# AN ANALOGUE OF HILBERT'S THEOREM 90 FOR THE RING OF ENTIRE FUNCTIONS 

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## 1. Introduction

Let us look at the following situation.
Let $\lambda \in C^{q}$ and $H(\lambda)$ be an $n$ by $n$ matrix whose entries are entire in $\lambda$. Regard $\lambda$ as a row vector and suppose that there exists a linear transformation $L: C^{q} \rightarrow C^{q}$ such that
(I) $L^{m}$ is the identity transformation,
(II) $H$ satisfies the condition

$$
\begin{equation*}
H(\lambda) H(\lambda L) H\left(\lambda L^{2}\right) \cdots H\left(\lambda L^{m-1}\right)=I \tag{1.1}
\end{equation*}
$$

where $I$ is the $n$ by $n$ identity matrix.
To compare with this situation, we shall state Theorem 90 of Hilbert.
Theorem 90 of Hilbert. Let $k$ be a field and $K$ a finite cyclic extension of $k$. Denote by $g$ a generator of $\operatorname{Gal}(K / k)$. Then

$$
\begin{equation*}
N_{K / k}(a)\left(=a g(a) g^{2}(a) \ldots g^{m-1}(a)\right)=1 \tag{1.2}
\end{equation*}
$$

for an element $a$ of $K$ if and only if

$$
\begin{equation*}
a=b g(b)^{-1} \tag{1.3}
\end{equation*}
$$

for some $b \in K^{*}(=K-\{0\})$, where $m$ is the order of $\operatorname{Gal}(K / k)(c f . D$. Hilbert [6]).

[^0]As to (1.1), a result analogous to this theorem is known.
Proposition 1.1. Suppose that (I) is satisfied. Furthermore assume that (III) $L^{h}$ is not the identity transformation for $0<h<m$.

Let $H$ be an $n$ by $n$ matrix whose entries are meromorphic in $\lambda$. Then $H$ satisfies condition (1.1) if and only if

$$
\begin{equation*}
H(\lambda)=W(\lambda) W(\lambda L)^{-1} \tag{1.4}
\end{equation*}
$$

for some matrix $W$ whose entries are meromorphic in $\lambda$.
In this paper we shall prove a better result:
Theorem 1.2. Assume that L satisfies condition (I). Let $H(\lambda)$ be an $n$ by $n$ matrix whose entries are entire in $\lambda$. Then $H$ satisfies condition (1.1) if and only $i f$
(i) $H(0)^{m}=L$; and there exists an $n$ by $n$ matrix $E(\lambda)$ such that
(ii) the entries of $E$ and $E^{-1}$ are entire in $\lambda$,
(iii) $H(\lambda)=E(\lambda) H(0) E(\lambda L)^{-1}$.

Note. In this theorem we do not assume (III) of Proposition 1.1.
The hardest part of the problem is to construct the matrix $E$ so that the entries of $E$ and $E^{-1}$ are entire in $\lambda$. If $q=1$ (i.e., $\lambda$ is a scalar) and if (III) is also satisfied, we can construct such an $E$ by using the $W$ of Proposition 1.1 and Weierstrass' representation of an entire function as an ifinite product. Such a technique does not work for other cases. To overcome this difficulty we use the following result:

Proposition 1.3. Let $(t, \lambda)=\left(t, \lambda_{1}, \ldots, \lambda_{q}\right) \in C^{q+1}$ and $F(t, \lambda)$ be an $n$ by $n$ matrix whose entires are entire in $(t, \lambda)$. Assume that there exists a linear transformation $L: C^{q} \rightarrow C^{q}$ such that
(i) $L^{m}$ is the identity transformation,
(ii) $F(t, \lambda) F(t, \lambda L) F\left(t, \lambda L^{2}\right) \ldots F\left(t, \lambda L^{m-1}\right)=I$.

Then, there exists an $n$ by $n$ matrix $A(t, \lambda)$ such that
(1) the entries of $A(t, \lambda)$ are entire in $(t, \lambda)$,
(2) the derivative of $F$ with respect to $t$ is given by

$$
\begin{equation*}
(\partial F / \partial t)(t, \lambda)=A(t, \lambda) F(t, \lambda)-F(t, \lambda) A(t, \lambda L) \tag{1.5}
\end{equation*}
$$

In fact, if we set $F(t, \lambda)=H(t \lambda)$, then condition (1.1) implies condition (ii) of Proposition 1.3. Now, define a matrix $E$ by

$$
\begin{equation*}
d E / d t=A(t, \lambda) E, \quad E=I \text { at } t=0 \tag{1.6}
\end{equation*}
$$

Then, the entries of $E$ and $E^{-1}$ are entire in $(t, \lambda)$ and

$$
\begin{equation*}
H(t \lambda)=F(t, \lambda)=E(t, \lambda) F(0, \lambda) E(t, \lambda L)^{-1} \tag{1.7}
\end{equation*}
$$

Setting $t=1$, we derive (iii) of Theorem 1.2 with $E=E(1, \lambda)$.

## 2. Proof of Proposition 1.3

In case $m=2$ (i.e., $L^{2}$ is the identity), we have $F(t, \lambda) F(t, \lambda L)=I$, and hence

$$
\begin{equation*}
K(t, \lambda) F(t, \lambda L)+F(t, \lambda) K(t, \lambda L)=0 \tag{2.1}
\end{equation*}
$$

where $K=\partial F / \partial t$. Set

$$
\begin{equation*}
B(t, \lambda)=K(t, \lambda) F(t, \lambda L)=K(t, \lambda) F(t, \lambda)^{-1} \tag{2.2}
\end{equation*}
$$

Then

$$
B(t, \lambda L)=K(t, \lambda L) F(t, \lambda)=-F(t, \lambda L) K(t, \lambda)
$$

Therefore

$$
K(t, \lambda)=B(t, \lambda) F(t, \lambda)=-F(t, \lambda) B(t, \lambda L)
$$

and

$$
\begin{equation*}
K(t, \lambda)=\frac{1}{2}[B(t, \lambda) F(t, \lambda)-F(t, \lambda) B(t, \lambda L)] . \tag{2.3}
\end{equation*}
$$

Hence, setting

$$
\begin{equation*}
A(t, \lambda)=\frac{1}{2} B(t, \lambda)\left(=\frac{1}{2} K(t, \lambda) F(t, \lambda)^{-1}\right) \tag{2.4}
\end{equation*}
$$

we derive (1.5). We can prove Proposition 1.3 for the general case in a similar manner. However, we shall provide here a much simpler proof.

Set

$$
\begin{align*}
& \omega=\exp [2 \pi i / m], \rho\left(\omega^{k}\right)=L^{k} \quad(k=0,1, \ldots, m-1) \\
& F\left(t, \lambda ; \omega^{0}\right)=I, F(t, \lambda ; \omega)=F(t, \lambda)  \tag{2.5}\\
& F\left(t, \lambda ; \omega^{k}\right)=F(t, \lambda) F(t, \lambda L) \ldots F\left(t, \lambda L^{k-1}\right) \\
& \quad(k=2, \ldots, m-1)
\end{align*}
$$

Then

$$
\begin{align*}
& F\left(t, \lambda ; \omega^{k+h}\right)=F\left(t, \lambda ; \omega^{k}\right) F\left(t, \lambda \rho\left(\omega^{k}\right) ; \omega^{h}\right)  \tag{2.6}\\
& \quad(k, h=0,1, \ldots, m-1)
\end{align*}
$$

Set

$$
\begin{equation*}
D\left(t, \lambda ; \omega^{k}\right)=(\partial F / \partial t)\left(t, \lambda ; \omega^{k}\right) F\left(t, \lambda ; \omega^{k}\right)^{-1} \tag{2.7}
\end{equation*}
$$

Then
(2.8) $D\left(t, \lambda ; \omega^{k+h}\right)$

$$
=D\left(t, \lambda ; \omega^{k}\right)+F\left(t, \lambda ; \omega^{k}\right) D\left(t, \lambda \rho\left(\omega^{k}\right) ; \omega^{h}\right) F\left(t, \lambda ; \omega^{k}\right)^{-1}
$$

Therefore, if we define $A$ by

$$
\begin{equation*}
A(t, \lambda)=\frac{1}{n} \sum_{h=0}^{m-1} D\left(t, \lambda ; \omega^{h}\right) \tag{2.9}
\end{equation*}
$$

we have

$$
A(t, \lambda)=D\left(t, \lambda ; \omega^{k}\right)+F\left(t, \lambda ; \omega^{k}\right) A\left(t, \lambda \rho\left(\omega^{k}\right)\right) F\left(t, \lambda ; \omega^{k}\right)^{-1}
$$

or

$$
\begin{align*}
(\partial F / \partial t)\left(t, \lambda ; \omega^{k}\right)= & A(t, \lambda) F\left(t, \lambda ; \omega^{k}\right)  \tag{2.10}\\
& -F\left(t, \lambda ; \omega^{k}\right) A\left(t, \lambda \rho\left(\omega^{k}\right)\right)
\end{align*}
$$

Setting $k=1$, we derive (1.5).

## 3. A generalization of Proposition 1.3 and Theorem 1.2

We can generalize Proposition 1.3 further.
Theorem 3.1. Let $t \in C, \lambda \in C^{q}$ and $\xi \in G$, where $G$ is a compact group. Let $\rho: G \rightarrow G L_{q}(C)$ be a continuous map. Suppose that

$$
\begin{equation*}
F: C \times C^{q} \times G \rightarrow G L_{n}(C) \tag{3.1}
\end{equation*}
$$

is a continuous map such that
(i) the entries of matrix $F(t, \lambda ; \xi)$ are entire in $(t, \lambda)$ for each fixed $\xi \in G$,
(ii) $F$ satisfies the condition

$$
\begin{equation*}
F(t, \lambda ; \xi \eta)=F(t, \lambda ; \xi) F(t, \lambda \rho(\xi) ; \eta) \tag{3.2}
\end{equation*}
$$

for $t \in C, \lambda \in C^{q}, \xi \in G$ and $\eta \in G$.
Then, there exists an $n$ by $n$ matrix $A(t, \lambda)$ such that
(i) the entries of $A$ are entire in $(t, \lambda)$,
(ii) the derivative of $F$ with respect to $t$ is given by
(3.3) $(\partial F / \partial t)(t, \lambda ; \xi)=A(t, \lambda) F(t, \lambda ; \xi)-F(t, \lambda ; \xi) A(t, \lambda \rho(\xi))$.

In fact, if $\nu(\xi)$ is the normalized Haar measure on $G$ with $\nu(G)=1$, then $A$ is given by

$$
\begin{equation*}
A(t, \lambda)=\int_{G} D(t, \lambda ; \eta) d \nu(\eta) \tag{3.4}
\end{equation*}
$$

where

$$
\begin{equation*}
D(t, \lambda ; \xi)=(\partial F / \partial t)(t, \lambda ; \xi) F(t, \lambda ; \xi)^{-1} \tag{3.5}
\end{equation*}
$$

Note that (3.2) implies

$$
\begin{equation*}
D(t, \lambda ; \xi \eta)=D(t, \lambda ; \xi)+F(t, \lambda ; \xi) D(t, \lambda \rho(\xi) ; \eta) F(t, \lambda ; \xi)^{-1} \tag{3.6}
\end{equation*}
$$

Let us define an $n$ by $n$ matrix $E(t, \lambda)$ by

$$
\begin{equation*}
d E / d t=A(t, \lambda) E, \quad E=I \quad \text { at } t=0 \tag{3.7}
\end{equation*}
$$

Then, (3.3) implies that

$$
\begin{equation*}
F(t, \lambda ; \xi)=E(t, \lambda) F(0, \lambda ; \xi) E(t, \lambda \rho(\xi))^{-1} \tag{3.8}
\end{equation*}
$$

or

$$
F(t, \lambda ; \xi) E(t, \lambda \rho(\xi)) F(0, \lambda ; \xi)^{-1}=E(t, \lambda)
$$

Note that the entries of $E$ and $E^{-1}$ are entire in $(t, \lambda)$. Thus we have proved the following theorem:

Theorem 3.2. Under the same assumptions as Theorem 3.1, there exists an $n$ by $n$ matrix $E(t, \lambda)$ such that
(1) the entries of $E$ and $E^{-1}$ are entire in $(t, \lambda)$,
(2) $F$ has the form (3.8), i.e.,

$$
F(t, \lambda ; \xi)=E(t, \lambda) F(0, \lambda ; \xi) E(t, \lambda \rho(\xi))^{-1}
$$

In the case when $F$ does not depend on $t$, introducing $t$ through the change of the variable replacing $\lambda$ by $t \lambda$, we can prove the following theorem:

Theorem 3.3. Let $\lambda \in C^{q}$ and $\xi \in G$, where $G$ is a compact group. Let $\rho: G \rightarrow G L_{q}(C)$ be a continuous map. Suppose that

$$
\Phi: C^{q} \times G \rightarrow G L_{n}(C)
$$

is a continuous map such that
(i) the entries of $\Phi$ are entire in $\lambda$ for each fixed $\xi \in G$,
(ii) $\Phi$ satisfies the condition

$$
\Phi(\lambda ; \xi \eta)=\Phi(\lambda ; \xi) \Phi(\lambda \rho(\xi) ; \eta)
$$

for $\lambda \in C^{q}, \xi \in G$ and $\eta \in G$.
Then there exists an $n$ by $n$ matrix $E(\lambda)$ such that
(1) the entries of $E$ and $E^{-1}$ are entire in $\lambda$,
(2) $\Phi$ has the form $\Phi(\lambda ; \xi)=E(\lambda) \Phi(0 ; \xi) E(\lambda \rho(\xi))^{-1}$.

Remark 3.4. (i) In Theorem 3.1, it is not necessary to assume that

$$
\begin{equation*}
\rho(\xi \eta)=\rho(\xi) \rho(\eta) \text { for } \xi \text { and } \eta \in G \tag{3.9}
\end{equation*}
$$

However, condition (3.2) implies that

$$
F(t, \lambda \rho(\xi \eta) ; \zeta)=F(t, \lambda \rho(\xi) \rho(\eta) ; \zeta) \quad \text { for } \xi, \eta \text { and } \zeta \in G
$$

Hence, it would be convenient to verify (3.9) when we want to check condition (3.2).
(ii) Condition (3.2) can be relaxed in the following way:

The quantity

$$
C=F(t, \lambda ; \xi \eta)^{-1} F(t, \lambda ; \xi) F(t, \lambda \rho(\xi) ; \eta)
$$

is independent of $t$.
In fact, (3.2') also implies (3.6).
(iii) In Theorem 1.2, Condition (1.1) can be replaced by

$$
\begin{equation*}
H(\lambda) H(\lambda L) H\left(\lambda L^{2}\right) \ldots H\left(\lambda L^{m-1}\right)=K \tag{1.1'}
\end{equation*}
$$

where $K$ is an $n$ by $n$ invertible constant matrix.
In fact, utilizing (2.5) we can verify that

$$
F\left(\lambda ; \omega^{k+h}\right)^{-1} F\left(\lambda ; \omega^{k}\right) F\left(\lambda L^{k} ; \omega^{h}\right)=\left\{\begin{array}{lll}
I & \text { if } & 2 \leq k+h \leq n \\
K & \text { if } & n+1 \leq k+h \leq 2 n
\end{array}\right.
$$

Hence we can apply Remark (ii). (Condition (i) of Theorem 1.2 should be replaced by $H(0)^{m}=K$.)
(iv) The requirement on the smoothness of the entries of $F(t, \lambda ; \xi)$ may be relaxed. For example, we may assume that the entries of $F$ and $\partial F / \partial t$ are continuous in $(t, \lambda ; \xi)$ in a domain: $l \times \mathscr{U} \times G$, where $l$ is a $t$-interval and $\mathscr{U}$ is an open set in the $\lambda$-space. Then the entries of $A$ are also continuous in $(t, \lambda)$. If we assume a continuous differentiability of the entries of $F$ and $\partial F / \partial t$ with respect to $\lambda$, then the entries of $A$ admit the same kind of smoothness. Those modifications are based on the observation concerning the differentiation of an integral of the type $\int_{G} g d \nu(\xi)$ with a function $g$ which is smooth with respect to parameters.
(v) Theoem 3.1 can be extended to $p$-adic functions if $G$ is finite. However, we do not know whether the matrix $E$ defined by (3.7) is entire.
(vi) The matrices $\Phi$ and $E$ of Theorem 3.3 satisfy condition (2) or

$$
\begin{equation*}
E(\lambda)=\Phi(\lambda ; \xi) E(\lambda \rho(\xi)) \Phi(0: \xi)^{-1} \tag{3.10}
\end{equation*}
$$

We can interpret this relation in terms of automorphy factors and automorphic forms (cf. A. Borel [3]). In fact, if we define a map

$$
\begin{equation*}
\alpha: C^{q} \times G \rightarrow \operatorname{Aut}\left(M_{n}(C)\right) \tag{3.11}
\end{equation*}
$$

by

$$
\begin{equation*}
\alpha(\lambda ; \xi)[X]=\Phi(\lambda ; \xi) X \Phi(0 ; \xi)^{-1} \tag{3.12}
\end{equation*}
$$

where $M_{n}(C)$ is the vector space of $n$ by $n$ complex matrices, and $\lambda \in C^{q}$, $\xi \in G$ and $X \in M_{n}(C)$, then, since $\Phi$ satisfies the relation $\Phi(\lambda ; \xi \eta)=$ $\Phi(\lambda ; \xi) \Phi(\lambda \rho(\xi) ; \eta)$, we have

$$
\begin{equation*}
\alpha(\lambda ; \xi \eta)=\alpha(\lambda ; \xi) \alpha(\lambda \rho(\xi) ; \eta) \tag{3.13}
\end{equation*}
$$

and relation (3.10) can be written in the form

$$
\begin{equation*}
\alpha(\lambda: \xi)[E(\lambda \rho(\xi))]=E(\lambda) \tag{3.14}
\end{equation*}
$$

This mans that $\alpha$ is an automorphy factor and that $E$ is an automorphic form relative to $\alpha$. For a given automorphy factor, automorphic functions are not unique. One of the most important problems in the study of automorphic functions is to find a basis for the space of automorphic functions. We shall investigate such a problem concerning the matrix $E$ of Theorems 3.2 and 3.3, elsewhere.

Note. In the case of Theorem 3.2, we define $\alpha$ by

$$
\alpha(t, \lambda ; \xi)[X]=F(t, \lambda ; \xi) X F(0, \lambda ; \xi)^{-1}
$$

Then condition (3.2) implies that

$$
\begin{equation*}
\alpha(t ; \lambda ; \xi \eta)=\alpha(t, \lambda ; \xi) \alpha(t, \lambda \rho(\xi) ; \eta) \tag{3.13'}
\end{equation*}
$$

and relation ( $3.8^{\prime}$ ) can be written in the form

$$
\alpha(t, \lambda ; \xi)[E(t, \lambda \rho(\xi))]=E(t, \lambda)
$$

Hence, $\alpha$ is an automorphy factor and $E$ is an automorphic form relative to $\alpha$.

## 4. Results in terms of Galois-cohomology

Let $X$ be a smooth complex anaytic manifold, $E$ a vector bundle over $X$, and $G$ a compact group acting continuously on the right on $E$ by vector bundle automorphisms; i.e., there is given a homomorphism $G^{0} \rightarrow \operatorname{Aut}(E)$. Let $\mathscr{R}$ denote $\Gamma\left(E, \mathcal{O}_{E}\right)$ (the ring of all global analytic functions on $E$ ) so that $G$ acts on $\mathscr{R}$, i.e., there is a homomorphism $\rho: G \rightarrow \operatorname{Aut}(\mathscr{R})$. Thus $G$ also acts on $G L_{n}(\mathscr{R})$. Now we can state an abstract version of Theorem 3.3 in terms of Galois-cohomology (cf. A. Grothendieck [5]).

Theorem 4.1. $\quad H^{1}\left(G, G L_{n}(\mathscr{R})\right)=H^{1}\left(G, G L_{n}\left(\Gamma\left(X, \mathcal{O}_{X}\right)\right)\right)$.
Note. In case of Theorem 3.3, $X$ consists of a point.
To prove Theorem 4.1, we introduce a new variable $t$ through $(t, e) \rightarrow t e$ $(e \in E)$. This replaces $E$ by $C \times E$. Let $\Gamma\left(C \times E, \mathcal{O}_{C \times E}\right)$ be denoted by $\mathscr{R}^{\prime}$. Then $G$ still acts on $C \times E$, and hence $\rho$ can be extended to a homomorphism $\rho^{\prime}: G \rightarrow \operatorname{Aut}\left(\mathscr{R}^{\prime}\right)$. Further if we denote $d / d t$ by $\mathscr{D}$, then we have $\mathscr{D} \rho^{\prime}(\xi)=$ $\rho^{\prime}(\xi) \mathscr{D}$ for $\xi \in G$. Let $\mathscr{A}=G L_{n}\left(\mathscr{R}^{\prime}\right)$ and $\mathscr{F}=\mathscr{M}\left(\mathscr{R}^{\prime}\right)$. Now we can state an abstract version of Theorem 3.1.

Theorem 4.2. For every one cocycle $f: G \rightarrow \mathscr{A}$ there exists $b \in \mathscr{F}$ such that

$$
\begin{equation*}
\mathscr{D}(f(\xi))=b f(\xi)-f(\xi) \rho^{\prime}(\xi)(b) \tag{4.1}
\end{equation*}
$$

for all $\xi \in G$.
Note. A map $f: G \rightarrow \mathscr{A}$ is a one cocycle if

$$
f(\xi \eta)=f(\xi) \rho^{\prime}(\xi)(f(\eta)) \quad \text { for } \xi \in G \text { and } \eta \in G
$$

To prove Theorem 4.2, we introduce on $\mathscr{F}$ a new structure of $G$-module by $(\xi, b) \rightarrow \xi \times b=f(\xi) \rho^{\prime}(\xi)(b) f(\xi)^{-1}$. Let us denote this new $G$-module by $\mathscr{F}_{f}$. Set

$$
\begin{equation*}
h(\xi)=\mathscr{D}(f(\xi)) f(\xi)^{-1} \tag{4.2}
\end{equation*}
$$

This is a map $G \rightarrow \mathscr{F}_{f}$. It is easily verified that

$$
\begin{equation*}
h(\xi \eta)=h(\xi)+\xi \times h(\eta) \tag{4.3}
\end{equation*}
$$

This means that $h$ is a one cocycle and hence defines an element in $H^{1}\left(G, \mathscr{F}_{f}\right)$.
Now the following lemma is the key to the proof of Theorem 4.2.
Lemma 4.3. $\quad H^{1}\left(G, \mathscr{F}_{f}\right)=\{0\}$.
To prove this, set

$$
b=\int_{G} h(\eta) d \nu(\eta)
$$

where $\nu$ is the normalized Haar measure on $G$ with $\nu(G)=1$. Then from (4.3) we derive

$$
\begin{equation*}
b=h(\xi)+\xi \times b \tag{4.4}
\end{equation*}
$$

Hence Lemma 4.3 follows immediately.

## 5. An example

A traditional method for solving differential equations which goes back to Riemann's treatment of hypergeometric differential equation has 5 stages:
(1) classification of differential equations by means of suitable transformations;
(2) identification of invariants under such transformations in terms of solutions;
(3) construction of a standard equation representing an equivalence class in terms of its invariants;
(4) computation of invariants for a given equation;
(5) reduction of a given equation to a standard equation.

Since 1970 a German-American School (W. Jurkat (Ulm-Syracuse), D.A. Lutz (Milwaukee-San Diego), W. Balser (Ulm) et al) has done extensive work on the classification of meromorphic differential equations (cf. W. Jurkat [8]). In particular they identified the invariants in terms of monodromy matrices and Stokes multipliers. Inspired by the German-American group, a French School (B. Malgrange (Grenoble), J.-P. Ramis (Strasbourg) et al) described the space of invariants in terms of certain cohomology groups related with differential equations.

Now, "computation of invariants" has become a point of interest. Such a computation may be carried out in many ways. This problem is essentially
related with the computation of monodromy matrices and Stokes multipliers. These quantities may be computed numerically, if a differential equation is given. We are interested in studying these quantities as functions of suitable parameters. In a local study, this leads us to a perturbation theory (regular and/or singular: cf. W. Balser [2]), or a deformation theory such as recent work on isomonodromic deformations (cf. Flaschka-Newell [4], Jimbo-MiwaUeno [7], T. Kimura [10], K. Okamoto [12] and Römer-Schröder [13]) and isoformal deformations (cf. Babbitt-Varadalajan [1]). A motivation of our researches is to study some special but important cases (cf. Y. Sibuya [14]).

Precisely speaking we study solutions of

$$
\begin{equation*}
\left(\delta^{2}-p \delta\right) y-P(x) y=0 \quad(\delta=x d / d x) \tag{5.1}
\end{equation*}
$$

where

$$
\begin{equation*}
P(x)=x^{m}+\sum_{h=1}^{m-1} a_{h} x^{m-h} \tag{5.2}
\end{equation*}
$$

$p$ is an integer such that $0 \leq p \leq m-1$, and $a_{h}(h=1, \ldots, m-1)$ are parameters. We assume that $x=0$ is an apparent singular point. This assumption implies that $a_{m-p}$ is a certain polynomial in other parameters $a_{h}$ ( $h \neq m-p$ ) for each pair $(m, p)$ : for example,

$$
\begin{align*}
& a_{m-1}=0(p=1), \quad a_{m-2}=a_{m-1}^{2}(p=2)  \tag{5.3}\\
& a_{m-3}=a_{m-1} a_{m-2}-\frac{1}{4} a_{m-1}^{3}(p=3), \text { etc. }
\end{align*}
$$

We shall denote by $a$ the vector $\left(a_{1}, \ldots, a_{m-p-1}, a_{m-p+1}, \ldots, a_{m-1}\right) \in C^{m-2}$. If $p=1$, Equation (5.1) becomes

$$
\begin{equation*}
d^{2} y / d x^{2}-Q(x) y=0 \tag{5.4}
\end{equation*}
$$

where

$$
\begin{equation*}
Q(x)=x^{m-2}+\sum_{h=1}^{m-2} a_{h} x^{m-h-2} \tag{5.5}
\end{equation*}
$$

Asymptotic solutions of Equation (5.4) with (5.5) were studied in Y. Sibuya [14]. Many results in this book can be extended to equation (5.1).

Proposition 5.1. There exist two linearly independent solutions of equation (5.1):

$$
\begin{align*}
& \varphi_{1}(x, a)=1+\sum_{h=1}^{p-1} \varphi_{1 h}(a) x^{h}+\sum_{h=p+1}^{\infty} \varphi_{1 h}(a) x^{h}  \tag{5.6}\\
& \varphi_{2}(x, a)=x^{p}+\sum_{h=p+1}^{\infty} \varphi_{2 h}(a) x^{h}
\end{align*}
$$

which are unique and entire in $(x, a)$.

This result is an application of the method of G. Frobenius to (5.1) at $x=0$. Note that $x=0$ is an apparent singular point.

Proposition 5.2. Equation (5.1) admits a solution $\varphi(x, a)$ such that
(i) $\varphi$ is entire in $(x, a)$,
(ii) $\varphi$ admits an asymptotic representation

$$
\begin{equation*}
\varphi=x^{-b(a)+(2 p-m) / 4}\left[1+O\left(x^{-1 / 2}\right)\right] \exp [-E(x, a)] \tag{5.7}
\end{equation*}
$$

as $x \rightarrow \infty$ in $|\arg x|<3 \pi / m$, where

$$
\begin{gathered}
{\left[1+\sum_{k=1}^{m-1} a_{k} x^{-k}\right]^{1 / 2}=1+\sum_{h=1}^{\infty} b_{h}(a) x^{-h}} \\
E(x, a)=(2 / m) x^{m / 2}+\sum_{1 \leq h<m / 2}(2 /(m-2 h)) b_{h}(a) x^{(m-2 h) / 2}
\end{gathered}
$$

and

$$
b(a)= \begin{cases}0 & \text { if } m \text { is odd } \\ b_{m / 2}(a) & \text { if } m \text { is even }\end{cases}
$$

Condition (ii) determines the solution $\varphi$ uniquely.
This is a simple modification of Theorem 6.1 of Y. Sibuya [14] (see also F.E. Mullin [11]).

Set

$$
\begin{equation*}
\omega=\exp (2 \pi i / m), \quad G(a)=\left(\omega a_{1}, \ldots, \omega^{k} a_{k}, \ldots, \omega^{m-1} a_{m-1}\right) \tag{5.8}
\end{equation*}
$$

and

$$
\begin{equation*}
f_{k}(x, a)=\varphi\left(\omega^{-k} x, G^{-k}(a)\right) \tag{5.9}
\end{equation*}
$$

Proposition 5.3. For every integer $k, f_{k}$ is a solution of equation (5.1) and

$$
\begin{align*}
f_{k}= & \left(\omega^{-k} x\right)^{-(-1)^{k} b(a)+(2 p-m) / 4}\left[1+O\left(x^{-1 / 2}\right)\right] \\
& \times \exp \left[-(-1)^{k} E(x, a)\right] \\
\delta f_{k}= & \left(\omega^{-k} x\right)^{-(-1)^{k} b(a)+(2 p+m) / 4}\left[-1+O\left(x^{-1 / 2}\right)\right]  \tag{5.10}\\
& \times \exp \left[-(-1)^{k} E(x, a)\right]
\end{align*}
$$

as $x \rightarrow \infty$ in $|\arg x-2 \pi k / m|<3 \pi / m$.

This result follows from (5.9), Proposition 5.2 and

$$
\begin{equation*}
b_{k}(G(a))=\omega^{k} b_{k}(a) \quad(k=1,2, \ldots) \tag{5.11}
\end{equation*}
$$

Proposition 5.4. For the solutions $\varphi_{1}$ and $\varphi_{2}$ we have

$$
\begin{equation*}
\varphi_{1}(\omega x, G(a))=\varphi_{1}(x, a), \quad \varphi_{2}(\omega x, G(a))=\omega^{p} \varphi_{2}(x, a) \tag{5.12}
\end{equation*}
$$

This result is a consequence of the uniqueness of $\varphi_{1}$ and $\varphi_{2}$.
Set

$$
\Phi_{k}(x, a)=\left[\begin{array}{cc}
f_{k} & f_{k+1}  \tag{5.13}\\
\delta f_{k} & \delta f_{k+1}
\end{array}\right], \quad W_{k}(x, a)=\operatorname{det} \Phi_{k}(x, a)
$$

Proposition 5.5. For every integer $k$, we have

$$
\begin{equation*}
\Phi_{k}(x, a)=\Phi_{0}\left(\omega^{-k} x, G^{-k}(a)\right) \tag{5.14}
\end{equation*}
$$

and

$$
\begin{equation*}
W_{k}(x, a)=2 x^{p} \omega^{-(-1)^{k} b(a)-k p-(2 p-m) / 4} . \tag{5.15}
\end{equation*}
$$

This result follows from (5.11) and Proposition 5.3.
Set

$$
\begin{equation*}
\Phi_{k}(x, a)=\Phi_{k+1}(x, a) S_{k}(a) \tag{5.16}
\end{equation*}
$$

The matrices $S_{k}$ are called Stokes multipliers.
Proposition 5.6. For every integer $k$, we have

$$
S_{k}(a)=\left[\begin{array}{ll}
C_{k}(a) & 1  \tag{5.17}\\
\tilde{C}_{k}(a) & 0
\end{array}\right]
$$

where

$$
\begin{gather*}
C_{k}(a)=\left(W_{k+1}\right)^{-1}\left[\begin{array}{cc}
f_{k} & f_{k+2} \\
\delta f_{k} & \delta f_{k+2}
\end{array}\right]  \tag{5.18}\\
\tilde{C}_{k}(a)=-\left(W_{k+1}\right)^{-1} W_{k}=-\omega^{-(-1)^{k} 2 b(a)+p} \tag{5.19}
\end{gather*}
$$

and

$$
\begin{equation*}
S_{k}(a)=S_{0}\left(G^{-k}(a)\right) \tag{5.20}
\end{equation*}
$$

These results can be verified by simple computations.
Proposition 5.7. The matrices $S_{k}(k=0, \ldots, m-1)$ satisfy the relation

$$
\begin{equation*}
S_{m-1}(a) S_{m-2}(a) \ldots S_{2}(a) S_{1}(a) S_{0}(a)=I \tag{5.21}
\end{equation*}
$$

where $I$ is the 2 by 2 identity matrix.
This result follows from the fact that $\Phi_{m+k}=\Phi_{k}$ (cf. (5.14)). Note that the monodromy group of equation (5.1) is trivial, since $x=0$ is an apparent singular point.

We are interested in the meaning of relation (5.21). Utilizing (5.20) we can write (5.21) in the form

$$
\begin{equation*}
S_{0}\left(G^{-m+1}(a)\right) S_{0}\left(G^{-m+2}(a)\right) \ldots S_{0}\left(G^{-1}(a)\right) S_{0}(a)=I \tag{5.22}
\end{equation*}
$$

or

$$
\begin{equation*}
S_{0}(a) S_{0}(G(a)) \ldots S_{0}\left(G^{m-2}(a)\right) S_{0}\left(G^{m-1}(a)\right)=I \tag{5.23}
\end{equation*}
$$

Theorem 1.2 applies to (5.23). Hence, there exists a 2 by 2 matrix $E(a)$ such that
(i) the entries of $E$ and $E^{-1}$ are entire in $a$,
(ii) $S_{0}$ has the form

$$
\begin{equation*}
S_{0}(a)=E(a) S_{0}(0) E(G(a))^{-1} \tag{5.24}
\end{equation*}
$$

On the other hand, let us look at relation (5.16). Setting $k=0$ and utilizing (5.14) we derive

$$
\Phi_{0}(x, a)=\Phi_{0}\left(\omega^{-1} x, G^{-1}(a)\right) S_{0}(a)
$$

or

$$
\begin{equation*}
S_{0}(a)=\Phi_{0}\left(\omega^{-1} x, G^{-1}(a)\right)^{-1} \Phi_{0}(x, a) \tag{5.25}
\end{equation*}
$$

Utilizing the two linearly independent solutions $\varphi_{1}$ and $\varphi_{2}$ (cf. Proposition 5.1), we write the two solutions $f_{0}$ and $f_{1}$ as linear combinations of $\varphi_{1}$ and $\varphi_{2}$, i.e.,

$$
\begin{equation*}
\Phi_{0}(x, a)=\Phi(x, a) \Gamma(a) \tag{5.26}
\end{equation*}
$$

where

$$
\Phi(x, a)=\left[\begin{array}{cc}
\varphi_{1} & \varphi_{2} \\
\delta \varphi_{1} & \delta \varphi_{2}
\end{array}\right]
$$

and the entries of $\Gamma$ and $\Gamma^{-1}$ are entire in $a$. The matrix $\Gamma(a)$ is called a central connection matrix. Note that

$$
\Phi_{0}\left(\omega^{-1} x, G^{-1}(a)\right)=\Phi(x, a)\left[\begin{array}{cc}
1 & 0  \tag{5.27}\\
0 & \omega^{-p}
\end{array}\right] \Gamma\left(G^{-1}(a)\right)
$$

(cf. (5.12)). Therefore, from (5.25), (5.26) and (5.27), we derive

$$
S_{0}(a)=\Gamma\left(G^{-1}(a)\right)^{-1}\left[\begin{array}{cc}
1 & 0  \tag{5.28}\\
0 & \omega^{p}
\end{array}\right] \Gamma(a)
$$

and hence

$$
S_{0}(0)=\Gamma(0)^{-1}\left[\begin{array}{cc}
1 & 0  \tag{5.29}\\
0 & \omega^{p}
\end{array}\right] \Gamma(0)
$$

Thus we have

$$
\begin{equation*}
S_{0}(a)=F\left(G^{-1}(a)\right)^{-1} S_{0}(0) F(a) \tag{5.30}
\end{equation*}
$$

where

$$
\begin{equation*}
F(a)=\Gamma(0)^{-1} \Gamma(a) \tag{5.31}
\end{equation*}
$$

(Note that the entries of $F$ and $F^{-1}$ are entire in a.) This means that the matrix $E$ of (5.24) and the matrix $\tilde{F}=F\left(G^{-1}(a)\right)^{-1}$ satisfy the same relation with $S_{0}$ (i.e., (5.24) and (5.30)). However, $E$ and $\tilde{F}$ were constructed through two totally different processes. We shall investigate the relation between $E$ and $\tilde{F}$ more carefully elsewhere. For doing this, we hope that it would be helpful to regard Stokes multipliers $S_{k}$ as an automorphy factor and $E$ and $\tilde{F}$ as associated automorphic forms (cf. Remark 3.4 (vi)).

We strongly believe that the study of Stokes multipliers as functions of suitably chosen parameters will lead us to a theory similar to that of automorphy factors and automorphic functions (cf., also, Jurkat-Zwiesler [9]).

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[^0]:    Received January 9, 1987.
    ${ }^{1}$ Supported in part by grants from the National Science Foundation.
    ${ }^{2}$ Supported in part by grants from the National Science Foundation and Alexander von Humboldt-Stiftung.

