $\mathcal{E}_{d_{0}-1,0}$ necessarily satisfies $L^{\mathrm{dom}} \varphi=0$. Consequently, we may rewrite $\varphi$ as a linear combination of the $h_{i}$ plus an asymptotically smaller solution to (1.1). Repeating this procedure, we conclude that $h_{1}, \ldots, h_{d_{0}}$ forms a basis for the solutions to (1.1) in $\mathcal{E}_{d_{0}-1,0}$.

Corollary 4.3. Let $L_{0}, \ldots, L_{r} \in \mathcal{S}$ and let $d_{0}$ be the order of $L^{\text {dom }}$. Then there exists a narrowing $\tilde{x}$ of $x$, such that the solutions to (1.1) in $\mathcal{E}_{d_{0}-1,0}^{\tilde{x}}$ form a vector space of dimension $d_{0}$.

Proof. Apply the previous corollary to $L / L_{d_{0}}^{\text {dom }}$, for any narrowing $x=\tilde{x}+\varphi \quad\left(\varphi \in \mathcal{P}^{\tilde{x}}\right)$, such that $L_{d_{0}}^{\text {dom }}$ does not vanish for $\tilde{x} \in \mathbb{R}$.

### 4.3. Finding the potential dominant terms

In this section, we are interested in finding potential dominant terms $c e^{-\mu x}$ of solutions to (4.1) with $v_{f}<0$. We already noticed that such terms coincide with the potential dominant terms of the solutions to (4.2).

We say that $\mu$ with $\omega<\mu<0$ is a potential dominant exponent of $f$, if there exist indices $j<k$ with $v_{L_{j}}+j \mu=v_{L_{k}}+k \mu$ and $v_{L_{l}}+l \mu \geqslant v_{L_{j}}+j \mu$ for all other indices $l$. There are only a finite number of such $\mu$, which can be read of graphically from the Newton polygon associated to (4.2); for instance, see (van der Hoeven, 1997).

Given any $\mu<0$, let $j$ be an index such that $v_{L_{l}}+l \mu \geqslant v_{L_{j}}+j \mu=\alpha$ for all other $l$. Then we call

$$
\begin{equation*}
\Lambda(\lambda)=L_{r, \alpha-r \mu} \lambda^{r}+\cdots+L_{0, \alpha} \tag{4.6}
\end{equation*}
$$

the Newton polynomial associated to $\mu$, where $L_{j, \beta}$ denotes the coefficient of $e^{-\beta x}$ in $L_{j}$. We call $c e^{-\mu x}$ a potential dominant term of $f$, if $c$ is a non zero root (in the algebraic closure of $\mathcal{P}$ ) of the Newton polynomial $\Lambda$ associated to $\lambda$. The multiplicity of $c e^{-\mu x}$ is the multiplicity of $c$ as a root of $\Lambda$.

Clearly, if $c e^{-\mu x}$ is a potential dominant term, then $\mu$ must be a potential dominant
exponent, since $\Lambda$ should contain at least two terms in order to admit a non zero root. It is also readily checked that the dominant term $c e^{-\mu x}$ of a solution $f$ to (4.1) with $v_{f}<0$ must necessarily be a potential dominant term: otherwise, $R(f)$ would be equal to $\Lambda(c) e^{-\alpha x}$ plus lower order terms.

The Newton degree $d$ of (4.1) is the largest index $d$, such that $v_{L_{d}}+d \omega \leqslant v_{L_{j}}+j \omega=\alpha$ for all other indices $j$. It can be shown that this degree either coincides with the largest possible degree of a Newton polynomial associated to a potential dominant exponent $\mu<0$, or with the order $d_{0}$ of the dominant part of $L$, if there are no potential dominant exponents.

Lemma 4.1. Let $d_{0}$ be the order of $L^{\text {dom }}$. Then there are precisely $d-d_{0}$ potential dominant terms $c e^{-\mu x}$ of $f$ with $\mu<0$, when counted with multiplicities.

Proof. Let $0>\mu_{1}>\cdots>\mu_{m}$ be the potential dominant exponents of $f$. Each potential dominant exponent $\mu_{i}$ is determined by two indices $j_{i}<k_{i}$, which are the first projections of the extremities of the corresponding edge of the Newton polygon; $j_{i}$ and $k_{i}$ are respectively the valuation and the degree of the Newton polynomial associated to $\mu_{i}$, therefore this polynomial has $k_{i}-j_{i}$ non zero roots. But $d_{0}=j_{1}<k_{1}=j_{2}<\cdots<k_{m-1}=j_{m}<$ $k_{m}=d$, whence, counting with multiplicities, there are $\left(k_{1}-j_{1}\right)+\cdots+\left(k_{m}-j_{m}\right)=d-d_{0}$ potential dominant terms of $f$.

### 4.4. NarRowings and refinements

Assuming that we know the potential dominant terms of solutions to (4.1), we now want to perform a narrowing followed by a refinement in order to find the next terms of the solutions.

## Lemma 4.2 .

(a) There exists a narrowing (2.1), such that all potential dominant terms of solutions to (4.1) are in $\mathcal{S}^{\tilde{x}}$.
(b) If there exists a potential dominant term whose multiplicity is equal to the Newton degree $d$ of (4.1), then this narrowing may be chosen with multiplicator $p=1$.
(c) The Newton degree of the asymptotic Ricatti equation (4.1) rewritten with respect to the new coordinate $\tilde{x}$ is again $d$.

Proof. Part (a) results from an iterative application of proposition 2.1 to all Newton polynomials associated to a potential dominant exponent of a solution to (1.1).

Now assume that the Newton polynomial $\Lambda$ associated to some potential dominant exponent $\mu$ has a root of multiplicity $d$. Then $\Lambda(\lambda)$ is a constant multiple of $(\lambda-c)^{d}$. In particular, $\Lambda_{0}=L_{0, \alpha}$ and $\Lambda_{1}=L_{1, \alpha-\mu}$ both do not vanish, so that $\mu \in \mathbb{Z}$. Furthermore, $c$ is actually the root of the polynomial $P^{(d-1)}$ of degree one with coefficients in $\mathcal{P}$. Consequently, $c$ is $2 \pi$-periodic and meromorphic on $\mathbb{R}$. Therefore, any narrowing (2.1) with $p=1$, such that $\Im \gamma_{\tilde{f}}$ contains no poles of $c$, meets our requirements in (b).

As to (c), let $f(x)=\tilde{f}(\tilde{x})$ and $\tilde{L}=L \circ \gamma$. Then (4.1) transforms into an asymptotic Ricatti equation

$$
\begin{equation*}
\tilde{L}_{r} R_{r}(\tilde{f})+\cdots+\tilde{L}_{0} R_{0}(\tilde{f})=0 \quad\left(\tilde{v}_{\tilde{f}}>p \omega\right) \tag{4.7}
\end{equation*}
$$

with coefficients $\tilde{L}_{0}, \ldots, \tilde{L}_{r}$ in $\mathcal{S}^{\tilde{x}}$ and where $\tilde{v}_{\tilde{f}}$ denotes the valuation of $\tilde{f}$ in $e^{-\tilde{x}}$. Since each $f^{(j)}(x)$ is a $\mathcal{P}^{\tilde{x}}$-linear combination of $\tilde{f}^{\prime}(\tilde{x}), \ldots, \tilde{f}^{(j)}(\tilde{x})$, each $\tilde{L}_{j}(\tilde{x})$ is a $\mathcal{P}^{\tilde{x}}$-linear combination of $L_{j}(\gamma(\tilde{x})), \ldots, L_{r}(\gamma(\tilde{x}))$. Notice that $\tilde{v}_{L_{j}(\gamma(\tilde{x}))}=p v_{L_{j}(x)}$ for each $j$. By the definition of the Newton degree, $\alpha=v_{L_{d}}+d \omega$ is such that $v_{L_{j}}+j \omega \geqslant \alpha$ for $j \geqslant d$ and $v_{L_{j}}+j \omega>\alpha$ for $j<d$. Consequently, $\tilde{v}_{\tilde{L}_{j}} \geqslant p \min (\alpha-j \omega, \ldots, \alpha-r \omega)=p \alpha-p \omega j$ for $j \geqslant d$ and similarly $\tilde{v}_{\tilde{L}_{j}}>p \alpha-p \omega j$ for $j<d$. Furthermore, $\tilde{v}_{\tilde{L}_{d}(\tilde{x})-L_{d}(\gamma(\tilde{x}))} \geqslant p \alpha-p \omega(d+1)$, whence $\tilde{v}_{\tilde{L}_{d}}=p v_{L_{d}}=p \alpha-p \omega d$. Therefore, the Newton degree of (4.7) is $d$.

Lemma (4.2) ensures us that modulo a narrowing, and without altering the Newton degree of (4.1), we may assume without loss of generality that all potential dominant terms of solutions to (4.1) are in $\mathcal{S}$.

Given such a potential dominant monomial $c e^{-\mu x}$, the change of variables $f=c e^{-\mu x}+\tilde{f}$ in the refinement (4.3) corresponds to the change of variables $h=e^{\int c e^{-\mu x}} \tilde{h}$ in the linear differential equation (1.1). Consequently, $\tilde{h}$ also satisfies a linear differential equation, whence (4.4) is again an asymptotic Ricatti equation; actually, $\tilde{L}=e^{-\int c e^{-\mu x}} L_{\times} \int c e^{-\mu x}$.

Furthermore, each solution $\tilde{h}$ to $\tilde{L} \tilde{h}=0$ in $\mathcal{E}_{r-1}$, whose logarithmic derivative $\tilde{f}=$ $\tilde{h}^{\prime} / \tilde{h}$ satisfies (4.4), induces a solution $h=e^{\int c e^{-\mu x}} \tilde{h}$ to (1.1) in $\mathcal{E}_{r-1}$, whose logarithmic derivative $f=h^{\prime} / h$ satisfies (4.1). Indeed, we may take $\int c e^{-\mu x}=e^{-\mu x}\left(e^{\mu x} \partial_{\times e^{-\mu x}}\right)^{-1} c \in$ $\mathcal{P} e^{-\mu x}$.

Lemma 4.3. Assume that $c e^{-\mu x}$ is a potential dominant term to a solution of (4.1) in $\mathcal{S}$. Then the Newton degree of (4.4) is equal to the multiplicity of c as a root of the Newton polynomial $\Lambda$ associated to $\mu$.

Proof. Let $d$ denote the Newton degree of (4.1) and let $\alpha=v_{L_{d}}+d \mu$. We notice that $v_{L_{j}}+j \mu \leqslant \alpha$ for all indices $j$, by the definition of the Newton degree $d$. Now using the fact that $R_{k}(f)$ and $f^{k}$ coincide up to lower order terms for all $j$, we may express $\tilde{L}_{j}$ in terms of the $L_{k}$ by

$$
\begin{aligned}
\tilde{L}_{j} & =\sum_{k=j}^{n}\binom{k}{j}\left(L_{k, \alpha-k \mu}+o(1)\right) e^{-(\alpha-k \mu) x}\left(c e^{-\mu x}\right)^{k-j} \\
& =\left(\Lambda^{(j)}(c)+o(1)\right) e^{-(\alpha+j \mu) x}
\end{aligned}
$$

Denoting by $\tilde{d}$ the multiplicity of $c$ as a root of $\Lambda$, we have in particular $v_{\tilde{L}_{\tilde{d}}}=\alpha+\tilde{d} \mu$, $v_{\tilde{L}_{j}} \geqslant \alpha+j \mu$ for $j \geqslant \tilde{d}$ and $v_{\tilde{L}_{j}}>\alpha+j \mu$ for $j<\tilde{d}$. In other words, the Newton degree of (4.4) equals $\tilde{d}$.

### 4.5. Solving the homogeneous Equation

Theorem 4.4. Assume that we are given an asymptotic Ricatti equation (4.1) of Newton degree $d$. Then there exists a narrowing $\tilde{x}$ of $x$, such that (1.1) admits $d$ linearly independent solutions in $\mathcal{E}_{d-1}^{\tilde{x}}$, whose logarithmic derivatives are solutions to (4.1).

Proof. We prove the theorem by a double induction over $r$ and $-\omega$. Clearly, the theorem holds for $r=0$ and for $\omega=0$. Assume therefore that $r>0,-\omega>0$ and that we have proved the theorem for all smaller $r$ and all smaller $-\omega$ with the same $r$. By
lemma 4.2(a), there exists a narrowing $x_{0}$ of $x$, such that the potential dominant terms $c_{1} e^{-\mu_{1} x_{0}}, \ldots, c_{m} e^{-\mu_{m} x_{0}}$ of solutions to (4.1) are in $\mathcal{S}^{x_{0}}$. Moreover, by lemma 4.2(b) if there exists only one such potential dominant term of multiplicity $d$, then we may assume the multiplicator of this narrowing to be 1 .

Now consider the refinement $f=c_{1} e^{-\mu_{1} x}+\tilde{f}_{1} \quad\left(v_{\tilde{f}_{1}}>p \mu_{1}\right)$. The Newton degree of the resulting asymptotic Ricatti equation in $\tilde{f}_{1}$ is $d_{1}$, by lemmas $4.2(\mathrm{c})$ and 4.3 . We have either $d_{1}<d$, or $p=1$ and $-p \mu_{1}<-\omega$. In both cases, the induction hypothesis implies that there exists a narrowing $x_{1}$ of $x_{0}$, such that there exist $d_{1}$ linearly independent solutions $h_{1,1}, \ldots, h_{1, d_{1}}$ to (1.1) in $\mathcal{E}_{d_{1}-1}^{x_{1}}$, which correspond to solutions to the asymptotic Ricatti equation in $\tilde{f}_{1}$.

Similarly, for $i$ running from 2 to $m$, assume that we are given a narrowing $x_{i-1}$ of $x_{1}$ and consider the refinement $f=c_{i} e^{-\mu_{i} x}+\tilde{f}_{i} \quad\left(v_{\tilde{f}_{i}}>p \mu_{i}\right)$. The Newton degree of the resulting asymptotic Ricatti equation in $\tilde{f}_{i}$ is $d_{i}<d$. Hence, by the induction hypothesis, there exists a narrowing $x_{i}$ of $x_{i-1}$, relative to which there exist $d_{i}$ linearly independent solutions $h_{i, 1}, \ldots, h_{i, d_{i}}$ to (1.1) in $\mathcal{E}_{d_{i}-1}^{x_{i}}$, which correspond to solutions to the asymptotic Ricatti equation in $\tilde{f}_{i}$.

Finally, by the second corollary of theorem 4.1, there exists a narrowing $\tilde{x}$ of $x_{m}$, such that (1.1) admits $d_{0}$ linearly independent solutions $h_{0,1}, \ldots, h_{0, d_{0}}$ in $\mathcal{E}_{d_{0}-1,0}^{\tilde{x}}$, whose logarithmic derivatives are solutions to (4.1). By construction, solutions $h_{i, j}$ and $h_{i^{\prime}, j^{\prime}}$ necessarily belong to different direct summands of $\mathcal{E}_{d-1}^{\tilde{x}}$ for $i^{\prime} \neq i$. Hence the $h_{i, j}$ are linearly independent. By lemma 4.1 , we have $d=d_{0}+\cdots+d_{m}$, which concludes the proof of the theorem.

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