# Publications mathématiques de l'I.H.É.S. 

# Marc Chaperon <br> $C^{k}$-conjugacy of holomorphic flows near a singularity 

Publications mathématiques de l'I.H.É.S., tome 64 (1986), p. 143-183
[http://www.numdam.org/item?id=PMIHES_1986__64__143_0](http://www.numdam.org/item?id=PMIHES_1986__64__143_0)
© Publications mathématiques de l'I.H.É.S., 1986, tous droits réservés.
L'accès aux archives de la revue « Publications mathématiques de l'I.H.É.S. » (http:// www.ihes.fr/IHES/Publications/Publications.html) implique l'accord avec les conditions générales d'utilisation (http://www.numdam.org/conditions). Toute utilisation commerciale ou impression systématique est constitutive d'une infraction pénale. Toute copie ou impression de ce fichier doit contenir la présente mention de copyright.

# $C^{k}$-CONJUGACY OF HOLOMORPHIC FLOWS <br> NEAR A SINGULARITY 

by Marc CHAPERON

## TABLE OF CONTENTS

I. Introduction ..... 144
(1.1) Notation and definitions ..... 144
(1.2) Holomorphic classification: known results and obstacles ..... 144
(1.3) $\mathrm{C}^{k}$-classification: definitions and main results ..... 145
(1.4) Plan of the article ..... 146
Aknowledgements ..... 147
2. General idea and main local tools of our proofs ..... 147
(2.1) Introduction: the real case ..... 147
(2.2) Proof of Sternberg's theorem ..... 148
(2.3) General idea of our proofs ..... 154
3. The complex flow of a weakly hyperbolic $S$ ..... 155
(3.1) Introduction and notation ..... 155
(3.2) First properties ..... I 55
(3.3) Structure of the flow $\left(\Phi_{c}^{l}\right)$ ..... ${ }^{1} 56$
(3.4) Concluding remarks ..... 160
4. Non-linear complex flows-proof of theorem 2 ..... 160
(4.1) Hypotheses and notation ..... 160
(4.2) " Isolating blocks" for the complex flow of S near the origin ..... 161
(4.3) " Isolating blocks" for perturbations of the complex flow of S near o ..... 163
(4.4) Conjugacy results; proof of Theorem 2 ..... 168
5. Proof of theorem $I$ ..... 172
(5.1) Algebraic background: normal forms and their formal flows ..... 172
(5.2) A generalisation of Theorem 1 ..... 175
(5.3) Proof of Theorem 3 ..... - 76
6. Concluding remarks ..... I 79
(6.1) $\mathrm{C}^{k}$-linearisations ..... 179
(6.2) Families ..... 180
(6.3) Historical and technical comments ..... 181
Main notation ..... 181
Definitions and main results ..... 182

## 1. INTRODUCTION

## (1.1) Notation and definitions

Let $n$ denote a fixed positive integer, and let $\mathbf{d}$ be the Lie algebra of those germs at $o \in \mathbf{C}^{n}$ of holomorphic vector fields which vanish at the origin. Call two elements $X$ and $Y$ of $\mathbf{d}$ conjugate if there exists a germ $h:\left(\mathbf{C}^{n}, o\right)$ ) of a holomorphic diffeomorphism such that $h^{*} \mathrm{Y}=\mathrm{X}$. For each $\mathrm{X} \in \mathbf{d}$, let $\mathrm{X}^{\mathbf{1}}$ stand for the linear part $d \mathbf{X}(\mathbf{0}) \in \mathfrak{g l}(n, \mathbf{C})$ of X .

Throughout the sequel, we denote by S a diagonalisable element of $\mathfrak{g l}(n, \mathbf{C})$. An S-vector field is an element $X$ of $\mathbf{d}$ such that $S$ is the semi-simple part of $X^{1}$ (thus, every $\mathbf{X} \in \mathbf{d}$ is an $S$-vector field for a unique $S$ ). An S-normal form is an S-vector field of the form ( ${ }^{1}$ ) $S+N$, where $N$ is polynomial and commutes whith $S$ in $\mathbf{d}$ (since $S+N$ is an S -vector field, the linear part $\mathrm{N}^{1}$ is nilpotent and commutes with S ).

## (1.2) Holomorphic classification: known results and obstacles

Define $S$ to be in the Poincaré domain when the convex hull of its spectrum in $\mathbf{C}$ does not contain the origin.

Theorem (Poincaré-Dulac). - If S is in the Poincaré domain, then
(i) the centraliser of $\mathbf{S}$ in $\mathbf{d}$ is finite dimensional and consists of polynomial vector fields-in particular, the degree of an S-normal form cannot be arbitrarily high;
(ii) every S-vector field is conjugate to an S-normal form.

The reader is referred to (5.1) below for a proof of (i). The "preparation lemma" we shall state in section 2 is a natural generalisation of (ii) to all elements of $\mathbf{d}$.

The Poincaré-Dulac theorem provides very good models: if $S$ is in the Poincaré domain, then (see (5.1))
a) every S-normal form generates a holomorphic $\mathbf{C}$-action on $\mathbf{C}^{n}$, given by an explicit formula;
b) if S lies outside an explicitly known closed subset of codimension one, then the only S -normal form is S itself (in this case, the above result is Poincarés linearisation theorem).
$\left.{ }^{( }{ }^{1}\right)$ We shall not distinguish between polynomial vector fields and their germs at 0.

The Siegel domain (i.e. the complementary subset of the Poincare domain in $\mathfrak{g l}(n, \mathbf{C})$ ) contains a full measure subset $\mathscr{S}$ with the following property: if S lies in $\mathscr{S}$, then every S -vector field is conjugate to S (this is the Siegel linearisation theorem). Here are some reasons why this non-trivial, remarkable result is not quite as satisfactory as the previous one:
I) To show that a given S belongs to $\mathscr{S}$, one should check infinitely many in-equalities-which might take some time in general.
2) The complementary subset $\mathscr{R}$ of $\mathscr{S}$ in the Siegel domain is dense, and even quite large. Moreover, the few known facts about the conjugacy classes of S-vector fields with $S \in \mathscr{R}$ show that there is little hope for simple results in that direction: for example, for a given $S$, the space of all formal conjugacy classes of $S$-vector fields may be really huge (and a given formal conjugacy class may contain an infinite dimensional space of conjugacy classes [MR]).

We shall prove that, most of the time (in a very simple sense), these pathological phenomena disappear if holomorphic conjugacy is replaced by $\mathrm{C}^{k}$-conjugacy, $k \in \mathbf{N}$, defined in (1.3) below. This viewpoint will prove especially useful when dealing with families of vector fields-see our final remarks.

Our methods are based upon a rather thorough geometric understanding of the complex flows under study (sections 3 and 4): even in cases when the Siegel theorem holds true, this is an addition to our knowledge of the subject, allowing one to estimate the extent to which two different (holomorphic) conjugacy classes are geometrically different.

## (1.3) $\mathrm{C}^{\boldsymbol{k}}$-conjugacy: definitions and main results

Given $X \in \mathbf{d}$, recall that a representative of $X$ is a holomorphic vector field $\widetilde{X}$ on some open neighbourhood dom $\tilde{X}$ of o in $\mathbf{C}^{n}$, such that $X$ is the germ of $\tilde{X}$ at o-in other words, germ $X$, viewed as a convergent power series, is the Taylor expansion of $\tilde{X}$ at $o$. The foliation defined by $X$ is the foliation of (dom $\tilde{X}) \backslash \tilde{X}^{-1}$ (o) by holomorphic curves everywhere tangent to $\tilde{\mathbb{X}}$. The germ of this foliation at $o$ depends only on X , and will be denoted by $\mathscr{F} \mathrm{X}$.

For each $k \in \mathbf{N}$, call two elements X and Y of $\mathbf{d}$

- $\mathrm{C}^{k}$-equivalent if there exists a germ of a $\mathbf{C}^{k}$-diffeomorphism ( $\mathbf{C}^{n}, \mathrm{o}$ ) 2 (viewing $\mathbf{C}^{n}$ as $\mathbf{R}^{2 n}$ ) sending $\mathscr{F} \mathrm{X}$ onto $\mathscr{F} \mathrm{Y}$;
- $\mathrm{C}^{k}$-conjugate if they admit representatives $\widetilde{\mathrm{X}}$ and $\widetilde{\mathrm{Y}}$ respectively with the following property: there exists a $\mathrm{C}^{k}$-diffeomorphism $h: \operatorname{dom} \widetilde{\mathrm{X}} \rightarrow \operatorname{dom} \widetilde{\mathrm{Y}}$ such that, for every $v \in \operatorname{dom} \widetilde{\mathbf{X}}$, the image $h \circ c$ of the local integral curve $c:(\mathbf{C}, o) \rightarrow\left(\mathbf{C}^{n}, v\right)$ of $\widetilde{\mathbf{X}}$ is the local integral curve $(\mathbf{C}, \mathrm{o}) \rightarrow\left(\mathbf{C}^{n}, h(v)\right)$ of $\widetilde{\mathrm{Y}}$ (in the language of [Ch 86] (p. 68-70), the $\mathbf{C}$-action germs generated by X and Y are $\mathrm{C}^{k}$-isomorphic).

Clearly, conjugacy implies $\mathrm{C}^{k}$-conjugacy, which implies $\mathrm{C}^{k}$-equivalence. Call S

- hyperbolic if its eigenvalues are simple and any two of them are $\mathbf{R}$-independant;
- weakly hyperbolic if the closed line segment between two of its eigenvalues never contains o (thus, if S is hyperbolic, or in the Poincaré domain, it is weakly hyperbolic).

In section 5, we shall state and prove a more general version of the following result [Ch 8o]:

Theorem 1. - If S is hyperbolic, every S -vector field is $\mathrm{C}^{0}$-conjugate to S .
Now, Guckenheimer [G], Camacho, Kuiper, Palis [CKP] and Ladis [I] were able to determine the $\mathrm{C}^{0}$-equivalence class of a hyperbolic S inside $\mathfrak{g l}(n, \mathbf{C})$, hence

Corollary 1. - Let S and $T$ be hyperbolic elements of $\mathfrak{g l}(n, \mathbf{C})$, and let $\mathrm{Spec}^{-1} \mathrm{~S}$ (resp. $\mathrm{Spec}^{-1} \mathrm{~T}$ ) denote the set of all inverses of eigenvalues of $\mathrm{S}($ resp. T$)$.
(i) If S and T are in the Poincaré domain, then any $\mathrm{S}-\mathrm{vector}$ field is $\mathrm{C}^{\mathbf{0}}$-equivalent to any T-vector field.
(ii) In the remaining case, the following two conditions are equivalent:

- Any S-vector field is $\mathrm{C}^{0}$-equivalent to any T -vector field.
- There exists $\mathrm{A} \in \mathrm{GL}(2, \mathbf{R})$ such that $\mathrm{A}\left(\operatorname{Spec}^{-1} \mathrm{~S}\right)=\operatorname{Spec}^{-1} \mathrm{~T}$.

In the Siegel domain, this result-stated as a conjecture in [CKP]-exhibits a rigidity phenomenon which makes Theorem I much more surprising than the GrobmanHartman linearisation theorem (see [A] or [Ch 86]), despite superficial analogy. For this very reason, the proof of Theorem 1 is hard and uses the following complex analogue of a theorem of Sternberg stated in (2.I) below:

Theorem 2. - If S is weakly hyperbolic, then, for every positive integer $k$, each S-vector field is $\mathbf{C}^{k}$-conjugate to an S -normal form ( ${ }^{1}$ ).

This statement is not quite as simple as Theorem 1 -however, see Theorem 4 in the Conclusion-but its much nicer proof is very instructive.

## (1.4) Plan of the article

In section 2, the real version of Theorem 2 is used as a good introduction to the general idea of our proofs (paragraph (2.3)), a good excuse for stating our main local lemmas (the Isolating Block Lemma and the Extension lemma in (2.2), the Preparation Lemma in (2.3)) and a good reason for introducing the basic notion of a strongly invariant manifold (paragraph (2.3)).

In section 3, we explain the (global) structure of the complex flow generated by a weakly hyperbolic $S$, and discuss some further applications of our analysis. In

[^0]section 4 , we show that, if $S$ is weakly hyperbolic, an $S$-vector field of the form provided by the Preparation Lemma can really be considered a small perturbation of $S$ near $o$, and prove Theorem 2.

In Section 5, we state and prove the generalisation of Theorem I already mentioned.
In the conclusion, we discuss some related problems and results.

## Aknowledgements.

Like most of what I have done in mathematics, this work owes much to the influence of René Thom and the advice of Alain Chenciner and Michael R. Herman. It is also a pleasure to thank Cesar Camacho, Freddy Dumortier, Xavier Gómez-Mont, André Haefliger, Ju. S. Il'iasenko, Nicolaas Kuiper, Ivan Kupka, François Laudenbach, Santiago López de Medrano, Robert Moussu, Jacob Palis and Robert Roussaric, whose-sceptical or confident-interest in this research was quite stimulating.

## 2. GENERAL IDEA AND MAIN LOCAL TOOLS OF OUR PROOFS

## (2. 1) Introduction: the real case

In this paragraph, we denote by R a diagonalisable (over $\mathbf{C}$ ) element of $\mathfrak{g l}(n, \mathbf{R})$. Define a real R -vector field and a real R -normal form as in ( $\mathrm{I} . \mathrm{I}$ ), replacing $\mathbf{d}$ by the Lie algebra of those germs at $o \in \mathbf{R}^{n}$ of $\mathbf{C}^{\infty}$-vector fields which vanish at 0 .

The stable subspace $\mathrm{E}^{+}$of R and its unstable subspace $\mathrm{E}^{-}$are defined as follows: $\mathrm{E}^{+}$is the unstable subspace of -R , and $\mathrm{E}^{-}$is the sum of those R -invariant subspaces of $\mathbf{R}^{n}$ corresponding to eigenvalues $c$ with $\mathscr{R} e c>0$, i.e. the set of those $v \in \mathbf{R}^{n}$ such that $\lim _{t \rightarrow-\infty} e^{i \mathrm{R}} v=0$. For this last reason, a real R -normal form is tangent to $\mathrm{E}^{-}$ (and to $\mathrm{E}^{+}$) at each of its points (see [Ch 86], p. I4r, Lemme-an alternative, silly proof would follow from the calculations in (5.r) below). The following result is a (classical) particular case of [Ch $86(4 \cdot 4 \cdot 2 b)$, Théorème I]:

Preparation Lemma. - For every $k \in \mathbf{N}$, each real $\mathbf{R}$-vector field is $\mathbf{C}^{\infty}$-conjugate (in the usual sense) to a real R -vector field X which has $k$-th order contact with a real R -normal form along $\mathrm{E}^{+} \cup \mathrm{E}^{-}$-in particular, by the above remark, X is tangent to $\mathrm{E}^{-}$(and to $\mathrm{E}^{+}$) " at each of its points".
(In fact, the result proven in [Ch 86] is that one can take the same X for every $k$. The advantage of our weaker statement is that its proof can lead to effective computations.)

Call R (real) hyperbolic if none of its eigenvalues lies on the imaginary axis, i.e. $\mathrm{E}^{+} \oplus \mathrm{E}^{-}=\mathbf{R}^{n}$ (thus, R need not be hyperbolic as an element of $\mathfrak{g l}(n, \mathbf{C})$ ). Here comes our real version of Theorem 2:

Theorem (Sternberg). - If R is (real) hyperbolic, then, for each positive integer $k$, every real R -vector field is $\mathrm{C}^{k}$-conjugate to a real R -normal form.

## (2.2) Proof of Sternberg's theorem

We shall first state two key lemmas, in the general form needed later on.
Notation. - Let Q denote the (riemannian) product $\mathrm{M} \times \mathrm{E}^{+} \times \mathrm{E}^{-}$, where M is a compact riemannian manifold and $\mathrm{E}^{+}, \mathrm{E}^{-}$denote two euclidean spaces with $\mathrm{E}^{+} \neq\{0\}$. Let $\mathrm{W}^{+}, \mathrm{W}^{-}$and $\Sigma$ be the three submanifolds of Q defined by

$$
\begin{aligned}
& \mathrm{W}^{+}=\mathrm{M} \times \mathrm{E}^{+} \times\{0\}, \quad \mathrm{W}^{-}=\mathrm{M} \times\{0\} \times \mathrm{E}^{-} \\
& \mathrm{\Sigma}=\mathrm{W}^{+} \cap \mathrm{W}^{-}=\mathrm{M} \times\{0\}\left(^{1}\right) .
\end{aligned}
$$

The canonical (orthogonal) projections of Q onto $\mathrm{E}^{+}$and $\mathrm{E}^{-}$will be denoted by $x \mapsto x_{+}$ and $x \mapsto x_{-}$respectively, and the euclidean norms will be written $v \mapsto|v|$. We let (see Fig. 1)

$$
\left\{\begin{array}{l}
\mathrm{B}=\left\{x \in \mathrm{Q}:\left|x_{+}\right| \leqslant \mathrm{I} \text { and }\left|x_{-}\right| \leqslant \mathrm{I}\right\} \\
\partial^{+} \mathrm{B}=\left\{x \in \mathrm{~B}:\left|x_{+}\right|=\mathrm{I}\right\}, \quad \partial^{-} \mathrm{B}=\left\{x \in \mathrm{~B}:\left|x_{-}\right|=\mathrm{I}\right\} \\
\mathrm{W}_{1}^{+}=\mathrm{W}^{+} \cap \mathrm{B}, \quad \mathrm{~W}_{1}^{-}=\mathrm{W}^{-} \cap \mathrm{B}
\end{array}\right.
$$

hence in particular $\partial \mathrm{B}=\partial^{+} \mathrm{B} \cup \partial^{-} \mathrm{B}$. Denote the scalar products of $\mathrm{E}^{+}$and $\mathrm{E}^{-}$ by $(x, y) \mapsto(x \mid y)$, and the differential $\mathrm{TQ} \rightarrow \mathrm{E}^{ \pm}$of the projection $x \mapsto x_{ \pm}$, by $v \mapsto v_{ \pm}$.

Isolating Block Lemma (Fig. 1). - Let $\xi$ be a smooth vector field on some open neighbourhood of B in Q . Assume that there exist positive constants $k_{+}$and $k_{-}$such that every $x \in \mathrm{~B}$ satisfies

$$
\begin{equation*}
\left(\xi(x)_{+} \mid x_{+}\right) \leqslant-k_{+}\left|x_{+}\right|^{2} \quad \text { and } \quad\left(\xi(x)_{-} \mid x_{-}\right) \geqslant k_{-}\left|x_{-}\right|^{2} . \tag{I}
\end{equation*}
$$

Then, the flow $\left(\Phi^{t}\right)$ generated by $\xi$ has the following properties:
(i) The two functions $r_{+}, r_{-}: \mathbf{B} \rightarrow[0, \infty]$ defined by

$$
\left\{\begin{array}{l}
r_{+}(x)=\sup \left\{t \geqslant 0: \Phi^{-s}(x) \in \mathbf{B} \text { for } 0 \leqslant s \leqslant t\right\} \\
r_{-}(x)=\sup \left\{t \geqslant 0: \Phi^{s}(x) \in \mathbf{B} \quad \text { for } 0 \leqslant s \leqslant t\right\}
\end{array}\right.
$$

satisfy

$$
\left\{\begin{array}{l}
r_{+}^{-1}(\infty)=\mathrm{W}_{1}^{-}, \quad r_{-}^{-1}(\infty)=\mathrm{W}_{1}^{+} \\
\Phi^{-r_{+}(x)}(x) \in \partial^{+} \mathrm{B} \text { for } x \in \mathrm{~B} \backslash \mathrm{~W}^{-}, \quad \Phi^{r_{-}(x)}(x) \in \partial^{-} \mathrm{B} \text { for } x \in \mathrm{~B} \backslash \mathrm{~W}^{+}
\end{array}\right.
$$

(ii) The two smooth functions $x \mapsto-r_{+}(x) / \log \left|x_{+}\right|$and $x \mapsto-r_{-}(x) / \log \left|x_{-}\right|$, defined on $\mathrm{B} \backslash\left(\mathrm{W}^{-} \cup \partial^{+} \mathrm{B}\right)$ and $\mathrm{B} \backslash\left(\mathrm{W}^{+} \cup \partial^{-} \mathrm{B}\right)$ respectively, are bounded and have positive lower bounds.
(iii) If $\mathscr{B}$ is a basis for the filter of neighbourhoods of $\partial \mathrm{W}_{1}^{+}$in $\partial^{+} \mathrm{B}$, then the sets $\tilde{\mathrm{V}}=\left\{\Phi^{t}(x): 0 \leqslant t \leqslant r_{-}(x)\right.$ and $\left.x \in \mathrm{~V}\right\} \cup \mathrm{W}_{1}^{-}$with $\mathrm{V} \in \mathscr{B}$ form a basis $\widetilde{\mathscr{B}}$ for the filter of neighbourhoods of $\mathrm{W}_{1}^{+} \cup \mathrm{W}_{1}^{-}$in B .

[^1]

Fig. I
Proof. - There exist positive constants $\mathrm{K}_{+}$and $\mathrm{K}_{-}$such that, for every $x \in \mathrm{~B}$, the following estimates are satisfied:

$$
\left\{\begin{array}{l}
-\mathrm{K}_{+}\left|x_{+}\right|^{2} \leqslant\left(\xi(x)_{+} \mid x_{+}\right) \leqslant-k_{+}\left|x_{+}\right|^{2} \\
k_{-}\left|x_{-}\right|^{2} \leqslant\left(\xi(x)_{-} \mid x_{-}\right) \leqslant \mathrm{K}_{-}\left|x_{-}\right|^{2} .
\end{array}\right.
$$

"Integrating" these inequations (see [Ch 86], p. 362, Lemme), we get

$$
\left\{\begin{array}{l}
e^{k_{+} t}\left|x_{+}\right| \leqslant\left|\Phi^{-t}(x)_{+}\right| \leqslant e^{\mathrm{K}_{+} t}\left|x_{+}\right| \quad \text { for } 0 \leqslant t \leqslant r_{+}(x)  \tag{2}\\
e^{k_{-} t}\left|x_{-}\right| \leqslant\left|\Phi^{t}(x)_{-}\right| \leqslant e^{\mathrm{K} t}\left|x_{-}\right| \quad \text { for } 0 \leqslant t \leqslant r_{-}(x) .
\end{array}\right.
$$

From this and the (therefore) obvious fact that each function $t \rightarrow\left|\Phi^{-t}(x)_{+}\right|, 0 \leqslant t \leqslant r_{+}(x)$ and $t \mapsto\left|\Phi^{t}(x)_{-}\right|, \quad 0 \leqslant t \leqslant r_{-}(x)$, is either zero, or strictly increasing, assertions (i) and (ii) follow at once.

Assertion (iii) is obtained as follows: for each $x \in \partial^{+} B$, the greatest distance $\Delta_{x}$ between $\Phi^{t}(x)$ and $\mathrm{W}_{1}^{+} \cup \mathrm{W}_{1}^{-}$for $0 \leqslant t \leqslant r_{-}(x)$ can be estimated quite easily: either $x \in \mathrm{~W}_{1}^{+}$, in which case $\Delta_{x}=0$, or, by the growth properties of $t \mapsto\left|\Phi^{-t}(x)_{-}\right|$and $t \mapsto\left|\Phi^{t}(x)_{+}\right|$,

$$
\Delta_{x}=\left|\Phi^{t_{x}(x)_{+}}\right|=\left|\Phi^{t_{x}(x)_{-}}\right|
$$

where $t_{x} \in\left[0, r_{-}(x)\right]$ is defined by this equality. Therefore, (2) yields

$$
\begin{equation*}
\Delta_{x} \leqslant\left|x_{-}\right|^{\mid k_{4}\left(\mathbb{K}_{+}+K_{-}\right)} . \tag{3}
\end{equation*}
$$

Similarly, the least distance $\delta_{x}$ between $\Phi^{t}(x)$ and $\mathrm{W}_{1}^{+} \cup \mathrm{W}_{1}^{-}$for $0 \leqslant t \leqslant r_{-}(x)$ is either zero, or $\min \left\{\left|x_{-}\right|,\left|\Phi^{r_{-}(x)}(x)_{+}\right|\right\}$, hence, by (2),

$$
\begin{equation*}
\delta_{x} \geqslant \min \left\{\left|x_{-}\right|,\left|x_{-}\right|^{\mid \mathrm{K}_{1} / k_{-}}\right\} . \tag{4}
\end{equation*}
$$

Assertion (iii) follows at once from (3) and (4).
Note. - The name of our lemma comes from the fact that B is an isolating block for ( $\Phi^{t}$ ) in the sense of Conley (see for example [CZ]).

Notation and definition. - For each smooth vector field X on a manifold (with corners) C , we let $f_{\mathrm{X}}$ denote the flow of X (defined on a subset $\operatorname{dom} f_{\mathrm{X}}$ of $\mathbf{R} \times \mathbf{C}$ ), and sometimes write $f_{\mathrm{X}}^{t}(x)$ instead of $f_{\mathrm{X}}(t, x)$ for $(t, x) \in \operatorname{dom} f_{\mathrm{x}}$. A subset U of $\mathbf{C}$ is called X -saturated if $\left\{t: f_{\mathrm{X}}^{t}(x) \in \mathrm{U}\right\}$ is an interval for every $x \in \mathrm{U}$.

Extension lemma. - For each $\xi$ as in the Isolating Block Lemma and each positive integer $k$, there exist an integer $m \geqslant k$ and a $\mathrm{C}^{1}$-neighbourhood $\mathfrak{N}$ of $\left.\xi\right|_{\mathrm{B}}$ in the space of smooth vector fields on B such that the following hold true:
(i) If $\mathscr{N}_{0}$ denotes the set of those $\mathrm{X} \in \mathscr{N}$ which are tangent to $\mathrm{W}_{1}^{+}$and to $\mathrm{W}_{1}^{-}$at each of its points, then, Hypothesis (1) of the Isolating Block Lemma is satisfied (with different positive constants) if $\left.\xi\right|_{\mathrm{B}}$ is replaced by any $\mathrm{X} \in \mathscr{N}_{0}$.
(ii) Let X and Y be two arbitrary elements of $\mathcal{N}_{0}$, having m-th order contact along $\mathrm{W}_{1}^{+} \cup \mathrm{W}_{1}^{-}$, and let $h: \omega \rightarrow \omega^{\prime}$ be a $C^{m}$-diffeomorphism with the following properties:
a) $\omega$ is the X -saturated intersection of an open neighbourhood of $\Sigma$ in Q and an open neighbourhood of $\mathrm{W}_{1}^{+} \backslash \Sigma$ in $\mathrm{B} \backslash\left(\mathrm{W}_{1}^{-} \cup \partial^{-} \mathrm{B}\right)$;
b) $\omega^{\prime}$ is Y -saturated, open in B , and $h_{*}\left(\left.\mathrm{X}\right|_{\omega}\right)=\left.\mathrm{Y}\right|_{\omega^{\prime}}$;
c) $h$ has $m$-th order contact with the identity along $\mathrm{W}^{+} \cap \omega$.

Then, the set

$$
\begin{aligned}
\Omega=\left\{f_{\mathrm{X}}^{t}(x): x \in \omega, t \geqslant\right. & \mathrm{o},(t, x) \in \operatorname{dom} f_{\mathrm{X}} \\
& \left.\quad \text { and }(t, h(x)) \in \operatorname{dom} f_{\mathrm{Y}} \backslash\left(f_{\mathrm{Y}}\right)^{-1}\left(\partial^{-} \mathrm{B}\right)\right\} \cup \mathrm{W}_{1}^{-} \backslash \partial^{-} \mathrm{B}
\end{aligned}
$$

is an X -saturated open subset of B , and $h$ extends to a unique $\mathrm{C}^{k}$-diffeomorphism H of $\Omega$ onto the Y-saturated open subset

$$
\begin{aligned}
& \Omega^{\prime}=\left\{f_{\mathrm{Y}}^{\prime}(y): y \in \omega^{\prime}, t \geqslant 0,(t, y) \in \operatorname{dom} f_{\mathrm{Y}}\right. \\
& \left.\quad \text { and }\left(t, h^{-1}(y)\right) \in \operatorname{dom} f_{\mathrm{X}} \backslash\left(f_{\mathrm{X}}\right)^{-1}\left(\partial^{-} \mathrm{B}\right)\right\} \cup \mathrm{W}_{\mathbf{1}}^{-} \backslash \partial^{-} \mathrm{B}
\end{aligned}
$$

such that $\mathrm{H}_{*}\left(\left.\mathrm{X}\right|_{\Omega}\right)=\left.\mathrm{Y}\right|_{\mathbf{\Omega}^{\prime}} . \quad$ Moreover, H has $k$-th order contact with the identity along $\mathrm{W}^{+} \cup \mathrm{W}^{-}$.
Proof. - (i) is clear, and so is the fact that $\Omega \backslash \mathrm{W}^{-}$(resp. $\Omega^{\prime} \backslash \mathrm{W}^{-}$) is an X-saturated (resp. Y-saturated) open subset of B-whatever $m$ may be. Moreover, still assuming $m \geqslant k$ arbitrary, $\left.\mathrm{H}\right|_{\Omega W^{W}}$ is uniquely determined by the conjugacy condition, which implies that it has to be the $\mathrm{C}^{m}$-diffeomorphism onto $\Omega^{\prime} \backslash \mathrm{W}^{-}$defined by $\mathrm{H}\left(f_{\mathrm{X}}^{\prime}(x)\right)=f_{\mathrm{Y}}^{\prime}(h(x))$ for $t \geqslant 0, x \in \omega$ and $f_{\mathrm{X}}^{\prime}(x) \in \Omega$. Therefore, uniqueness comes
from the following fact: if $\Omega$ is a neighbourhood of $\mathrm{W}_{1}^{-} \backslash^{-} \mathrm{B}$ in B , then, as $\mathrm{E}^{+}$is not trivial, $\left.\mathrm{H}\right|_{\Omega \backslash W}$ has at most one continuous extension to $\Omega$. Thus, we just have to prove

Lemma 1. - If $m$ is large enough and $\mathcal{N}$ small enough, then, under the hypotheses of (ii), there exists an open subset U of B with $\mathrm{\Sigma} \subset \mathrm{U} \subset \Omega$ such that the mapping

$$
\mathrm{U} \ni v \mapsto \begin{cases}\mathrm{H}(v) & \text { if } v \notin \mathrm{~W}^{-} \\ v & \text { if } v \in \mathrm{~W}^{-}\end{cases}
$$

is of class $\mathrm{C}^{k}$ and has $k$-th order contact with Id along $\mathrm{W}^{-}$.
Indeed, Lemma i implies our result: if its conclusion is true, then, by the analogue of (2) for X and Y (which implies that $\mathrm{W}_{1}^{-}$is their unstable manifold at $\Sigma$ ) and the fact that they coincide up to order $k$ along $\mathrm{W}_{1}^{-}$, we shall have $\mathrm{H}\left(f_{\mathbf{X}}^{\prime}(v)\right)=f_{\mathbf{Y}}^{\prime}(\mathrm{H}(v))$ for $t \geqslant 0, v \in \mathrm{U}$ and $f_{\mathrm{X}}^{\prime}(v) \in \mathrm{B} \backslash \partial^{-} \mathrm{B}$ and, by the flow-box theorem, $\Omega$ will be a neighbourhood of $\mathrm{W}_{1}^{-} \backslash \partial^{-} \mathrm{B}$ and H will have $k$-th order contact with Id along $\mathrm{W}_{1}^{-} \backslash \partial^{-} \mathrm{B}$.

We shall give a rather tricky proof of Lemma 1 , again adapted from [Ch 86] ((4.2.3), théorème 2). Of course, we exclude the case when $\mathrm{E}^{-}$-and, therefore, Lemma I -is trivial. Let $u \in \mathbf{C}^{\infty}(\mathbf{R},[\mathrm{O}, \mathrm{I}])$ satisfy $u^{-1}(\mathrm{I})=(-\infty, \mathrm{I} / 3]$ and $u^{-1}(0)=[2 / 3, \infty)$, and let $\mathrm{X}_{0}, \widetilde{\mathrm{X}}, \widetilde{\mathrm{Y}}$ be the smooth vector fields on Q defined by

$$
\begin{cases}\mathrm{X}_{0}(x)=\left(0,-x_{+}, x_{-}\right) \\ \tilde{\mathrm{X}}(x)= \begin{cases}\mathbf{X}_{0}(x)+u\left(\left|x_{+}\right|\right) u\left(\left|x_{-}\right|\right)\left(\mathrm{X}(x)-\mathrm{X}_{0}(x)\right) & \text { if } x \in \mathrm{~B} \\ \mathrm{X}_{0}(x)=\widetilde{\mathrm{Y}}(x) & \text { if } x \notin \mathrm{~B}\end{cases} \\ \widetilde{\mathrm{Y}}(x)=\mathrm{X}_{0}(x)+u\left(\left|x_{+}\right|\right) u\left(\left|x_{-}\right|\right)\left(\mathrm{Y}(x)-\mathrm{X}_{0}(x)\right) & \text { if } x \in \mathrm{~B} .\end{cases}
$$

As $\mathrm{X}_{0}$ generates an $\mathbf{R}$-action on Q , so do $\widetilde{\mathrm{X}}$ and $\widetilde{\mathrm{Y}}$.
Step 1. - If $\mathcal{N}$ has been chosen small enough, then
(i) $\widetilde{\mathrm{X}}$ and $\widetilde{\mathrm{Y}}$ satisfy inequations of type (2) in the whole of Q , and have m-th order contact along $\mathrm{W}^{+} \cup \mathrm{W}^{-}$;
(ii) let $\mathrm{B}^{\prime}=\left\{x \in \mathrm{Q}:\left|x_{-}\right|<\mathrm{I} / 3\right.$ and $\left.\left|x_{+}\right|<\mathrm{I} / 3\right\}$ and $\omega^{\prime \prime}=\mathrm{B}^{\prime} \cap \mathrm{H}^{-1}\left(\mathrm{~B}^{\prime}\right) \backslash \mathrm{W}^{-}$; then, $\omega^{+}=\left\{f_{\tilde{\mathrm{x}}}^{t}(x): x \in \omega^{\prime \prime}\right.$ and $\left.t \leqslant \mathrm{o}\right\}$ is an $\widetilde{\mathrm{X}}$-saturated open neighbourhood of $\mathrm{W}^{+} \backslash \Sigma$ in $\underset{\sim}{\mathrm{Q}} \backslash \mathrm{W}^{-}$, and $\left.\mathrm{H}\right|_{\omega^{\prime \prime}}$ extends to a unique $\mathrm{C}^{m}$-embedding $\widetilde{\mathrm{H}}: \omega^{+} \rightarrow \mathbf{Q}$ such that $\widetilde{\mathrm{H}}_{*}\left(\left.\widetilde{\mathrm{X}}\right|_{\omega^{+}}\right)=\left.\widetilde{\mathrm{Y}}\right|_{\tilde{\mathrm{H}}\left(\omega^{+}\right)}$, having m-th order contact with Id along $\mathrm{W}^{+} \backslash \Sigma$.

Step 2. - Let $\mathrm{W}_{r}=\left\{x \in \mathbf{Q}:\left|x_{-}\right|<r\right\}, r>o$, and let $x \mapsto x_{0}$ denote the projection $\mathrm{Q} \rightarrow \mathrm{M}$. If $r$ is small enough, then
(i) the mapping $\mathrm{H}_{0}: \mathrm{W}_{r} \cap\left(\omega^{+} \cup \mathrm{B}\right) \rightarrow \mathbf{Q}$ defined by

$$
\mathrm{H}_{0}(v)=\left\{\begin{array}{l}
\left(\tilde{\mathrm{H}}\left(v_{0}, v_{+}, u\left(\mathrm{I}-\left|v_{+}\right|\right) v_{-}\right)_{0}, \tilde{\mathrm{H}}(v)_{+}+u\left(\left|v_{+}\right|\right)\left(v_{+}-\tilde{\mathrm{H}}(v)_{+}\right),\right. \\
\left.\quad \tilde{\mathrm{H}}(v)_{-}+u\left(\left|v_{+}\right|\right)\left(v_{-}-\widetilde{\mathrm{H}}(v)_{-}\right)\right) \\
\quad \text { for } \mathrm{I} / 3 \leqslant\left|v_{+}\right| \leqslant 2 / 3 \\
v \quad \text { for }\left|v_{+}\right| \leqslant \mathrm{I} / 3 \\
\tilde{\mathrm{H}}(v) \quad \text { for }\left|v_{+}\right| \geqslant 2 / 3
\end{array}\right.
$$

is a $\mathrm{C}^{m}$-embedding, having m-th order contact with Id along $\mathrm{W}^{+} \cup \mathrm{W}^{-}$;
(ii) for $r$ small enough, the vector field Z on $\mathrm{W}_{r}$ given by

$$
\mathbf{Z}(v)=\mathrm{H}_{0}^{*} \widetilde{\mathrm{Y}}(v) \quad \text { if } v \in \mathbf{B} \quad \text { and } \quad \mathrm{Z}(v)=\mathbf{X}_{\mathbf{0}}(v) \quad \text { if } v \notin \mathbf{B}
$$

is of class $\mathrm{C}^{m-1}$ and has $(m-\mathrm{I})$-th order contact with $\widetilde{\mathrm{X}}$ along $\mathrm{W}^{+} \cup \mathrm{W}^{-}$.
Outside B , we have that $\mathrm{H}_{0}^{-1} \circ \tilde{\mathrm{H}}=\mathrm{Id}$ and $\tilde{\mathrm{X}}=\mathrm{Z}$; therefore, replacing $(\tilde{\mathrm{X}}, \tilde{\mathrm{Y}}, \tilde{\mathrm{H}})$ by $\left(\tilde{\mathrm{X}}, \mathrm{Z}, \mathrm{H}_{0}^{-1} \circ \widetilde{\mathrm{H}}\right)$, we get a more or less standard extension problem, which we shall now solve.

Step 3 ([Ch 86], théorème A6-5, p. 361). Define $\widetilde{\mathbf{X}}_{-}: \mathbf{Q} \rightarrow \mathrm{E}^{-}$by $\widetilde{\mathbf{X}}_{-}(x)=\widetilde{\mathbf{X}}(x)_{-}$, and let

$$
\left\{\begin{array}{l}
\lambda_{\mathrm{x}}=\inf \left\{\left(\left.\frac{\partial \widetilde{\mathrm{X}}_{-}}{\partial x_{-}}(x) \cdot y \right\rvert\, y\right) /|y|^{2}: x \in \mathrm{~W}^{+}, y \in \mathrm{E}^{-} \backslash\{o\}\right\} \\
\mu_{\mathrm{X}}^{+}=\sup \left\{\mathrm{L}_{\tilde{\mathrm{x}}} \mathrm{G}(x)(y, y) / 2|y|^{2}: y \in \mathrm{~T}_{x} \mathrm{Q} \backslash\{0\}, x \in \mathrm{~W}^{+}\right\} \\
\mu_{\mathrm{x}}=\inf \left\{\mathrm{L}_{\tilde{\mathrm{x}}} \mathrm{G}(x)(y, y) / 2|y|^{2}: y \in \mathrm{~T}_{x} \mathrm{Q} \backslash\{0\}, x \in \mathrm{~W}^{+}\right\}
\end{array}\right.
$$

where $\mathrm{L}_{\tilde{\mathrm{x}}} \mathrm{G}$ is the Lie derivative of the riemannian metric of Q with respect to $\widetilde{\mathrm{X}}$. If we have $\lambda_{\mathrm{X}}>\mathrm{o}$ (hence $\mu_{\mathrm{X}}^{+}>0$ ), $\mu_{\overline{\mathrm{x}}}^{-}<\mathrm{o}$ and $m>k+\mathrm{I}+\left(\mu_{\mathrm{X}}^{+}-k \mu_{\mathrm{x}}^{-}\right) / \lambda_{\mathrm{x}}$, then, for $r$ small enough, the family $\left(f_{\mathrm{Z}}^{t} \circ f_{\tilde{\mathrm{X}}} \boldsymbol{t}_{\mathrm{w}_{r}}\right)_{t \geqslant 0}$ is well-defined and converges in the $\mathrm{C}^{k}$ sense to an embedding $\mathrm{H}_{r}$ which has $k$-th order contact with Id along $\mathrm{W}^{+} \cup \mathrm{W}^{-}$.

We can now explain how to choose $m$ and $\mathscr{N}: m$ is the least integer greater than $k+\mathrm{I}+\left(\mu_{\xi}^{+}-k \mu_{\xi}^{-}\right) / \lambda_{\xi}$ (we have $\mu_{\xi}^{-}<0<\lambda_{\xi}$ because $\mathrm{E}^{+}$and $\mathrm{E}^{-}$are nontrivial), and $\mathscr{N}$ is such that each $\mathrm{X} \in \mathscr{N}_{0}$ satisfies (i), $\mu_{\mathrm{x}}<\mathrm{o}<\lambda_{\mathrm{x}}$ and $\boldsymbol{m}>k+\mathrm{I}+\left(\mu_{\mathrm{X}}^{+}-k \mu_{\mathrm{x}}^{-}\right) / \lambda_{\mathrm{x}}$.

Then, under the above hypotheses, the definition of $\mathrm{H}_{r}$ implies that $\mathrm{H}_{r}^{*}\left(\left.\mathrm{Z}\right|_{\mathrm{H}_{r}\left(\mathrm{w}_{r}\right)}\right)=\left.\widetilde{\mathrm{X}}\right|_{\mathrm{w}_{r}}$ and $\mathrm{H}_{r}(v)=v$ for $\left|v_{+}\right| \geqslant \mathrm{I}$-by Step I (i) and the fact that $\mathrm{Z}=\widetilde{\mathrm{X}}=\mathrm{X}_{0}$ in $\left\{x \in \mathrm{~W}_{r}:\left|x_{+}\right| \geqslant \mathrm{I}\right\}$. Therefore, if $\mathrm{D}_{r}$ denotes the X -saturated domain of $\mathrm{H}_{0} \circ \mathrm{H}_{r}=\tilde{\mathrm{H}}_{r}$, we obtain

$$
\left\{\begin{array}{l}
\widetilde{\mathrm{H}}_{r}^{*}\left(\left.\widetilde{\mathrm{Y}}\right|_{\mathrm{H}_{r}\left(\mathrm{D}_{r}\right)}\right)=\left.\widetilde{\mathrm{X}}\right|_{\mathrm{D}_{r}} \\
x \in \mathrm{D}_{r} \quad \text { and } \quad \widetilde{\mathrm{H}}_{r}(x)=\widetilde{\mathrm{H}}(x) \quad \text { for every } x \in \omega^{+} \text {with }\left|x_{+}\right| \text {large enough. }
\end{array}\right.
$$

Thus, by Step I (i), we have that $\tilde{\mathrm{H}}_{r}=\widetilde{\mathrm{H}}$ in $\mathrm{W}_{r} \cap \omega^{+}$, hence in particular

$$
\mathrm{U}=\left\{x \in \mathrm{~W}_{r}:\left|x_{+}\right|<r\right\} \subset \Omega \quad \text { and }\left.\quad \mathrm{H}\right|_{\mathrm{U}}=\left.\widetilde{\mathrm{H}}_{r}\right|_{\mathrm{U}}
$$

if $r$ is small enough for U and $\tilde{\mathrm{H}}_{r}(\mathrm{U})$ to lie in $\mathrm{B}^{\prime}$ and be contained in $\left\{f_{\mathrm{X}}^{t}(\omega): t \geqslant 0\right\} \cup \mathrm{W}^{-}$ (which is a neighbourhood of $\Sigma$ in $Q$ by the Isolating Block Lemma (iii), applied in $\left\{x \in \overline{\mathrm{~W}}_{\mathrm{R}}:\left|x_{+}\right| \leqslant \mathrm{R}\right\}$ for some $\mathrm{R} \leqslant \mathrm{I}$; indeed, by Step I (i) and the definition of $\Omega$, $\left.\mathrm{H}\right|_{\mathrm{U} \cap \Omega}$ has to be well-defined and equal to $\left.\widetilde{\mathrm{H}}_{r}\right|_{\mathrm{U}}$ in the whole domain of definition of $\left.\tilde{H}_{r}\right|_{U}$, i.e. U. This proves Lemma .

We shall now see how to construct $\mathrm{C}^{k}$-conjugacies using the Extension Lemma. The "Cauchy problem method" introduced here will be omnipresent in the sequel:

Corollary 2. - Let $k, m$ and $\mathscr{N}_{0}$ be as in the Extension Lemma. If two elements $\mathbf{X}$ and Y of $\mathscr{N}_{0}$ have m-th order contact along $\mathrm{W}^{+} \cup \mathrm{W}^{-}$, then, they are $\mathrm{C}^{k}$-conjugate in a neighbourhood of $\Sigma$. More precisely, let V be the open subset of $\left(\partial^{+} \mathbf{B} \backslash \partial^{-} \mathbf{B}\right) \times[0, \infty)$ consisting of those $(x, t)$ such that both $f_{\mathrm{X}}^{t}(x)$ and $f_{\mathrm{Y}}^{t}(x)$ are well-defined and lie in $\mathrm{B} \backslash \partial^{-} \mathrm{B}$, and let $\varphi$ and $\psi$ denote the smooth embeddings of V into $\mathbf{B} \backslash \partial^{-} \mathrm{B}$ given by $\varphi(x, t)=f_{\mathrm{X}}^{t}(x)$ and $\psi(x, t)=f_{\mathrm{Y}}^{t}(x)$; then, the hypotheses of the Extension Lemma (ii) are satisfied by $\omega=\varphi(\mathrm{V})$ and $h=\psi \circ \varphi^{-1}$.

Proof. - By the Extension Lemma (i) and the Isolating Block Lemma, $\varphi$ and $\psi$ are smooth embeddings of $V$ into $B \backslash\left(W^{-} \cup \partial^{-} B\right)$, the images of which are open neighbourhoods of $\mathrm{W}_{1}^{+} \backslash \Sigma$. Moreover, as X and Y have $m$-th order contact along $\mathrm{W}_{1}^{+}$ and are tangent to it, $\varphi$ and $\psi$ have $m$-th order contact along $\varphi^{-1}\left(\mathbf{W}^{+}\right)=\psi^{-1}\left(\mathbf{W}^{+}\right)$, hence Corollary 2.

Proof of Sternberg's theorem. - Given a positive integer $k$ and a (real) hyperbolic $\mathbf{R} \in \mathfrak{g l}(n, \mathbf{R})$, with stable subspace $\mathrm{E}^{+}$and unstable subspace $\mathrm{E}^{-}$, we shall prove there exists an integer $m \geqslant k$ with the following property: if a real R -vector field Z has $m$-th order contact with a real R -normal form $\mathrm{R}+\mathrm{N}$ along $\mathrm{E}^{+} \cup \mathrm{E}^{-}$, then Z is $\mathrm{C}^{k}$-conjugate to $\mathrm{R}+\mathrm{N}$.

This and the Preparation Lemma (2.1) clearly imply Sternberg's theorem. Now, there obviously exists a euclidean structure on $\mathbf{R}^{n}$ for which $\mathrm{E}^{+}$and $\mathrm{E}^{-}$are orthogonal, and such that the hypotheses of the Isolating Block Lemma are satisfied with $\xi=\mathrm{R}, \mathrm{Q}=\mathbf{R}^{n}=\mathrm{E}^{+} \oplus \mathrm{E}^{-} \simeq \mathrm{E}^{+} \times \mathrm{E}^{-}, \mathrm{W}^{+}=\mathrm{E}^{+}, \mathrm{W}^{-}=\mathrm{E}^{-}$and $\Sigma=\{\mathrm{o}\}$. Let $\mathscr{N}_{0}, m$ be as in the Extension Lemma (with this choice of $\xi$ ), and let $Z$ be a real R -vector field having $m$-th order contact with a real R -normal form $\mathrm{R}+\mathrm{N}$ along $E^{+} \cup E^{-}$; if $\widetilde{Z}$ denotes a representative of $Z$, there exists (see the end of (4.4) below) $\mathrm{A} \in \mathrm{GL}(n, \mathbf{R})$, with $\mathrm{A}^{*} \mathrm{R}=\mathrm{R}$, such that $\mathrm{X}=\left.\left(\mathrm{A}^{*}(\mathrm{R}+\mathrm{N})\right)\right|_{\mathrm{B}}$ and $\mathrm{Y}=\left.\left(\mathrm{A}^{*} \widetilde{\mathrm{Z}}\right)\right|_{\mathrm{B}}$ satisfy the hypotheses of Corollary 2, hence our result.

Important remark. - In Corollary 2, we obtained a local conjugacy $h$ between X and $Y$ as the solution of the Cauchy Problem " $h=I d$ on $\partial^{+} B \backslash \partial^{-} B$ ", which was wellposed because every flow-line of $X$ (or $Y$ ) which lies outside $W^{-} \cup\left(\partial^{+} B \cap \partial^{-} B\right)$ intersects $\partial^{+} \mathbf{B}$ transversally, at exactly one point. When $\xi$ is a (real) hyperbolic $R$, as in Sternberg's theorem, the essential reason for this is that the cylinder $\mathrm{Q}^{\prime}=\left\{x \in \mathbf{Q}=\mathbf{R}^{n}:\left|x_{+}\right|=\mathrm{I}\right\}$ is "a quotient of the $f_{\mathrm{R}^{-}}$invariant open subset $\mathrm{Q} \backslash \mathrm{W}^{-}$ by $f_{\mathrm{R}}{ }^{\prime \prime}$, meaning that each flow-line of R which is not contained in $\mathrm{W}^{-}$intersects $\mathrm{Q}^{\prime}$ transversally, at precisely one point. Now, such a quotient (of course diffeomorphic to the orbit space of $\left.\left.f_{\mathbf{R}}\right|_{\mathbf{R} \times\left(Q \backslash W^{-}\right)}\right)$can be constructed in many other ways; here is one: for simplicity, assume that R has only real eigenvalues $c_{1}, \ldots, c_{n}$; then, there exists a system $\left(x_{1}, \ldots, x_{n}\right)$ of real linear coordinates on $\mathbf{R}^{n}$ such that R is the gradient of the non-degenerate quadratic form $\mathrm{F}=\Sigma_{c} x_{j}^{2} / 2$ with respect to the euclidean metric $\Sigma d x_{j}^{2}$,
and $\mathrm{Q}^{\prime}$ can be replaced by $\mathrm{F}^{-1}(c)$ for every negative $c$. Under the hypotheses of Corollary 2 with $\xi=\mathrm{R}$-restricting $\mathscr{N}$ if necessary-, if $c$ is close enough to $o$ for the sphere $\mathrm{F}^{-1}(c) \cap \mathrm{W}^{+}$to lie in the interior of B , a local conjugacy between X and Y can be obtained by extending to $\mathrm{W}^{-}$the unique solution of the Cauchy problem " $\mathrm{H}=\mathrm{Id}$ in $\mathrm{V}^{\prime \prime}$, where V is a neighbourhood of $\mathrm{F}^{-1}(c) \cap \mathrm{W}^{+}$in $\mathrm{F}^{-1}(c)$.

## (2.3) General idea of our proofs

Here and in the sequel, we again denote by S a diagonalisable element of $\mathfrak{g l}(n, \mathbf{C})$. If we wish to prove Theorem 2 as the Sternberg theorem, we have to answer the following two questions:

Question 1. - In the complex case, what would a good Preparation Lemma be-in particular, is there any natural analogue of the stable and unstable subspaces?

Question 2. - If Question 1 admits a positive answer, can we establish Theorem 2 by a "Cauchy problem method" as in the proof of Corollary 2?

The answer to Question I is very simple: define a strongly invariant manifold (s.i.m.) of S to be a subspace of $\mathbf{C}^{n}$ which is the unstable subspace $\mathrm{E}_{a}^{-}$of $a \mathrm{~S}$, viewed as an element of $\mathfrak{g l}(2 n, \mathbf{R})$, for some $a \in \mathbf{C}$-in other words, $\mathrm{E}_{a}^{-}$is the direct sum of those eigenspaces of S associated to eigenvalues $c$ with $\mathscr{R}_{e}(a c)>0$, which shows, in particular, that the s.i.m.'s of $\mathbf{S}$ are complex vector subspaces of $\mathbf{C}^{n}$ and there is but a finite number of them. Moreover, every S-normal form is tangent to every s.i.m. W of S at each of its points (proof: if $\mathrm{W}=\mathrm{E}_{a}^{-}$, then, considering $\mathbf{C}^{n}$ as $\mathbf{R}^{2 n}, a(\mathrm{~S}+\mathrm{N})$ is a real $a \mathrm{~S}$-normal form; therefore, it is tangent to $\mathrm{E}_{a}^{-}$at each of its points, hence our result, for $\mathrm{E}_{a}^{-}$is a complex subspace).
throughout the sequel, $\mathscr{V}$ denotes the union of the s.i.m.'s of S .
Here comes the answer to Question I:
Complex Preparation Lemma. - For every $k \in \mathbf{N}$, each S-vector field is (holomorphically) conjugate to an S -vector field X which has $k$-th order contact with an S-normal form along $\mathscr{V}$-in particular, by the above remark, X is tangent to every s.i.m. of S "at each of its points".

This result was announced in [Ch 85]; its detailed proof is contained in [Ch 86a], together with further information on strongly invariant manifolds.

Notice that S is in the Poincaré domain if and only if it admits $\mathbf{C}^{n}$ itself as a s.i.m. Thus, the Poincare-Dulac theorem is a particular case of the Complex Preparation Lemma.

It will take us the next two sections to see that the answer to Question 2 is yes. The general idea is the following: if ( $x_{1}, \ldots, x_{n}$ ) denotes a system of complex linear coordinates on $\mathbf{C}^{n}$ in which the matrix of S is diagonal, $x_{j} \circ \mathrm{~S}=c_{j} x_{j}, \quad \mathrm{I} \leqslant j \leqslant n$, the role
of the quadratic form introduced in the final remark of (2.2) will be played by the complex function $\mathbf{F}: \mathbf{C}^{n} \rightarrow \mathbf{C}$ given by

$$
\begin{equation*}
\mathbf{F}(v)=\Sigma c_{j}\left|x_{j}(v)\right|^{2} / 2 \tag{5}
\end{equation*}
$$

More precisely, we shall obtain a $\mathrm{C}^{k}$-conjugacy H between a "prepared" S-vector field and the corresponding S-normal form as the solution of the Cauchy problem " $\mathrm{H}=\mathrm{Id}$ in V " in some kind of an isolating block for the complex flow under study; here, V denotes a suitable neighbourhood of the (compact) subset $\mathbf{F}^{-1}(b) \cap \mathscr{V}$ in $\mathbf{F}^{-1}(b)$ for some regular value $b$ of $\mathbf{F}$.

The proof of Theorem I rests on the same idea, but requires a finer analysis; the reason why Theorem 2 is obtained by softer methods is that we do not strive for normal forms of the least possible degree-which can be obtained from Theorem 2, using " explicit" calculations as in the proof of Theorem I (see [Ch 86c]).

## 3. THE COMPLEX FLOW OF A WEAKLY HYPERBOLIC S

## (3. 1) Introduction and notation

We assume $S$ weakly hyperbolic and denote by $c_{1}, \ldots, c_{n}$ its eigenvalues (repeated according to their multiplicities), and by ( $x_{1}, \ldots, x_{n}$ ) a system of coordinates as in (5) above. We shall study the holomorphic flow $\sigma$ defined by

$$
\sigma(t, v)=e^{t \mathrm{~S}} v, \quad t \in \mathbf{C}, \quad v \in \mathbf{C}^{n}
$$

with the help of the function $\mathbf{F}$ defined by (5). In the sequel, we denote by $a$ a complex number such that $a \mathrm{~S} \in \mathfrak{g l}(2 n, \mathbf{R})$ is (real) hyperbolic, and by $\sigma_{1}$ the real flow $f_{a \mathrm{~S}}$ of $a \mathrm{~S}$. We let $\mathbf{F}_{1}$ be the real quadratic form $v \mapsto \mathscr{R e}(a \mathbf{F}(v))$.

The idea is to " split" our complex (indeed) flow into two comprehensible real flows: the action $\sigma_{1}$, and the action of $\mathbf{C} / a \mathbf{R}$ on the orbit space of $\sigma_{1}{ }^{(1)}$ defined by $\sigma$, which will be studied in (3.3). Most of the proofs are straightforward and will be only sketched, for they can be found-in a more general setting-in Section 5 of [Ch 86].

## (3.2) First properties

We shall start with a collection of essentially trivial results, which remain valid when S is not weakly hyperbolic, provided it has no zero eigenvalue (so that some $a \mathrm{~S}$ can be (real) hyperbolic).

[^2]Proposition 1. - The function $\mathbf{F}_{1}$ is a "Lyapunov function" for $\sigma_{1}$ : its only critical point is o, and, for each $v \in \mathbf{C}^{n} \backslash\{0\}$,
(i) the function $\psi_{v}: \mathbf{R} 2$ given by $\psi_{v}(t)=\mathbf{F}_{\mathbf{1}}\left(\sigma_{\mathbf{1}}(t, v)\right)$ is increasing, and
(ii) $\psi_{v}$ is bounded above (resp. below) if and only if $v$ belongs to the stable (resp. unstable)


Proof. - This is trivial, since $\psi_{v}(t)=\Sigma \mathscr{R} e\left(a c_{j}\right)\left|e^{a \sigma_{j} t} x_{j}(v)\right|^{2} / 2$ and (therefore) $\psi_{v}^{\prime}(t)=\Sigma\left(\mathscr{R}_{e}\left(a c_{j}\right)\right)^{2}\left|e^{a_{j} t} x_{j}(v)\right|^{2}$.

Corollary 3. - For each negative real number $c, \mathrm{Q}_{\epsilon}=\mathbf{F}_{1}^{-1}(c)$ is a quotient of the $\sigma$-invariant open subset $\mathrm{E}_{\mathrm{c}}=\mathbf{C}^{n} \backslash \mathbf{W}_{0}^{-}$by $\sigma_{1}$ in the sense of (2.2), final remark.

Here comes the "splitting" of $\sigma$ mentioned at the end of (3.1):
Corollary 4 ([Ch 86], (5.2), proposition 3). - Under the hypotheses of Corollary 3, the mapping $\varphi_{c}: \mathbf{Q}_{c} \times \mathbf{R} \rightarrow \mathbf{C}^{n}$ given by $\varphi_{c}(x, t)=\sigma_{\mathbf{l}}(t, x)$ is a diffeomorphism onto $\mathbf{E}_{c}$. For every $(x, s) \in \mathbf{Q}_{c} \times \mathbf{R}$ and every $\left(s^{\prime}, t\right) \in \mathbf{R}^{2}$,

$$
\left(\varphi_{c}^{*} e^{\left(s^{\prime}+i t\right) a s}\right)(x, s)=\left(\Phi_{c}^{t}(x), s+s^{\prime}+\int_{0}^{t} g_{c} \circ \Phi_{c}^{u}(x) d u\right),
$$

where the flow ( $\Phi_{c}^{t}$ ) on $\mathbf{Q}_{c}$ and the function $g_{c}: \mathbf{Q}_{c} \rightarrow \mathbf{R}$ are defined as follows: if $\xi_{c}$ denotes the infinitesimal generator of $\left(\Phi_{c}^{l}\right)$, then, for each $x \in \mathrm{Q}_{c}, \xi_{c}(x)=i a \mathrm{~S}(x)-g_{c}(x) a \mathrm{~S}(x)$. Thus, the flowines of $\xi_{c}$ are precisely the intersections of $\mathrm{Q}_{\varepsilon}$ with the orbits of $\left.\sigma\right|_{\mathbf{0} \times \mathrm{E}_{c}}$.

## (3.3) Structure of the flow ( $\Phi_{c}^{t}$ )

Hypotheses and notation. - We assume S weakly hyperbolic, and endow $\mathbf{C}^{n}$ with the hermitian norm $v \mapsto|v|$ defined by $|v|^{2}=\Sigma\left|x_{j}(v)\right|^{2}$. For each $\mathrm{I} \subset\{1, \ldots, n\}$, we let $\mathrm{E}_{\mathrm{I}}=\bigoplus_{j \in \mathrm{I}} \mathrm{O} x_{j}$. We still assume $c<0$, and $a$ is fixed as in (3.I).

Proposition 2 (see Fig. 2). - (i) Let $\mathscr{K}$ be the set of all equivalence classes of the equivalence relation $c_{k} \in \mathbf{R} c_{j}$ between elements $j, k$ of $\{\mathrm{I}, \ldots, n\}$. Each $\mathrm{I} \in \mathscr{K}$ determines a unique $g_{\mathrm{I}} \in \mathbf{R}$ such that

$$
a \mathbf{R} c_{j}=\mathbf{R}\left(\mathrm{I}-i g_{\mathrm{I}}\right)\left(=\left\{u \in \mathbf{C}: \mathscr{R e}\left(u\left(i-g_{\mathrm{I}}\right)\right)=\mathrm{o}\right\}\right) \quad \text { for every } j \in \mathrm{I} .
$$

(ii) For each $\mathrm{I} \in \mathscr{K}$, let $\mathrm{I}_{0}^{+}$and $\mathrm{I}_{0}^{-}$be the subsets of $\{\mathrm{I}, \ldots, n\}$ given by $\mathrm{I}_{0}^{ \pm}=\left\{j: \pm \mathscr{R} e\left(a c_{j}\left(i-g_{\mathrm{I}}\right)\right) \leqslant \mathrm{o}\right\}$. Then, $\mathrm{E}_{\mathrm{I}_{\mathrm{t}}}$ and $\mathrm{E}_{\mathrm{I}_{\mathrm{e}}}$ are s.i.m.'s of S . Conversely, every s.i.m. of S which is maximal for the inclusion is both of the form $\mathrm{E}_{\mathrm{I}_{\mathrm{t}}}, \mathrm{I} \in \mathscr{K}$ and of the form $\mathrm{E}_{\mathrm{J}_{0}}$, $\mathrm{J} \in \mathscr{K}$.


Fig. 2
We can now see how the s.i.m.'s of S appear in the structure of $\left(\Phi_{c}^{t}\right)$.
Notation. - Let $\mathscr{I}_{c}$ be the set of all $\mathrm{I} \in \mathscr{K}$ with $\mathscr{R} e\left(a c_{j}\right)<\mathrm{o}$ for some (and therefore every) $j \in \mathrm{I}$; equivalently, $\mathscr{I}_{c}$ is the set of those $\mathrm{I} \in \mathscr{K}$ such that

$$
\Sigma_{c, \mathrm{I}}=\mathrm{E}_{\mathrm{I}} \cap \mathrm{Q}_{\iota}
$$

is not empty-in which case it is a sphere. For each $I \in \mathscr{I}_{c}$, let $E_{c, 1}$ (resp. $Q_{c, I}$ ) denote the set of those $v \in \mathbf{C}^{n}$ whose orthogonal projection $v_{\mathrm{I}}$ onto $\mathrm{E}_{\mathrm{I}}$ lies in $\mathrm{E}_{c}$ (resp. $\mathbf{Q}_{c}$ ), and let $\mathrm{I}^{ \pm}=\left\{j: \pm \mathscr{R e}\left(a c_{j}\left(i-g_{\mathrm{I}}\right)\right)<\mathrm{o}\right\}=\mathrm{I}_{0}^{ \pm} \backslash \mathrm{I}$. Clearly, we can identify $\mathrm{Q}_{\mathrm{c}, \mathrm{I}}$ to $\Sigma_{c, \mathrm{I}} \times \mathrm{E}_{\mathrm{I}^{+}} \times \mathrm{E}_{\mathrm{I}^{-}}$via the canonical isomorphism of $\mathbf{C}^{n}=\mathrm{E}_{\mathrm{I}} \oplus \mathrm{E}_{\mathrm{I}^{+}} \oplus \mathrm{E}_{\mathrm{I}^{-}}$onto $\mathrm{E}_{\mathrm{I}} \times \mathrm{E}_{\mathrm{I}^{+}} \times \mathrm{E}_{\mathrm{I}^{-}}$, and the open subset $\mathrm{E}_{\mathrm{c}, \mathrm{I}}$ of $\mathbf{C}^{n}$ in $\sigma$-invariant. Therefore, it is easy to check that both $\mathrm{Q}_{c, \mathrm{I}}$ and $\mathrm{Q}_{c} \cap \mathrm{E}_{c, 1}$ are quotients of $\mathrm{E}_{c, \mathrm{I}}$ by $\sigma_{1}$ in the sense of (2.2), last remark. We let

$$
\mathscr{C}_{c, 1}: Q_{c} \cap E_{c, 1} \rightarrow Q_{c, I}
$$

denote the canonical diffeomorphism obtained by following the orbits of $\sigma_{1}$.


Fig. 3

Proposition 3 ([Ch 86], (5.2), Proposition 6 and Théorème 1). - For each $\mathrm{I} \in \mathscr{F}_{c}$, $\xi_{c}$ is " normally hyperbolic at $\Sigma_{c, \mathrm{I}}$, with global stable manifold $\mathrm{W}_{c, \mathrm{I}}^{+}=\mathrm{E}_{c, \mathrm{I}} \cap \mathrm{Q}_{c} \cap \mathrm{E}_{\mathrm{I}_{0}^{+}}$and global unstable manifold $\mathrm{W}_{c, \mathrm{I}}^{-}=\mathrm{E}_{c, \mathrm{I}} \cap \mathrm{Q}_{c} \cap \mathrm{E}_{\mathrm{I}_{0}^{-}} "$. More precisely, $\xi_{c, \mathrm{I}}=\left(\mathscr{C}_{c, \mathrm{I}}\right)_{*} \xi_{c}$ is given by $\quad \xi_{c, \mathrm{I}}=\left.\left(i-g_{\mathrm{I}}\right) a \mathrm{~S}\right|_{\mathrm{Q}_{c, 1} \mathrm{I}} ;$ thus, if we write $\mathrm{Q}_{c, \mathrm{I}}$ as $\Sigma_{c, \mathrm{I}} \times \mathrm{E}_{\mathrm{I}^{+}} \times \mathrm{E}_{\mathrm{I}^{-}}$, $\xi_{c, \mathrm{I}}$ splits into a vector field on $\Sigma_{c, \mathrm{I}}$ (namely, $\left.\xi_{c}\right|_{\Sigma_{c, 1}}$ ), a linear vector field on $\mathrm{E}_{\mathrm{I}^{+}}$whose stable subspace is $\mathrm{E}_{\mathrm{I}^{+}}$itself, and a linear vector field on $\mathrm{E}_{\mathrm{I}^{-}}$whose unstable subspace is the whole of $\mathrm{E}_{\mathrm{I}^{-}}$. In particular, the hypotheses of the Isolating Block Lemma (2.2) are satisfied by $\left(\mathbf{Q}, \mathbf{M}, \mathrm{E}^{+}, \mathrm{E}^{-}, \xi\right)=\left(\mathbf{Q}_{c, \mathrm{I}}, \Sigma_{c, \mathrm{I}}, \mathrm{E}_{\mathrm{I}^{+}}, \mathrm{E}_{\mathrm{I}^{-}}, \xi_{c, \mathrm{I}}\right)$.

Thus, in each " chart " $\mathscr{C}_{c, \text { I }}$, the situation is quite simple; we shall now see how these simple situations are glued together. Here is a first step:

Proposition 4 ([Ch 86], (5.2), Proposition 6 and Théorème 1). - We have that

$$
\bigcup_{\mathrm{I} \in \mathscr{G}_{c}} \mathrm{E}_{c, \mathrm{I}}=\mathrm{E}_{c} \quad \text { and } \quad \bigcup_{\mathrm{I} \in \mathscr{G}_{c}}\left(\mathrm{~W}_{c, \mathrm{I}}^{+} \cup \mathrm{W}_{c, \mathrm{I}}^{-}\right)=\mathrm{Q}_{c} \cap \mathscr{V} .
$$

We can now state an analogue of Proposition 1:
Proposition 5 ([Ch 86], (5.2), Théorème 2). - The function $\mathbf{f}_{c}: \mathbf{Q}_{c} \rightarrow \mathbf{R}$ given by $\mathbf{f}_{c}(x)=\mathscr{R e}(i a \mathbf{F}(x))$ is a "Lyapunov function" for the flow $\left(\Phi_{c}^{t}\right)$ : it is proper on $\mathbf{Q}_{c} \cap \mathscr{V}$, its critical set is $\Sigma_{c}=\bigcup_{\mathrm{I} \in \mathscr{G}_{c}} \Sigma_{c, \mathrm{I}}$ and, for each $x \in \mathrm{Q}_{c} \backslash \Sigma_{c}$,
(i) the function $\psi_{c, x}: \mathbf{R}$ २ defined by $\psi_{c, x}(t)=\mathbf{f}_{c}\left(\Phi_{c}^{t}(x)\right)$ is increasing and
(ii) $\psi_{c, x}$ is bounded from above (resp. below) if and only if $x$ belongs to $\mathrm{W}_{c, \mathrm{I}}^{+}\left(r e s p . \mathrm{W}_{c, \mathrm{I}}^{-}\right)$ for some $\mathbf{I} \in \mathscr{I}_{c}$.

Proposition 6. - For each $\mathrm{I} \in \mathscr{I}_{c}, \mathbf{f}_{c}\left(\Sigma_{c, \mathrm{I}}\right)=\left\{c g_{\mathrm{I}}\right\}$ (therefore, these critical values are distinct-see Fig. 2); moreover, we have $c g_{\mathrm{I}} \leqslant c g_{\mathrm{J}}$ if and only if $\mathrm{W}_{c, \mathrm{I}}^{-} \cap \mathrm{W}_{c, \mathrm{~J}}^{+}$is not empty, in which case it is contained in the $\left(\Phi_{c}^{t}\right)$-invariant sphere $\mathrm{W}_{0}^{+} \cap \mathrm{Q}_{c}$. Thus, the latter is the set of those $x \in Q_{c}$ for which the function $\psi_{c, x}$ of Proposition 5 is bounded, and

$$
\mathbf{f}_{c}\left(\mathrm{~W}_{0}^{+} \cap \mathrm{Q}_{c}\right)=\left[\min _{\mathrm{I} \in \mathscr{\xi}_{c}} c g_{\mathrm{I}}, \max _{\mathrm{I} \in \boldsymbol{g}_{c}} c g_{\mathrm{I}}\right]
$$

By Proposition 4, for each $\mathrm{I} \in \mathscr{I}_{c}$, we have that

$$
\mathrm{W}_{c, \mathrm{I}}^{-}=\left(\mathrm{Q}_{c} \cap \mathrm{E}_{\mathrm{I}_{\bullet}}\right) \backslash_{g_{\mathrm{J}}<g_{\mathrm{I}}} \mathrm{~W}_{c, \mathrm{~J}}^{-} \quad \text { and } \quad \mathrm{W}_{c, \mathrm{I}}^{+}=\left(\mathrm{Q}_{c} \cap \mathrm{E}_{\mathrm{I}_{\mathrm{o}}}\right) \backslash_{g_{\mathrm{J}}>g_{\mathrm{I}}} \mathrm{~W}_{c, \mathrm{~J}}^{+}
$$

Proof. - For each $x \in \Sigma_{c, \mathrm{I}}$, the definition of $g_{\mathrm{I}}$ implies that

$$
\begin{aligned}
\mathbf{f}_{c}(x)=\sum_{j} \mathscr{R} e\left(a\left(i-g_{\mathrm{I}}\right) c_{j}\right)\left|x_{j}(x)\right|^{2} / 2 & +\mathscr{R} e\left(g_{\mathrm{I}} a \mathbf{F}(x)\right) \\
& =\mathscr{R} e\left(g_{\mathrm{I}} a \mathbf{F}(x)\right)=g_{\mathrm{I}} \mathbf{F}_{1}(x)=c g_{\mathrm{I}}
\end{aligned}
$$

For the other assertions, see Figure 2 and Propositions 2, 4 and 5.
Proposition 7. - The critical points of $\mathbf{F}$ are those $v \in \mathbf{C}^{n}$ with $\underset{x_{j}(v) \neq 0}{ } \mathbf{R} c_{j} \neq \mathbf{C}$. Thus, the set of its critical values is $\bigcup_{1 \leqslant j \leqslant n} \mathbf{R}_{+} c_{j}$.

Corollary 5. - A real number $e$ is a regular value of $\mathbf{f}_{c}$ if and only if $b=(c-i e) / a$ is a regular value of $\mathbf{F}$, and $Q_{b}^{\prime}=\mathbf{F}^{-1}(b)$ then equals $\mathbf{f}_{c}^{-1}(e)$; if we have $e<\min \left\{c g_{\mathrm{I}}: \mathbf{I} \in \mathscr{I}_{c}\right\}$, then
(i) the union of those s.i.m.'s of S which do not intersect $\mathrm{Q}_{b}^{\prime}$ is

$$
\hat{\mathscr{V}}_{b}=\mathrm{W}_{0}^{-} \cup \bigcup_{\mathrm{I} \in \boldsymbol{S}_{c}} \mathrm{E}_{\mathrm{I}_{0}^{-}}, \quad \text { and }
$$

(ii) $\mathrm{Q}_{b}^{\prime}$ is a quotient of $\mathrm{Q}_{c} \backslash \mathrm{U}_{\mathrm{I} \in \boldsymbol{S}_{c}} \mathrm{~W}_{\mathrm{I}}^{-}$by $\left(\Phi_{c}^{t}\right)$ in the sense of (2.2), and
(iii) each $Q_{b}^{\prime} \cap \mathrm{E}_{\mathrm{I}_{+}}$with $\mathrm{I} \in \mathscr{I}_{c}$ is a compact submanifold of $\mathrm{Q}_{c}$, and

$$
\mathrm{Q}_{b}^{\prime} \cap \mathscr{V}=\bigcup_{\mathrm{I} \in \mathscr{S}_{c}}\left(\mathrm{Q}_{b}^{\prime} \cap \mathrm{W}_{\mathrm{I}}^{+}\right)=\bigcup_{\mathrm{I} \in \mathscr{F}_{c}}\left(\mathrm{Q}_{b}^{\prime} \cap \mathrm{E}_{\mathrm{I}_{0}}\right)
$$

Proof. - The first assertion is clear. Under the hypothesis of (i), it is easily checked ([Ch 86], (5.1), Corollaire 2) that a s.i.m. $\mathrm{E}_{\mathrm{I}}$ of $\mathrm{S}, \mathrm{I} \subset\{\mathrm{I}, \ldots, n\}$, is contained in $\hat{\mathscr{V}}_{b}$ if and only if $b$ does not lie in $\sum_{j \in \mathrm{I}} \mathbf{R}_{+} c_{j}$, hence (i) (see Fig. 2). Assertion (ii) is a straightforward consequence of Propositions 5 and 6. For each $I \in \mathscr{I}_{c}, Q_{Q_{b}}^{\prime} \cap \mathrm{E}_{\mathrm{I}^{+}}$ is a submanifold because $b$ is a regular value of $\left.\mathbf{F}\right|_{\mathrm{E}_{\mathrm{I}_{\mathrm{b}}}}$ (by Proposition 7 , applied to $\left.\sigma\right|_{\mathbf{C}^{2} \times \mathrm{E}_{\mathrm{I}_{\phi}}}$ instead of $\sigma$ ); it is compact because $\left.\boldsymbol{f}_{c}\right|_{Q_{c} \cap \mathscr{V}}$ is proper. The last but one equality is a consequence of (ii) and Proposition 4, and the last equality follows from it and the inclusions $\mathrm{W}_{\mathrm{I}}^{+} \subset \mathrm{E}_{\mathrm{I}^{+}} \subset \mathscr{V}, \mathrm{I} \in \mathscr{I}_{c}$.

Let us now see what Corollary 5 implies for $\sigma$ itself:
Corollary 6. - For every regular value b of $\mathbf{F}$,
(i) there exists $a \in \mathbf{C}$ such that $a \mathrm{~S}$ is (real) hyperbolic, that $c=\mathscr{R e}(a b)$ is negative and that (choosing this $a$ and this $c$ in the above theory) $e=\mathscr{R e}(i a b)$ is less than $c g_{\mathrm{I}}$ for every $\mathrm{I} \in \mathscr{I}_{c}$, and
(ii) if $\hat{\mathscr{V}}_{b}$ denotes the ( $\sigma$-invariant) union of those s.i.m.'s of S which do not intersect $\mathbf{Q}_{b}^{\prime}=\mathbf{F}^{-1}(b)$, then $Q_{b}^{\prime}$ is a quotient by $\sigma$ of the open $\sigma$-invariant subset $\mathrm{E}_{b}=\mathbf{C}^{n} \backslash \hat{\boldsymbol{V}}_{b}$, in the sense of the final remark of (2.2).

Proof. - (i) is clear, and (ii) can therefore be deduced from Corollary 5 (i), using the $\sigma$-invariance of $\hat{\mathscr{V}}_{b}$, Corollaries 3 and 4 , and the last assertion of Proposition 6 (see [Ch 86], (5.1), Corollaire 3 for a simple direct proof).

## (3.4) Concluding remarks

With the above notation, the gradient of $\mathbf{f}_{c}$ with respect to the ambiant metric of course has roughly the same properties as $\xi_{c}$. This can be used to study the topology of the cone $\mathbf{F}^{-1}(0)$-and, more generally (using paragraph (5.1)) of [Ch 86] with $r=2$ ), of the intersection of two real projective quadrics "in general position ": the advantage of this method would be to avoid the $h$-cobordism theorem (and its "bad dimensions "), so far necessary in the proof of such results [LM].

Going back to our subject, the advantage of replacing the whole of $\mathbf{C}^{n}$ by $\mathrm{E}_{b}$ (notations of Corollary 6) is clear, since the orbit space of $\left.\sigma\right|_{\mathbf{0} \times \mathrm{E}_{b}}$ is a Hausdorff manifold, diffeomorphic to $Q_{b}^{\prime}$, whereas the orbit space of $\sigma$ itself is a non-Hausdorff stratified set-the structure of which can easily be investigated via the above theory. As there is much current interest in the "quotient structures" induced by foliations or group actions, these non-trivial examples might prove useful. Anyway, it would certainly be interesting to use the above method in the hunting of topological invariants such as that defined by the Camacho-Kuiper-Ladis-Palis theorem (see (I.3), Corollary I).

## 4. NON-LINEAR COMPLEX FLOWS PROOF OF THEOREM 2

## (4.1) Hypotheses and notation

We still assume S weakly hyperbolic, and denote by $b$ an arbitrary regular value of the function $\mathbf{F}$ introduced in (3.1); in order to avoid empty statements, we assume that $\mathbf{Q}_{b}^{\prime}=\mathbf{F}^{-1}(b)$ is nonempty-or, equivalently, that $b$ lies in $\Sigma \mathbf{R}_{+} c_{j}$ (which is always the case when S is in the Siegel domain). The notations will be those of Section 3, with $a, c$ and $e$ as in Corollary 6 (i).

As $a S$ is (real) hyperbolic, the hypotheses of the Isolating Block Lemma (2.2) are satisfied by $\left(\mathrm{Q}, \mathrm{E}^{+}, \mathrm{E}^{-}, \xi\right)=\left(\mathbf{C}^{n}, \mathrm{~W}_{0}^{+}, \mathrm{W}_{0}^{-}, a \mathrm{~S}\right), \mathbf{C}^{n}$ being endowed with the hermitian metric introduced at the beginning of (3.3). Let B denote the corresponding " isolating block"; replacing $b$ by $k b$ (or the hermitian norm $|$.$| by || /$.$k ) for some$ positive $k$, small enough, we may-and shall-assume that the compact (since $\left.\mathbf{f}_{c}\right|_{Q_{c} \cap \mathcal{V}}$ is proper) subset $\mathscr{V} \cap \mathbf{f}_{c}^{-1}\left(\left[e, \max _{I \in \mathcal{S}_{c}} c g_{\mathrm{I}}\right]\right)$ lies in the interior of B ; we choose a real number $e^{\prime}>\max _{\mathrm{I} \in \varsigma_{c}} c g_{\mathrm{I}}$ such that the compact subset $\mathbf{f}_{c}^{-1}\left(\left[e, e^{\prime}\right]\right) \cap \mathscr{V}$ lies in the interior of B , and let

$$
\mathrm{L}=\mathbf{f}_{c}^{-1}\left(\left[e, e^{\prime}\right]\right) .
$$

## (4.2) "Isolating blocks" for the complex flow of S near the origin

Proposition 8 (see Fig. 4). - (i) For each compact tubular neighbourhood V of the compact subset $\mathrm{Q}_{b}^{\prime} \cap \mathscr{V}$ in $\mathrm{Q}_{b}^{\prime}$ (see Corollaries 5 and 6), the set $\tilde{\mathrm{V}}=\mathrm{L} \cap\left(\bigcup_{t \in \mathrm{R}} \Phi_{c}^{t}(\mathrm{~V}) \cup \bigcup_{\mathrm{I} \in \boldsymbol{s}_{c}} \mathrm{~W}_{c, \mathrm{I}}^{-}\right)$ is a compact neighbourhood of $\mathrm{W}_{0}^{+} \cap \mathrm{Q}_{\mathrm{c}}$ in $\mathrm{Q}_{c}$.
(ii) Such sets $\widetilde{\mathrm{V}}$ form a basis of the filter of neighbourhoods of $\mathscr{V} \cap \mathrm{L}$ in L .
(iii) For each such $\mathrm{V}, \tilde{\widetilde{V}}=\mathrm{B} \cap\left(\mathrm{W}_{0}^{-} \cup \bigcup_{t \in \mathrm{R}} \sigma_{1}(\{t\} \times \tilde{\mathrm{V}})\right)$ is a compact neighbourhood of o in $\mathbf{C}^{n}$.


Fig. 4

Proof. - (i) Given $\mathrm{I} \in \mathscr{I}_{c}$, assume that $\tilde{\mathrm{V}}$ is a neighbourhood of $\mathrm{L} \cap \mathscr{V} \backslash_{g_{j} \leqslant g_{\mathrm{I}}} \mathrm{W}_{c, \mathrm{~J}}^{-}$ -an induction hypothesis which is satisfied if $c g_{1}$ is minimal, by Corollary 5 (ii) and Propositions 3-6. Then, by the last assertion of Proposition 3 and the Isolating Block Lemma (2.2), $\widetilde{\mathrm{V}}$ is a neighbourhood of $\Sigma_{c, 1}$, hence of $\mathrm{L} \cap \mathscr{V} \backslash_{g_{\mathrm{I}}>g_{J}} \mathrm{~W}_{c, \mathrm{~J}}^{-}$by Proposition 3. By induction, this proves that $\tilde{\mathrm{V}}$ is a neighbourhood of $\mathrm{L} \cap \mathscr{V}$, which contains $W_{0}^{+} \cap Q_{c}$ by Proposition 6.

To see that $\tilde{\mathrm{V}}$ is compact, we shall use
Lemma 2. - Let $\left(a_{m}\right)=\left(u_{m}, v_{m}\right)$ be a sequence in $\mathbf{C} \times \mathbf{C}^{n}$, such that $v_{m}$ converges to a limit $v$ and that $\mathbf{F}\left(\sigma\left(a_{m}\right)\right)$ is bounded. Then, there is an increasing sequence $\left(m_{k}\right)$ in $\mathbf{N}$ such that $\sigma\left(a_{m_{k}}\right)$ converges to a limit $w$; moreover, if $\left(u_{m}\right)$ is unbounded, then $v$ lies in $\mathscr{V}$, and so does $w$ for a suitable choice of $\left(m_{k}\right)$.

Proof of Lemma 2. - If ( $u_{m}$ ) is bounded, we can extract from ( $a_{m}$ ) a convergent sequence. If ( $\dot{u}_{m}$ ) is not bounded, then, taking subsequences, we may assume that $\left|u_{m}\right| \rightarrow+\infty$, that $u_{m}| | u_{m} \mid \rightarrow u \in \mathbf{C}$ and that $\mathbf{F}\left(\sigma\left(a_{m}\right)\right) \rightarrow \mathrm{L} \in \mathbf{C}$. As S is weakly hyperbolic, there exists $u^{\prime} \in \mathbf{C}$ such that we have $\mathscr{R e}\left(u^{\prime} c_{j}\right)<0$ if and only if $\mathscr{R e}\left(u c_{j}\right)<0$, and $\mathscr{R e}\left(u^{\prime} c_{j}\right)>0$ if and only if $\mathscr{R e}\left(u c_{j}\right) \geqslant 0$. Now, $\mathscr{R} e\left(u^{\prime} \mathbf{F}\left(\sigma\left(a_{m}\right)\right)\right)=\sum_{j} \mathscr{R} e\left(u^{\prime} c_{j}\right)\left|x_{j}\left(\sigma\left(a_{m}\right)\right)\right|^{2} / 2$, and

$$
\begin{equation*}
\left|x_{j}\left(\sigma\left(a_{m}\right)\right)\right|^{2}=\left|x_{j}\left(v_{m}\right)\right|^{2} e^{2\left|u_{m}\right| x_{t}\left(\left(u-\left(u-u_{m}\left|u_{m}\right|\right) c_{j}\right)\right.} ; \tag{6}
\end{equation*}
$$

therefore, we obtain

$$
\begin{equation*}
\lim x_{j}\left(\sigma\left(a_{m}\right)\right)=0 \quad \text { for } \mathscr{R} e\left(u c_{j}\right)<0, \tag{7}
\end{equation*}
$$

hence $\mathscr{R}\left(u^{\prime} \mathrm{L}\right)=\lim \underset{\mathscr{R}\left(u^{\prime} c_{j}\right)>0}{\Sigma} \mathscr{R} e\left(u^{\prime} c_{j}\right)\left|x_{j}\left(\sigma\left(a_{m}\right)\right)\right|^{2}$. This proves that $\sigma\left(a_{m}\right)$ is bounded; therefore, by (6), $\mathscr{R e}\left(u c_{j}\right)>0$ implies $x_{j}(v)=0$, hence $v \in \mathscr{V}$; moreover, we can extract from $\sigma\left(a_{m}\right)$ a convergent subsequence, the limit of which belongs to $\mathscr{V}$ by (7), hence our lemma.

Now, let $\left(y_{m}\right)$ be a sequence in $\widetilde{\mathrm{V}}$. If it has infinitely many terms in the compact subset $\mathscr{V} \cap \mathrm{L}$, then it has a convergent subsequence with limit in $\mathscr{V} \cap \mathrm{L} \subset \tilde{\mathrm{V}}$. In the remaining case, extracting subsequences, we may assume that every $y_{m}$ is of the form $\Phi_{c}^{t_{m}}\left(v_{m}\right)$, alias $\sigma\left(u_{m}, v_{m}\right), v_{m} \in \mathrm{~V}, t_{m} \in \mathbf{R}, u_{m} \in \mathbf{C}$, and (since V is compact) that $v_{m}$ converges to some $v \in \mathrm{~V}$. If $t_{m}$ is bounded, we may assume that it converges to some $t$, and then $\lim y_{m}=\Phi_{c}^{t}(v)$ lies in L (which is closed), hence in $\tilde{\mathrm{V}}$. If $t_{m}$, hence $u_{m}$ (see Corollary 4), is unbounded, then, as $\mathbf{F}$ is bounded on $\mathbf{L}$, the hypotheses of Lemma 2 are satisfied by $\left(u_{m}, v_{m}\right)$; therefore, extracting subsequences, we may assume that $y_{m}$ tends to some $y \in \mathscr{V}$, which has to lie in L, hence in $\widetilde{\mathrm{V}}$, which is therefore compact.

Proof of (ii). - We just have to show that if a sequence $y_{m}=\Phi_{c}^{t_{m}}\left(v_{m}\right)$ in $\mathrm{L} \backslash \mathscr{V}$, $\left.v_{m} \in \mathbf{Q}_{b}^{\prime}\right) \mathscr{V}, t_{m} \geqslant 0$, is such that $v_{m}$ tends to $\mathscr{V} \cap \mathbf{Q}_{b}^{\prime}$, then some subsequence of $\left(y_{m}\right)$ converges to $\mathscr{V}$; now, extracting subsequences, we may assume that $v_{m}$ converges to some $v \in \mathscr{V} \cap \mathbf{Q}_{b}^{\prime}$, hence (ii) by Lemma 2.

Proof of (iii). - Following the orbits of $\sigma_{1}$, we get a diffeomorphism of $\mathbf{Q}_{c}$ onto the cylinder $\left\{\left|x_{+}\right|=\mathrm{I}\right\}$ (notations of the Isolating Block Lemma, with the above choice of B), hence (iii), by (i) and the Isolating Block Lemma (iii).

## (4.3) "Isolating Blocks" for perturbations of the complex flow of $S$ near o

Hypotheses and notation. - We denote by V a fixed compact tubular neighbourhood of $\mathscr{V} \cap \mathbf{Q}_{b}^{\prime}$ in $\mathbf{Q}_{b}^{\prime}$, small enough for $\widetilde{\mathrm{V}}$ to lie in the interior of B . For each $\mathrm{I} \in \mathscr{I}{ }_{c}$, we endow $Q_{c, \mathrm{I}}=\Sigma_{c, \mathrm{I}} \times \mathrm{E}_{\mathrm{I}^{+}} \times \mathrm{E}_{\mathrm{I}^{-}}$with the riemannian metric obtained by multiplying the ambiant hermitian norm by a positive constant $\alpha_{I}$ such that, if

$$
\mathrm{B}_{c, \mathrm{I}}=\left\{v \in \mathrm{Q}_{c, \mathrm{I}}: \alpha_{\mathrm{I}}\left|v_{\mathrm{I}}+\right| \leqslant \mathrm{I} \quad \text { and } \quad \alpha_{\mathrm{I}}\left|v_{\mathrm{I}}\right| \leqslant \mathrm{I}\right\},
$$

$\mathscr{C}_{c, 1}^{-1}\left(\mathrm{~B}_{c, 1}\right)$ lies in the interior of $\tilde{\mathrm{V}}$ in $\mathrm{Q}_{c}$ (of course, the hypotheses of the Isolating Block Lemma are satisfied by $\xi=\xi_{c, I}$ in $B=B_{c, I}$ for this new metric).

Complex Isolating Block Lemma, First Part. - There exists a $\mathrm{C}^{1}-n e i g h b o u r h o o d ~ \mathcal{N}^{\prime}$ of $\left.a \mathrm{~S}\right|_{\mathrm{B}}$ in the space of smooth vector fields on B holomorphic in the interior of $\mathbf{B}$ such that, denoting by $\mathscr{N}_{0}^{\prime}$ the set of those elements of $\mathscr{N}^{\prime}$ which are tangent to every s.i.m. of S , each $\mathrm{X} \in \mathcal{N}_{0}^{\prime}$ has the following properties:
(i) The Hypotheses of the Isolating Block Lemma (2.2) are satisfied if $\xi_{\mathrm{B}}=\mathrm{X}$ (considered as a real vector field), and the function $\mathrm{F}_{1}$ is increasing along every nonzero real flowline of X .
(ii) Let $\mathrm{X}_{c}$ denote the vector field on $\mathrm{Q}_{\mathrm{c}} \cap \mathrm{B}$ obtained from X as $\xi_{c}$ was obtained from aS in Corollary 4. Then, the set

$$
\tilde{\mathrm{V}}_{\mathrm{X}}=\mathrm{L} \cap\left(\bigcup_{t \geqslant 0} f_{\mathrm{x}_{c}}^{t}(\mathrm{~V}) \cup \bigcup_{\mathrm{I} \in \mathcal{F}_{c}} \mathrm{~W}_{c, \mathrm{I}}^{-}\right)
$$

is a compact $\mathrm{X}_{c}$-saturated neighbourhood of $\mathscr{V} \cap \mathrm{L}$ in L , contained in the interior of B .
(iii) The set $\widetilde{\mathrm{V}}_{\mathrm{x}}=\mathrm{B} \cap\left(\mathrm{W}_{0}^{-} \cup \bigcup_{t} f_{\mathrm{x}}^{t}\left(\widetilde{\mathrm{~V}}_{\mathrm{x}}\right)\right)$ is a compact neighbourhood of o in $\mathbf{C}^{n}$.
(iv) All the properties of $\xi_{c}$ stated in Propositions 3, 5, 6 and Corollary 5 remain true if we replace $\mathbf{Q}_{c}$ by $\tilde{\mathrm{V}}_{\mathrm{x}}, \xi_{c}$ by $\left.\mathrm{X}_{c}\right|_{\tilde{\mathrm{V}}_{\mathrm{x}}}$ and each $\mathrm{W}_{c, \mathrm{I}}^{ \pm}$by its intersection with L , with the following modifications:

- In Propositions 5-6, replace "bounded from above (resp. below)" by " bounded from above by a real number $\alpha<e^{\prime}$ (resp. bounded from below by a real number $\alpha>e$ )".
- In Proposition 3, replace the assertions after "More precisely..." by the following: " Moreover, if $\mathscr{C}_{\mathrm{x}, \mathrm{I}}:\left(\tilde{\mathrm{V}}_{\mathrm{x}}, \Sigma_{c, \mathrm{I}}\right) \rightarrow\left(\mathrm{Q}_{c, \mathrm{I}}, \Sigma_{c, \mathrm{I}}\right)$ is the local diffeomorphism which consists
in following the real orbits of X and if $\mathrm{X}_{\mathrm{I}}$ denotes the image of $\mathrm{X}_{c}$ by $\mathscr{C}_{\mathrm{X}, \mathrm{I}}$, then, the image of $\mathscr{C}_{\mathrm{X}, \mathrm{I}}$ contains $\mathrm{B}_{\mathrm{c}, \mathrm{I}}$, the hypotheses of the Isolating Block Lemma (2.2) are satisfied in $\mathrm{B}_{0, \mathrm{I}}$ by $\left(\mathrm{Q}, \mathrm{M}, \mathrm{E}^{+}, \mathrm{E}^{-}, \xi\right)=\left(\mathrm{Q}_{c, \mathrm{I}}, \Sigma_{c, \mathrm{I}}, \mathrm{E}_{\mathrm{I}^{+}}, \mathrm{E}_{\mathrm{I}^{-}}, \mathrm{X}_{\mathrm{I}}\right)$, and $\left.\mathrm{X}_{\mathrm{I}}\right|_{\mathrm{B}_{c, 1}}$ tends to $\left.\xi_{c, \mathrm{I}}\right|_{\mathrm{B}_{c, 1}}$ in the $\mathrm{C}^{1}$-topology if X tends to aS $\left.\right|_{\mathrm{B}}$ in the $\mathrm{C}^{1}$-topology."

Proof. - Let X be a smooth vector field on B, holomorphic in the interior of B and tangent to every s.i.m. of S . If X is $\mathrm{C}^{1}$-close enough to $\left.a \mathbf{S}\right|_{\mathrm{B}}$, then, by the Extension Lemma (i), assertion (i) is true; therefore, (iii) follows from (ii) as in the proof of Proposition 8 (iii). Checking (ii) and (iv) takes several steps, in which the locution " if X is $\mathrm{C}^{1}$-close enough to $\left.a \mathrm{~S}\right|_{\mathrm{B}}$ " will be implicit:

Step 1. - The second (i.e. local) part of our perturbed Proposition 3 is true. Indeed, as X is arbitrarily $\mathrm{C}^{1}$-close to $\left.a \mathrm{~S}\right|_{\mathrm{B}}$, the image of $\mathscr{C}_{\mathrm{X}, \mathrm{I}}$ does contain $\mathrm{B}_{e, \mathrm{I}}$; now, for each $v \in \mathbf{Q}_{c, 1}, \mathbf{X}_{\mathbf{I}}(v)$ is the image of $i \mathrm{X}(v)$ by the linear projection $\mathbf{C}^{n} \rightarrow \mathrm{~T}_{v} \mathbf{Q}_{c, \mathrm{I}}$ with kernel $\mathbf{R X}(v)$, which proves that $\left.\mathrm{X}_{\mathrm{I}}\right|_{\mathrm{B}_{6, \mathrm{I}}}$ is tangent to $\Sigma_{c, \mathrm{I}} \times \mathrm{E}_{\mathrm{I}^{+}}$and $\Sigma_{c, \mathrm{I}} \times \mathrm{E}_{\mathrm{I}^{-}}$and tends to $\left.\xi_{c, 1}\right|_{B, 1}$ in the $\mathrm{C}^{1}$-topology when X tends to $\left.a \mathrm{~S}\right|_{\mathrm{B}}$ in the $\mathrm{C}^{1}$-topology, hence Step I by the Extension Lemma (i).

Step 2. - The Lie derivative $\mathrm{L}_{\mathrm{x}_{c}} \mathbf{f}_{c}$ is positive outside $\Sigma_{c}$ : as $\mathrm{X}_{c}$ is arbitrarily $\mathrm{C}^{1}$-close to $\left.\xi_{c}\right|_{Q_{c} \cap B}$, this is true outside any fixed neighbourhood $U$ of $\Sigma_{c}$. To construct a neighbourhood U of $\Sigma_{c}$ in which it is also true, notice that, for each $\mathrm{I} \in \mathscr{I}_{c}, \mathrm{X}_{c}$ is tangent to $\mathrm{W}_{c, \mathrm{I}}^{+}$and to $\mathrm{W}_{c, \mathrm{I}}^{-}$, hence to $\Sigma_{c, \mathrm{I}}$, and that $\mathbf{f}_{c}$ can be written $\mathbf{f}_{c}(v)=c g_{\mathrm{I}}+\sum_{j \notin \mathrm{I}} \mathscr{R}\left(\left(i-g_{\mathrm{I}}\right) a c_{j}\right)\left|x_{j}(v)\right|^{2} / 2 ;$ therefore, by Taylor's Formula, there does exist a compact tubular neighbourhood $\mathrm{U}_{\mathrm{I}}$ of $\Sigma_{c, \mathrm{I}}$ in $\mathrm{Q}_{\rho}$, independent of X , such that $\left.\mathrm{L}_{\mathrm{X}_{c}} \mathbf{f}_{c}\right|_{\mathrm{U}_{\mathrm{I}}}$ is of the form $v \mapsto q_{\mathrm{X}}(v)\left(\left(x_{j}(v)\right)_{j \notin \mathrm{I}}\right)$, where $q_{\mathrm{X}}(v)$ denotes a quadratic form, depending continuously on $v$ and $\mathbf{X}$ (in the $\mathbf{C}^{1}$-topology), and such that every $q_{\mathrm{X}}(v)$ is positive definite if $\mathrm{X}=\left.a \mathrm{~S}\right|_{\mathrm{B}}$-hence if X is $\mathrm{C}^{1}$-close enough to $\left.a \mathrm{~S}\right|_{\mathrm{B}}$.

Step 3. - For each $\mathrm{I} \in \mathscr{I}_{c}, \quad \mathrm{~W}_{c, \mathrm{I}}^{+} \cap \mathrm{L} \quad$ (resp. $\quad \mathrm{W}_{c, \mathrm{I}}^{-} \cap \mathrm{L}$ ) is the global stable (resp. unstable) manifold of $\left.\mathrm{X}_{c}\right|_{\mathrm{B} \cap \mathrm{L}}$ at $\Sigma_{c, \mathrm{I}}$ : indeed, given $v \in \mathrm{~L} \cap \mathrm{~W}_{c, \mathrm{I}}^{+}$, every $f_{\mathrm{X}_{c}}^{t}(v)$ lies in $\mathrm{W}_{c, \mathrm{I}}^{+}$(since it belongs to $\mathrm{E}_{\mathrm{I}_{0}}$ by Proposition 2 (ii) and our hypothesis on X , this is deduced from the last equality of Proposition 6, using induction on the value of $\operatorname{cg}_{\mathrm{I}}$ ). Moreover, $f_{\mathrm{X}_{c}}^{t}(v)$ exists and lies in $\mathrm{L} \cap \mathrm{W}_{c, \mathrm{I}}^{+}$for every $t \geqslant 0$, since it can escape from the compact subset $\mathrm{L} \cap \mathscr{V} \cap \mathrm{B}$ neither through $\partial \mathrm{B}$-which $\mathrm{L} \cap \mathscr{V}$ does not intersect-, nor through $\mathbf{f}_{c}^{-1}\left(\left\{e, e^{\prime}\right\}\right)$-because of Step 2 and the inequality

$$
\begin{equation*}
e<\min \mathbf{f}_{c}\left(\mathrm{~W}_{c, \mathrm{~J}}^{-}\right)=c g_{\mathrm{J}}=\max \mathbf{f}_{c}\left(\mathrm{~W}_{c, \mathrm{~J}}^{+}\right)<e^{\prime}, \quad \mathrm{J} \in \mathscr{I}_{c}, \tag{8}
\end{equation*}
$$

obvious from Propositions 3-6. Now, for each positive $\varepsilon<\min _{J \neq K}\left|c g_{J}-c g_{\mathrm{K}}\right| / 2$, the last equalities of Proposition 6 and (8) imply that $\Sigma_{\varepsilon, J}^{+}=\mathbf{f}_{c}^{-1}\left(c g_{\mathrm{J}}-\varepsilon\right) \cap \mathrm{W}_{c, \mathrm{~J}}^{+}$(resp. $\left.\Sigma_{\varepsilon, \mathrm{J}}^{-}=\mathbf{f}_{c}^{-1}\left(c g_{\mathrm{J}}+\varepsilon\right) \cap \mathrm{W}_{c, \mathrm{~J}}^{-}\right)$is the compact submanifold $\mathbf{f}_{c}^{-1}\left(c g_{\mathrm{J}}-\varepsilon\right) \cap \mathrm{E}_{\mathrm{I}_{0}^{+}}$(resp. $\left.\mathbf{f}_{c}^{-1}\left(c g_{J}+\varepsilon\right) \cap \mathrm{E}_{\mathrm{I}_{0}}\right)$ for every $\mathrm{J} \in \mathscr{I}_{c}$. Moreover, as

$$
\mathrm{W}_{c, \mathrm{~J}}^{+} \cap \mathbf{f}_{c}^{-1}\left(c g_{\mathrm{J}}\right)=\Sigma_{\mathrm{c}, \mathrm{~J}}=\mathrm{W}_{c, \mathrm{~J}}^{-} \cap \mathbf{f}_{c}^{-1}\left(c g_{\mathrm{J}}\right),
$$

again by (8), we can choose $\varepsilon$ so small that $\Sigma_{\varepsilon, J}^{+}$and $\Sigma_{\varepsilon, J}^{-}$lie in the interior of $\mathscr{C}_{\overline{\mathrm{x}}, \mathrm{J}}^{-1}\left(\mathrm{~B}_{\mathrm{c}, \mathrm{J}}\right)$ for every J. Then, by Steps $\mathrm{I}-2$ and the Isolating Block Lemma, the interior of each $\mathscr{C}_{\mathrm{X}, \mathrm{J}}^{-1}\left(\mathrm{~B}_{c, \mathrm{~J}}\right)$ in $\mathrm{Q}_{c}$ contains a compact neighbourhood $\mathrm{C}_{\mathrm{X}, \mathrm{J}}$ of $\Sigma_{c, \mathrm{~J}}$, of the form (see Fig. 5 below)

$$
\mathrm{C}_{\mathrm{X}, \mathrm{~J}}=\mathbf{f}_{c}^{-1}\left(\left[c g_{\mathrm{J}}-\varepsilon, c g_{\mathrm{J}}+\varepsilon\right]\right) \cap\left(\mathrm{W}_{c, \mathrm{~J}}^{-} \cup \bigcup_{t \geqslant 0} f_{\mathrm{X}_{c}}^{t}\left(\mathrm{~K}_{\mathrm{J}}\right)\right)
$$

for some compact tubular neighbourhood $\mathrm{K}_{\mathrm{J}}$ of $\Sigma_{\varepsilon, J}^{+}$in $\mathbf{f}_{c}^{-\mathbf{1}}\left(c g_{J}-\varepsilon\right)$. By Step 2, $\mathrm{L}_{\mathrm{X}_{c}} \mathbf{f}_{c}$ is bounded from below by a positive constant in $\mathscr{V} \cap \mathrm{L} \backslash \mathrm{U}_{\mathrm{J}} \mathrm{C}_{\mathrm{X}, \mathrm{J}}$, where $f_{\mathrm{X}_{c}}^{t}(v)$ can therefore spend only a finite amount of non-negative time. Moreover, for every $\mathrm{J} \in \mathscr{I}_{c}$, each flowline of $\mathrm{X}_{c}$ which does not lie in $\mathrm{W}_{c, \mathrm{~J}}^{+}$and enters $\mathrm{C}_{\mathrm{X}, \mathrm{J}}$ clearly has to leave it through $\mathbf{f}_{c}^{-1}\left(c g_{J}+\varepsilon\right)$-hence forever, by Step 2-after a finite time. Therefore, $f_{\mathrm{X}_{c}}^{t}(v)$ has to be in $\mathrm{C}_{\mathrm{X}, \mathrm{I}}$ for every large enough $t$, hence our result by Step I (the case of $\mathrm{W}_{c, 1}^{-}$is of course entirely analogous).

Step 4. - End of the proof: let $\partial \mathrm{V}$ denote the boundary of our tube V in $\mathrm{Q}_{b}^{\prime}$. As $\mathrm{X}_{c}$ is arbitrarily $\mathrm{C}^{1}$-close to $\left.\xi_{c}\right|_{Q_{c} \cap B}, \mathbf{f}_{c}$ takes every value between $e$ and $e^{\prime}$ on each flowline of $\mathrm{X}_{c}$ through $\partial \mathrm{V}$, and $\mathrm{L} \cap \bigcup_{t \geqslant 0} f_{\mathrm{X}_{c}}^{t}(\partial \mathrm{~V})$ is arbitrarily $\mathrm{C}^{1}$-close to $\mathrm{L} \cap \bigcup_{t \geqslant 0} \Phi_{c}^{t}(\partial \mathrm{~V})$. Let $\widetilde{V}_{X}$ denote the " box" in $Q_{c}$, bounded laterally by $\mathrm{L} \cap \bigcup_{t \geqslant 0} f_{\mathrm{X}_{c}}^{t}(\partial \mathrm{~V})$, at one of its ends by V , and at its other end by the perturbed version of $\mathbf{f}_{c}^{-1}\left(e^{\prime}\right) \cap \tilde{\mathrm{V}}$ (i.e. the submanifold of $\mathbf{f}_{c}^{-1}\left(e^{\prime}\right)$ with boundary $\mathbf{f}_{c}^{-1}\left(e^{\prime}\right) \cap \bigcup_{t \geqslant 0} f_{\mathrm{X}_{c}}^{t}(\partial \mathrm{~V})$ which contains $\left.\mathscr{V} \cap \mathbf{f}_{c}^{-1}\left(e^{\prime}\right)\right)$. We may assume that the compact subset $\widetilde{\mathrm{V}}_{\mathrm{X}}$ contains every $\mathscr{C}_{\overline{\mathrm{X}}, \mathrm{I}}^{1}\left(\mathrm{~B}_{\mathrm{B}, \mathrm{I}}\right)$, and our problem is to show that $\widetilde{\mathrm{V}}_{\mathrm{X}}$ does admit the definition given in (ii), which is easy: given compact subsets $\mathrm{C}_{\mathrm{X}, \mathrm{I}}, \mathrm{I} \in \mathscr{I}_{c}$, as in Step 3, Steps I-3 imply that, for each $v \in \mathrm{~V}$ and each $\mathrm{I} \in \mathscr{I}_{c}$, either $v$ belongs to $\mathrm{W}_{c, \mathrm{I}}^{+}$, or there exists a real number $t_{\mathrm{I}}(v) \geqslant 0$ such that $f_{\mathrm{X}_{c}}^{t}(v)$ lies outside $\mathrm{C}_{\mathrm{X}, \mathrm{I}}$ for $t>t_{\mathrm{I}}(v)$. Now, $f_{\mathrm{X}_{c}}^{t}(v)$ can leave $\widetilde{\mathrm{V}}_{\mathrm{X}}$ for $t>0$ only through $\mathbf{f}_{c}^{-1}\left(e^{\prime}\right)$, and has to leave $\widetilde{\mathrm{V}}_{\mathrm{X}}$ if $v$ does not belong to $\mathrm{U}_{\mathrm{I}} \mathrm{W}_{c, \mathrm{I}}^{+}$, for $\mathrm{L}_{\mathrm{X}_{c}} \mathbf{f}_{c}$ is bounded from below by a positive constant in $\widetilde{\mathrm{V}}_{\mathrm{X}} \backslash \bigcup_{\mathrm{I}} \mathrm{C}_{\mathrm{X}, \mathrm{I}}$, where $f_{\mathrm{X}_{c}}^{t}(v)$ lies for $t>\max _{\mathrm{I}} t_{\mathrm{I}}(v)$. This proves what we wished.

We can now give a complex analogue of the Isolating Block Lemma (i)-(ii); the estimates in (vii) will not be needed before Section 5:

Complex Isolating Block Lemma, Second Part. - With the notation of the Complex Isolating Block Lemma, First Part, if we choose $\mathscr{N}^{\prime}$ small enough, then there exists a positive constant C such that every $\mathrm{X} \in \mathscr{N}_{0}^{\prime}$ has the following properties:
(v) $\mathrm{X} \mid \tilde{\mathrm{v}}_{\mathrm{x}}$ defines a global complex flow $\mathrm{R}_{\mathrm{x}}$ on $\widetilde{\widetilde{\mathrm{V}}}_{\mathrm{X}}$, which means the following: if $\mathrm{U}_{1}$ and $\mathrm{U}_{2}$ are connected 2-dimensional submanifolds with corners of $\mathbf{C}$, containing o, if $\gamma_{j}: \mathrm{U}_{j} \rightarrow \tilde{\widetilde{V}}_{\mathrm{x}}$, $j=1,2$, denote two smooth mappings which are holomorphic and satisfy $\gamma_{j}^{\prime}(u)=\mathrm{X}\left(\gamma_{j}(u)\right)$
inside $\mathrm{U}_{j}$ and if, moreover, $\gamma_{1}(\mathrm{o})=\gamma_{2}(\mathrm{o})=v \in \tilde{\widetilde{V}}_{\mathrm{X}}$, then, $\gamma_{1}(u)=\gamma_{2}(u)=\mathrm{R}_{\mathrm{X}}(u, v)=\mathbf{R}_{\mathrm{X}}^{u}(v)$ (which is thus defined) for every $u \in \mathrm{U}_{1} \cap \mathrm{U}_{2}$.
(vi) For each $v \in \tilde{\widetilde{V}}_{\mathrm{X}}$, the set of those $u \in \mathbf{C}$ for which $\mathrm{R}_{\mathrm{X}}^{u}(v) \in \widetilde{\widetilde{\mathrm{V}}}_{\mathrm{X}}$ is defined (in the sense of (i)) is a connected 2-dimensional submanifold with corners of $\mathbf{C}$, which is compact if and only if $v$ does not lie in $\mathscr{V}$. More precisely, let $\varphi_{\mathrm{x}}$ be the diffeomorphism of $\Omega_{\mathrm{X}}=\left\{(x, t) \in \widetilde{\mathrm{V}}_{\mathrm{X}} \times \mathbf{R}: f_{\mathrm{X}}^{t}(x) \in \widetilde{\widetilde{\mathrm{V}}}_{\mathrm{X}}\right\} \quad$ onto $\quad \tilde{\widetilde{V}}_{\mathrm{X}} \backslash \mathrm{W}_{0}^{-}$given by $\quad \varphi_{\mathrm{X}}(x, t)=f_{\mathrm{X}}^{t}(x)$, and let $g_{\mathrm{X}}: \tilde{\mathrm{V}}_{\mathrm{X}} \rightarrow \mathbf{R}$ be the smooth function defined by $i \mathrm{X}(x)=\mathrm{X}_{c}(x)+g_{\mathrm{X}}(x) \mathrm{X}(x), \quad x \in \tilde{\mathrm{~V}}_{\mathrm{X}}$. Then, for every $(x, s) \in \Omega_{\mathrm{x}}$ and every $\left(s^{\prime}, t\right) \in \mathbf{R}^{2}$,

$$
\begin{equation*}
\mathrm{R}_{\mathrm{X}}^{s^{\prime}+i t}\left(\varphi_{\mathrm{X}}(x, s)\right)=\varphi_{\mathrm{X}}\left(f_{\mathrm{X}_{c}}^{t}(x), s+s^{\prime}+\int_{0}^{t} g_{\mathrm{X}} \circ f_{\mathrm{X}_{c}}^{u}(x) d u\right) \tag{9}
\end{equation*}
$$

meaning in particular that both sides have the same definition domain.
(vii) One defines a smooth map $r_{\mathrm{X}}: \widetilde{\tilde{V}}_{\mathrm{X}} \backslash \hat{\mathscr{V}}_{b} \rightarrow \mathbf{C}$, analytic ( ${ }^{\mathbf{1}}$ ) off $\partial \mathrm{B}$, by

$$
\mathrm{R}_{\mathrm{X}}^{-r_{\mathbf{X}}(v)}(v) \in \mathrm{V}
$$

For each $v \in \widetilde{\tilde{V}}_{\mathrm{X}} \backslash \hat{\mathscr{V}}_{b}$, let $\gamma_{\mathrm{X}, v}$ be the path from o to $r_{\mathrm{X}}(v)$ in $\mathbf{C}$ such that (for some parametrisation of $\gamma_{\mathrm{X}, v}$ ) the path $s \mapsto \mathrm{R}_{\mathrm{X}}^{-\gamma_{\mathrm{x}}, v^{(s)}(v)}$ consists in following first the (real) orbit $\left\{f_{\mathrm{X}}^{t}(v)\right\}$ from $v$ to its intersection $x$ with $\widetilde{\mathrm{V}}_{\mathrm{X}}$, and then the orbit $\left\{f_{\mathbf{X}_{c}}^{t}(x)\right\}$ from $x$ to its intersection $\mathbf{R}_{\mathrm{X}}^{-{ }^{-}{ }^{(v)}(v)}$ with V ; then, denoting by d the hermitian distance in $\mathbf{C}^{n}$, we have

$$
\begin{equation*}
\left|\gamma_{\mathrm{X}, v}(s)\right| \leqslant \mathrm{C} \log \frac{2}{d\left(v, \hat{\mathscr{V}}_{b}\right)} \quad \text { for every } s \text { and every } v \in \tilde{\mathrm{~V}}_{\mathrm{X}} \backslash \hat{\boldsymbol{V}}_{b} \tag{10}
\end{equation*}
$$

Proof. - Assertions (v)-(vi) follow from the definition of $\widetilde{\mathrm{V}}_{\mathrm{x}}$. Under the hypotheses and with the notation of (vii), let $s_{\mathrm{X}}: \widetilde{\widetilde{V}}_{\mathrm{X}} \backslash \mathrm{W}_{0}^{-} \rightarrow \mathbf{R}$ and $t_{\mathrm{X}}: \widetilde{\mathrm{V}}_{\mathrm{X}} \backslash \hat{\mathscr{V}}_{b} \rightarrow \mathbf{R}$ be defined (see (i)-(vi)) by $f_{\mathbf{X}}^{-s_{\mathbf{x}}(v)}(v) \in \mathrm{Q}_{c}$ and $f_{\mathbf{X}_{c}}^{-t_{\mathbf{x}}(x)}(x) \in \mathrm{Q}_{b}^{\prime}$. As X (resp. $\mathrm{X}_{c}$ ) is transversal to $Q_{c}$ (resp. $\mathrm{Q}_{b}^{\prime}$ ) at every point of $\widetilde{\mathrm{V}}_{\mathrm{X}}$ (resp. V ), these two functions are smooth, and analytic off $\partial \mathrm{B}$. Since ( 9 ) yields

$$
\left\{\begin{array}{l}
r_{\mathrm{X}}(v)=s_{\mathrm{X}}(v)+\int_{0}^{-t_{\mathrm{X}}(x)} g_{\mathrm{X}} \circ f_{\mathrm{X}_{c}}^{u}(x) d u+i t_{\mathrm{X}}(x), \quad \text { where } x=f_{\mathrm{X}}^{-s_{\mathbf{X}}(v)}(v)  \tag{II}\\
\left|\gamma_{\mathrm{X}, v}(s)\right| \leqslant\left|s_{\mathrm{X}}(v)\right|+\left|t_{\mathrm{X}}(x)\right|\left(\mathrm{I}+\max _{y \in \widetilde{\mathrm{~V}}_{\mathbf{X}}}\left|g_{\mathrm{X}}(y)\right|\right) \quad \text { for every } s,
\end{array}\right.
$$

$r_{\mathrm{X}}$ is smooth, and analytic off $\partial \mathrm{B}$. We shall now prove (io):
Step 1. - Restricting $\mathscr{N}^{\prime}$ if necessary, there exists a positive constant $k$ such that, for $\mathrm{X} \in \mathscr{N}_{0}^{\prime}$, we have $d\left(f_{\mathrm{X}}^{-t}(v), \hat{\mathscr{V}}_{b}\right) \geqslant e^{-k|t|} d\left(v, \hat{\mathscr{V}}_{b}\right), t \in \mathbf{R}, v \in \mathrm{~B}$, and $d\left(f_{\mathrm{Xc}}^{-t}(x), \hat{\mathscr{V}}_{b}\right) \geqslant e^{-k t} d\left(x, \hat{\mathscr{V}}_{b}\right)$, $x \in \widetilde{\mathrm{~V}}_{\mathrm{X}}, \quad \mathrm{o} \leqslant t \leqslant t_{\mathrm{X}}(x)$ : let $h: \mathrm{Q}_{c} \rightarrow[\mathrm{o}, \mathrm{I}]$ be smooth, compactly supported, equal to I in a neighbourhood of $\tilde{\mathrm{V}}$, let $\mathrm{Y}=h \mathrm{X}_{c}$, and let $k_{\mathrm{X}}=\max \left\{|d \mathrm{Y}(x)|: x \in \mathrm{Q}_{c}\right\}$. If X is
${ }^{(1)}$ Analytic means real analytic.
$\mathrm{C}^{1}$-close enough to $a \mathrm{~S}$, then, for $x \in \widetilde{\mathrm{~V}}_{\mathrm{X}}$ and $\mathrm{o} \leqslant t \leqslant t_{\mathrm{X}}(x)$, we have $f_{\mathrm{X}_{c}}^{-t}(x)=f_{\mathrm{Y}}^{-t}(x)$ and, for every smooth path $\gamma:[0, \mathrm{I}] \rightarrow \mathrm{Q}_{c}$ from $f_{\mathrm{x}_{c}}^{-t}(x)$ to $\hat{\mathscr{V}}_{b} \cap \mathrm{Q}_{c}$,

$$
\begin{aligned}
& \frac{d}{d u}\left|\left(f_{Y}^{u} \circ \gamma\right)^{\prime}(s)\right|^{2}=2\left(\left(f_{Y}^{u} \circ \gamma\right)^{\prime}(s) \mid d Y\left(f_{Y}^{u} \circ \gamma(s)\right) \cdot\left(f_{Y}^{u} \circ \gamma\right)^{\prime}(s)\right) \\
& \leqslant 2 k_{\mathrm{X}}\left|\left(f_{\mathrm{Y}}^{u} \circ \gamma\right)^{\prime}(s)\right|^{2}, \quad 0 \leqslant u \leqslant t,
\end{aligned}
$$

hence, by integration, $\int_{0}^{1}\left|\left(f_{\mathrm{Y}}^{t} \circ \gamma\right)^{\prime}(s)\right| d s \leqslant e^{k_{\mathrm{X}} t} \int_{0}^{1}\left|\gamma^{\prime}(s)\right| d s$. Let $d_{c}$ denote the geodesic distance on $Q_{c}$; as $f_{\mathrm{Y}}^{t} \circ \gamma$ is a path from $x$ to $\hat{\mathscr{V}}_{b} \cap \mathrm{Q}_{c}$, we get

$$
d_{c}\left(x, \hat{\mathscr{V}}_{b} \cap Q_{c}\right) \leqslant e^{h x t} d_{c}\left(f_{\mathrm{X}_{c}}^{-t}(x), \hat{\mathscr{V}}_{b} \cap \mathrm{Q}_{c}\right) ;
$$

now, $k_{\mathrm{X}}$ depends continuously on X in the $\mathrm{C}^{1}$-topology, and, on the compact subset $h^{-1}(\mathrm{I})$, the distances $d$ and $d_{c}$ are equivalent. This proves that-restricting $\mathscr{N}^{\prime}$ if necessary-our second inequality holds for some positive constant $k$; modifying the latter, the same argument yields our first inequality.

Step 2. - Reduction of the problem: clearly, for each $\mathrm{X} \in \mathscr{N}_{0}^{\prime}, s_{\mathrm{X}}$ is bounded below, and its minimum depends continuously on X ; moreover, $s_{\mathrm{X}}$ is bounded above by the function $r_{+}$associated to $\xi=\mathrm{X}$ by the Isolating Block Lemma, which proves that $v \mapsto\left|s_{\mathrm{X}}(v) / \log \frac{2}{\left|v_{+}\right|}\right|$has a finite supremum $\mathrm{M}_{\mathrm{X}}$; by the proof of the Isolating Block Lemma, $\mathrm{M}_{\mathrm{x}}$ depends continuously on X , and we can therefore choose $\mathscr{N}^{\prime}$ so that $\mathrm{M}=\max \left\{\mathrm{M}_{\mathrm{X}}: \mathrm{X} \in \mathscr{N}_{0}^{\prime}\right\}$ is finite. As $\left|v_{+}\right|=d\left(v, \mathrm{~W}_{0}^{-}\right) \geqslant d\left(v, \hat{\mathscr{V}}_{b}\right)$, we get $\left|s_{\mathrm{X}}(v)\right| \leqslant \mathrm{M} \log \frac{2}{d\left(v, \hat{\mathscr{V}}_{b}\right),}$ hence, by Step I, $d\left(f_{\mathrm{X}}^{-s_{\mathrm{X}}(v)}(v), \hat{\mathscr{V}}_{b}\right) \geqslant 2^{-k \mathrm{M}} d\left(v, \hat{\mathscr{V}}_{b}\right)^{1+k \mathrm{M}}$ for each $v \in \tilde{\mathrm{~V}}_{\mathrm{X}} \backslash \mathrm{W}_{0}^{-}$. Therefore, by (II), we just have to prove

Step 3. - The non-negative function $x \mapsto-t_{\mathrm{x}}(x) / \log d\left(x, \hat{\mathscr{V}}_{b}\right)$ is bounded on $\widetilde{\mathrm{V}}_{\mathrm{x}} \backslash \hat{\mathcal{V}}_{b}$, and its supremum depends continuously on $\mathrm{X} \in \mathscr{N}_{0}^{\prime}$ in the $\mathrm{C}^{1}$-topology: let the compact subsets $\mathrm{C}_{\mathrm{X}, \mathrm{I}}, \mathrm{I} \in \mathscr{I}_{c}$, be as in Step 3 of the proof of the Complex Isolating Block Lemma, First Part. Let the elements $\mathrm{I}_{1}, \ldots, \mathrm{I}_{m}$ of $\mathscr{I}_{c}$ be so ordered as to satisfy $c g_{\mathrm{I}_{j}}>c g_{\mathrm{I}_{k}}$ for $j<k$; for $0 \leqslant j \leqslant m$, let $p_{j}^{+}, p_{j+1}^{-}: \widetilde{\mathrm{V}}_{\mathrm{X}} \backslash \hat{\mathscr{V}}_{b} \rightarrow \widetilde{\mathrm{~V}}_{\mathrm{X}}$ and $t_{j}^{+}, t_{j+1}^{-}: \widetilde{\mathrm{V}}_{\mathrm{X}} \backslash \widehat{\mathcal{V}}_{b} \rightarrow \mathbf{R}_{+}$ be defined inductively as follows (see Fig. 5):

$$
\left\{\begin{array}{l}
t_{0}^{+}(x)=0, \quad t_{m+1}^{-}(x)=t_{\mathrm{x}}(x) \\
\text { for } \mathrm{o}<j \leqslant m, \quad\left[t_{j}^{-}(x), t_{j}^{+}(x)\right]=\left\{\begin{array}{l}
\left\{t: f_{\mathrm{x}_{c}}^{-t}(x) \in \mathrm{C}_{\mathrm{X}, \mathrm{I}_{j}}\right\} \text { if non-empty } \\
\left\{t_{j-1}^{+}(x)\right\} \text { in the remaining case }
\end{array}\right. \\
\text { for } 0 \leqslant j \leqslant m, \quad p_{j}^{+}(x)=f_{\mathrm{X}_{c}}^{-t_{j}^{+}(x)}(x) \quad \text { and } \quad p_{i+1}^{-}=f_{\mathrm{X}_{c}}^{-t_{+1}^{-(x)}(x) .}
\end{array}\right.
$$



Fig. 5

As $L_{\mathrm{X}_{c}} \mathbf{f}_{c}$ has a positive lower bound on $\widetilde{\mathrm{V}}_{\mathrm{X}} \backslash \mathrm{U}_{\mathrm{I}} \mathrm{C}_{\mathrm{X}, \mathrm{I}}$ and $\mathbf{f}_{c} \mid \tilde{\mathrm{V}}_{\mathrm{X}}$ is bounded,
the length of $\left[0, t_{\mathrm{x}}(x)\right] \backslash \bigcup_{j}\left[t_{j}^{-}(x), t_{j}^{+}(x)\right]$ is uniformly bounded with respect to $x$ and (choosing $\mathcal{N}^{\prime}$ small enough) X.
Therefore, by Step I, $x \mapsto-t_{1}^{-}(x) / \log d\left(x, \hat{\mathscr{V}}_{b}\right)$ is bounded on $\widetilde{\mathrm{V}}_{\mathrm{X}} \backslash \hat{\mathscr{V}}_{b}$, uniformly with respect to X . Let us make the induction hypothesis that so is $x \rightarrow-t_{j}^{-}(x) / \log d\left(x, \hat{\mathscr{V}}_{b}\right)$ for some $j \leqslant m$; then by Step I, there is a positive constant $\alpha$ such that we have $d\left(p_{j}^{-}(x), \hat{\mathscr{V}}_{b}\right) \geqslant d\left(x, \hat{\mathscr{V}}_{b}\right)^{\alpha}, x \in \widetilde{\mathrm{~V}}_{\mathrm{X}} \backslash \hat{\mathscr{V}}_{b}, \mathrm{X} \in \mathscr{N}_{0}^{\prime}$. Thus, as $d\left(p_{j}^{-}(x), \mathrm{W}_{c, \mathrm{I}_{j}}^{-} \cap \mathrm{C}_{\mathrm{X}, \mathrm{I}_{j}}\right)$ is not less than $d\left(p_{j}^{-}(x), \hat{\mathscr{V}}_{b}\right)$ and, on $\mathrm{C}_{\mathrm{x}, \mathrm{I}_{j}}$, the distance induced by that of $\mathrm{Q}_{d, \mathrm{I}_{j}}$ via $\mathscr{C}_{\mathrm{X}, \mathrm{I}_{j}}$ is equivalent to $d$, it follows from the Extension Lemma (i)-applied to $\mathrm{X}_{\mathrm{I}_{j}}$ in $\mathrm{B}_{6, \mathrm{I}_{j}}$ that $-t_{j}^{+}(x) / \log d\left(x, \hat{\mathscr{V}}_{b}\right)$ is bounded in $\widetilde{\mathrm{V}}_{\mathrm{X}} \backslash \hat{\mathcal{V}}_{b}$, uniformly with respect to X ; therefore, by (12) and Step I , so is $-t_{j+1}^{-}(x) / \log d\left(x, \hat{\mathscr{V}}_{b}\right)$, hence our lemma.

## (4.4) Conjugacy results; proof of Theorem 2

Notation. - We assume $\mathscr{N}^{\prime}$ small enough for every $\mathrm{X} \in \mathscr{N}_{0}^{\prime}$ to satisfy properties (i)-(vii) of the Complex Isolating Block Lemma. For each $\mathrm{X} \in \mathscr{N}_{0}^{\prime}$, we let

$$
\dot{\mathrm{V}}_{\mathrm{x}}=\left\{x \in \tilde{\mathrm{~V}}_{\mathrm{x}}: \mathbf{f}_{c}(x)<e^{\prime}\right\}
$$

and

$$
\dot{\mathrm{V}}_{\mathrm{x}}=\left(\left\{v \in \widetilde{\tilde{\mathrm{~V}}}_{\mathrm{x}} \backslash \mathrm{~W}_{0}^{-}: \varphi_{\mathrm{x}}^{-1}(v) \in \dot{\mathrm{V}}_{\mathrm{x}} \times \mathbf{R}\right\} \cup \mathrm{W}_{0}^{-}\right) \backslash \partial \mathbf{B} .
$$

We shall construct conjugacies as in Corollary 2; here is the easy part:
Proposition 9. - (i) For each $\mathrm{X} \in \mathscr{N}_{0}^{\prime}$, the set $\omega_{\mathrm{x}}$ of those $(x, u) \in \mathrm{V} \times \mathbf{C}$ such that $\mathbf{R}_{\mathbf{X}}^{u}(x)$ is well-defined and lies in $\dot{\mathrm{V}}_{\mathrm{X}}$ is an open subset of $\mathrm{V} \times\left(\mathbf{R}+i \mathbf{R}_{+}\right)$, and the mapping $\Phi_{\mathrm{X}}: \omega_{\mathrm{X}} \rightarrow \dot{\mathrm{V}}_{\mathrm{X}} \backslash \hat{\mathcal{V}}_{b}$ given by $\Phi_{\mathrm{X}}(x, u)=\mathrm{R}_{\mathrm{X}}^{u}(x)$ is an analytic diffeomorphism, the inverse of which is $v \mapsto\left(\mathbf{R}_{\mathbf{X}}^{-{ }^{r}(v)}(v), r_{\mathrm{X}}(v)\right)$.
(ii) Given X and Y in $\mathscr{N}_{0}^{\prime}$, the mapping $h:=\Phi_{\mathrm{Y}} \circ \Phi_{\mathrm{X}}^{-1}: \Phi_{\mathrm{X}}\left(\omega_{\mathrm{X}} \cap \omega_{\mathrm{Y}}\right) \rightarrow \Phi_{\mathrm{Y}}\left(\omega_{\mathrm{X}} \cap \omega_{\mathrm{Y}}\right)$ is an analytic diffeomorphism, sending the restriction of $z \mathrm{X}$ onto that of $z \mathrm{Y}$ for every $z \in \mathbf{C}$, and, by (i),

$$
\begin{equation*}
h(v)=\mathrm{R}_{\mathrm{Y}}^{r_{\mathrm{Y}}(v)} \circ \mathrm{R}_{\mathrm{X}}^{-r_{\mathrm{X}}(v)}(v) \quad \text { and } \quad h^{-1}(v)=\mathrm{R}_{\mathrm{X}}^{r_{\mathrm{Y}}(v)} \circ \mathrm{R}_{\overline{\mathrm{Y}}}^{-r_{\mathrm{Y}}(v)}(v) . \tag{13}
\end{equation*}
$$

We can now prove a complex analogue of Corollary 2:
Complex Extension Lemma. - For every positive integer $k$, there exist an integer $p \geqslant k$ and $a \mathrm{C}^{1}$-neighbourhood $\mathscr{N}^{\prime \prime}$ of $\left.a \mathrm{~S}\right|_{\mathrm{B}}$ in $\mathscr{N}^{\prime}$ such that, if two elements $\mathrm{X}, \mathrm{Y}$ of $\mathscr{N}^{\prime \prime} \cap \mathscr{N}_{0}^{\prime}$ have $p$-th order contact along $\mathscr{V} \cap \mathrm{B}$, then, with the notation of Proposition 9,
(i) the sets $\mathrm{U}=\Phi_{\mathrm{X}}\left(\omega_{\mathrm{X}} \cap \omega_{\mathrm{Y}}\right) \cup\left(\dot{\mathrm{V}}_{\mathrm{X}} \cap \hat{\mathcal{V}}_{b}\right)$ and $\mathrm{U}^{\prime}=\Phi_{\mathrm{Y}}\left(\omega_{\mathrm{Y}} \cap \omega_{\mathrm{X}}\right) \cup\left(\dot{\mathrm{V}}_{\mathrm{Y}} \cap \hat{\mathscr{V}}_{b}\right)$ are open neighbourhoods of $\mathscr{V} \cap \dot{\mathbf{V}}_{\mathrm{X}}=\mathscr{V} \cap \dot{\mathrm{V}}_{\mathrm{Y}}$ in $\dot{\mathrm{V}}_{\mathrm{X}}$ and $\dot{\mathrm{V}}_{\mathrm{Y}}$ respectively, and
(ii) the mapping $\mathrm{H}: \mathrm{U} \rightarrow \mathrm{U}^{\prime}$, equal to $h$ off $\hat{\mathscr{V}}_{b}$ and to Id on $\hat{\mathscr{V}}_{b}$, is $a \mathrm{C}^{k}$-diffeomorphism, such that $\mathrm{H}_{*}\left(\left.z \mathrm{X}\right|_{\mathrm{U}}\right)=\left.z \mathrm{Y}\right|_{\mathrm{U}^{\prime}}$ for every $z \in \mathbf{C}$.

Proof. - Let $\delta>\mathrm{o}$ be small enough for $\sigma([\mathrm{o}, \delta] \times \mathrm{V})$ to lie in the interior of B , let $\mathbf{T}=\mathbf{R} / \delta \mathbf{Z}$, and let $q: \mathrm{Q}_{\iota} \times \mathbf{R} \rightarrow \mathrm{Q}_{\iota} \times \mathbf{T}$ denote the canonical projection. For every $\mathrm{X} \in \mathscr{N}_{0}^{\prime}$, close enough to $\left.a \mathrm{~S}\right|_{\mathrm{B}}$, we have $\widetilde{\mathrm{V}}_{\mathrm{X}} \times[0, \delta] \subset \varphi_{\mathrm{X}}{ }^{1}(\mathbf{B} \backslash \overrightarrow{\mathrm{~B}})$; thus, as $[\mathrm{X}, i \mathrm{X}]=\mathrm{o}$, there is a unique analytic vector field $\mathbf{X}_{c}$ on $\widetilde{\mathrm{V}}_{\mathrm{X}} \times \mathbf{T}$ such that $q^{*} \mathbf{X}_{c}=\varphi_{\mathbf{X}}^{*}\left(i \mathbf{X} \mid \tilde{\tilde{v}}_{\mathbf{x}} \backslash W_{0}\right)$, the flow of which is given by

$$
\begin{equation*}
f_{\mathbf{x}_{c}}^{t} \circ q(x, s)=q\left(f_{\mathrm{X}_{c}}^{t}(x), s+\int_{0}^{t} g_{\mathrm{x}} \circ f_{\mathrm{X}_{c}}^{u}(x) d u\right) \tag{I4}
\end{equation*}
$$

meaning in particular that both sides of this identity are defined in the same domain (recall that the definition domain of $g_{\mathrm{X}}$ is $\widetilde{\mathrm{V}}_{\mathrm{x}}$ ).

In the rest of the proof, we assume $\mathrm{X}, \mathrm{Y} \in \mathscr{N}_{0}^{\prime}$ close enough to $\left.a \mathrm{~S}\right|_{\mathrm{B}}$. The general idea is as follows: using the "charts" $\varphi_{\mathrm{X}}$ and $\varphi_{\mathrm{Y}}$, we transform our initial complex conjugacy problem into another one, concerning real flows (Steps i and 2). We can then use the Extension Lemma (2.2) to solve this reduced problem (Step 3); this is why we have introduced $\mathbf{T}$ and $q$, as a more direct approach would make use of a less simple tool.

Step 1. - Let $\mathrm{U}_{b}=\mathrm{U} \backslash \hat{\boldsymbol{V}}_{b}, \mathrm{U}_{b}^{\prime}=\mathrm{U}^{\prime} \backslash \hat{\boldsymbol{V}}_{b}$, let $d_{0}$ be the set of those $x \in \dot{\mathrm{~V}}_{\mathrm{x}}$ such that, with the notation of (II), $f_{\mathbf{Y}_{c}(x)}^{t_{c}} \circ f_{\mathrm{X}_{c}}^{-t_{\mathbf{x}}(x)}(x)$ is defined and lies in $\dot{\mathrm{V}}_{\mathrm{Y}}$, and let $h_{0}: d_{0} \times \mathbf{T} \rightarrow \dot{\mathrm{V}}_{\mathbf{Y}} \times \mathbf{T}$ and $q^{*} h_{0}: d_{0}^{c} \times \mathbf{R} \rightarrow \dot{\mathrm{V}}_{\mathbf{Y}} \times \mathbf{R}$ be given by

$$
h_{0}(x, \theta)=f_{\mathbf{X}_{c}}^{t_{x}(x)} \circ f_{\mathbf{x}_{c}}^{-t_{\mathbf{x}}(x)}(x, \theta) \quad \text { and } \quad q^{*} h_{0}(x, s)=f_{q^{*}+\mathbf{x}_{c}}^{t_{x}(x)} \circ f_{q^{-} \times \mathbf{x}_{c}}^{-t_{\mathbf{x}}(x)}(x, s) .
$$

Then, $\mathrm{U}_{b}=\varphi_{\mathrm{X}}\left(\Omega_{\mathrm{X}} \cap\left(q^{*} h_{0}\right)^{-1}\left(\Omega_{\mathrm{Y}}\right)\right)$ and

$$
\begin{equation*}
h=\left.\varphi_{\mathrm{Y}} \circ\left(q^{*} h_{0}\right) \circ \varphi_{\mathrm{X}}^{-1}\right|_{\mathrm{U}_{b}} . \tag{15}
\end{equation*}
$$

This is obvious from (9), (II) and (I3), since

$$
f_{q^{*} \mathbf{X}_{c}}^{t}(x, s)=\left(f_{\mathbf{X}_{c}}^{t}(x), s+\int_{0}^{t} g_{\mathrm{X}} \circ f_{\mathbf{X}_{c}}^{u}(x) d u\right)
$$

Step 2. - Let $m \geqslant k$ and $\mathscr{N}_{0}$ be as in the Extension Lemma with $\xi=a S$; assume that
(i) X and Y belong to $\mathscr{N}_{0}$ and have $m$-th order contact along $\mathscr{V}$,
(ii) $\mathrm{D}_{\mathbf{0}}=d_{\mathbf{0}} \cup\left(\hat{\mathscr{V}}_{b} \cap \dot{\mathrm{~V}}_{\mathrm{X}}\right)$ is an open neighbourhood of $\mathscr{V} \cap \dot{\mathrm{V}}_{\mathrm{x}}$ in $\dot{\mathrm{V}}_{\mathrm{X}}$, and
(iii) $h_{0}$ extends to a $\mathrm{C}^{m}$-embedding $\mathrm{H}_{0}: \mathrm{D}_{0} \times \mathbf{T} \rightarrow \dot{\mathrm{V}}_{\mathrm{Y}} \times \mathbf{T}$, having m-th order contact with the identity along ( $\left.\mathscr{V} \cap \dot{\mathrm{V}}_{\mathbf{x}}\right) \times \mathbf{T}$.

Then, the conclusion of the Complex Extension Lemma is satisfied.

Indeed, $q^{*} h_{0}$ clearly extends to a $\mathrm{C}^{m}$-embedding $q^{*} \mathrm{H}_{0}: \mathrm{D}_{0} \times \mathbf{R} \rightarrow \dot{\mathrm{V}}_{\mathbf{Y}} \times \mathbf{R}$, having $m$-th order contact with Id along $\left(\mathscr{V} \cap \dot{V}_{\mathbf{X}}\right) \times \mathbf{R}$; as X and Y have $m$-th order contact along every s.i.m. of $S$ and are tangent to it, it follows that

$$
\begin{equation*}
\mathrm{U}_{c}:=\mathrm{U} \backslash \mathrm{~W}_{0}^{-}=\varphi_{\mathrm{X}}\left(\Omega_{\mathrm{X}} \cap\left(q^{*} \mathrm{H}_{0}\right)^{-1}\left(\Omega_{\mathrm{Y}}\right)\right) \tag{16}
\end{equation*}
$$

is an open neighbourhood of $\dot{\mathrm{V}}_{\mathrm{X}} \cap \mathscr{V} \backslash \mathrm{W}_{0}^{-}$in $\dot{\mathrm{V}}_{\mathrm{X}}$, and that $\left.\mathrm{H}\right|_{\mathrm{U}_{c}}=\left.\varphi_{\mathrm{Y}} \circ\left(q^{*} \mathrm{H}_{0}\right) \circ \varphi_{\mathrm{X}}^{-1}\right|_{\mathrm{U}_{c}}$ is a $\mathrm{C}^{m}$-diffeomorphism onto $\mathrm{U}^{\prime} \backslash \mathrm{W}_{0}^{-}$, having $m$-th order contact with Id along $\dot{\mathrm{V}}_{\mathrm{x}} \cap \mathscr{V} \backslash \mathrm{W}_{0}^{-}$, hence in particular along $\mathrm{W}_{0}^{+} \backslash\{0\}$. As $\mathrm{U}_{c}$ is X-saturated by (16), we can apply the Extension Lemma (2.2) with $(\omega, h)=\left(\stackrel{\circ}{\mathrm{U}}_{c},\left.\mathrm{H}\right|_{\mathrm{U}_{c}} ^{\circ}\right)$; since the set $\Omega$ provided by the Extension Lemma clearly satisfies $\Omega \backslash \mathrm{W}_{0}^{-}=\stackrel{\circ}{\mathrm{U}}_{c}$, this prove Step 2.

For each $\mathrm{I} \in \mathscr{I}_{c}$, let $q_{\mathrm{I}}: \mathbf{Q}_{c, \mathrm{I}} \times \mathbf{R} \rightarrow \mathbf{Q}_{c, \mathrm{I}} \times \mathbf{T}=\mathbf{Q}_{c, \mathrm{I}}$ be the canonical projection, and let $\varphi_{c, \mathrm{I}}: \mathrm{Q}_{c, \mathrm{I}} \times \mathbf{R} \rightarrow \mathrm{E}_{c, \mathrm{I}}$ be defined by $\varphi_{c, \mathrm{I}}(x, s)=\sigma_{\mathbf{1}}(s, x)$. As $[a \mathrm{~S}, i a \mathrm{~S}]=\mathrm{o}$, there exists a unique analytic vector field $\xi_{c, \mathrm{I}}$ on $\mathrm{Q}_{c, \mathrm{I}} \times \mathbf{T}$ such that $q_{\mathrm{I}}^{*} \xi_{c, \mathrm{I}}=\varphi_{c, \mathrm{I}}^{*}(i a \mathrm{~S})$, given by $\xi_{c, \mathrm{I}}(x, \theta)=\left(\xi_{c, \mathrm{I}}(x), g_{\mathrm{I}}\right)$; therefore, setting $\mathbf{B}_{c, \mathrm{I}}=\mathrm{B}_{c, \mathrm{I}} \times \mathbf{T}$ and $\boldsymbol{\Sigma}_{c, \mathrm{I}}=\Sigma_{c, \mathrm{I}} \times \mathbf{T}$, the hypotheses of the Isolating Block Lemma (2.2) are satisfied by $(\mathbf{Q}, \mathbf{B}, \boldsymbol{\xi})=\left(\mathbf{Q}_{c, \mathrm{I}}, \mathbf{B}_{c, \mathrm{I}}, \boldsymbol{\xi}_{c, \mathrm{I}}\right)$. Here is our last step:

Step 3. - Let the elements $\mathrm{I}_{1}, \ldots, \mathrm{I}_{\ell}$ of $\mathscr{I}_{c}$ be so numbered as to satisfy $c g_{\mathrm{I}_{j}}<c g_{\mathrm{I}_{j}}$ for $j<j^{\prime}$, and let the integers $m=m_{\ell+1} \leqslant m_{\ell} \leqslant \ldots \leqslant m_{1}$ be defined as follows: $m$ is the same as in Step 2 and, for $1 \leqslant j \leqslant \ell, \quad m_{j}$ is the integer $m$ associated to ( $\left.\mathrm{B}, \boldsymbol{\xi}^{2}, \boldsymbol{\Sigma}, k\right)=\left(\mathbf{B}_{c, \mathrm{I}_{j}}, \boldsymbol{\xi}_{c, \mathrm{I}_{j}}, \mathbf{\Sigma}_{c, \mathrm{I}_{j}}, m_{j+1}\right)$ by the Extension Lemma (2.2). Then, the hypotheses of Step 2 are satisfied if X and Y have $m_{1}-$ th order contact along $\mathscr{V}$.

The proof is by induction. As above, set $\mathbf{Q}_{c}=Q_{c} \times \mathbf{T}, \mathbf{d}_{0}=d_{0} \times \mathbf{T}$, etc. Let $\dot{\mathrm{L}}=\mathrm{L} \backslash \mathbf{f}_{\mathrm{c}}^{-1}\left(e^{\prime}\right)$, let $\mathrm{H}_{\mathbf{0}}: \mathbf{D}_{\mathbf{0}} \rightarrow \dot{\mathrm{V}}_{\mathrm{Y}}$ equal $h_{0}$ off $\hat{\mathscr{V}}_{b}$ and Id on $\hat{\mathscr{V}}_{b}$, and let $\mathscr{W}_{j}=\dot{\mathrm{L}} \cap \mathscr{V} \backslash_{j^{\prime} \geqslant j} \mathrm{~W}_{c, \mathrm{I}_{j^{\prime}}}^{-}, \mathrm{I} \leqslant j \leqslant \ell+\mathrm{I}$. We shall establish that, for $\mathrm{I} \leqslant j \leqslant \ell+\mathrm{I}$, $d_{0} \cup \mathscr{W}_{j}$ is an open neighbourhood of $\mathscr{W}_{j}$ in $\dot{\mathrm{L}}$, and $\left.\mathrm{H}_{0}\right|_{d_{0} \cup w_{j}}$ is of class $\mathrm{C}^{m}$ and has $m_{j}$-th order contact with Id along $\mathscr{W}_{j}$.
This will imply our result for $j=\ell+1$, since $m_{\ell+1}=m$ and $\mathscr{W}_{\ell+1}=\mathscr{V} \cap \dot{\mathrm{L}}$.
Notice that (14) and the Complex Isolating Block Lemma (iv) yield (17) ${ }_{1}$; indeed, $\mathbf{V}$ is a quotient of $\tilde{\mathbf{V}}_{\mathbf{X}} \backslash \hat{\boldsymbol{V}}_{b}$ (resp. $\tilde{\mathbf{V}}_{\mathbf{Y}} \backslash \hat{\boldsymbol{V}}_{b}$ ) by $f_{\mathbf{X}_{c}}$ (resp. $f_{\mathbf{Y}_{c}}$ ), and $h_{0}$ is that conjugacy between $\mathbf{X}_{c}$ and $\mathbf{Y}_{c}$ which is the " maximal" solution of the Cauchy problem " $h_{0}=$ Id on $\mathbf{V}$ " with domain in $\dot{\mathbf{V}}_{\mathrm{X}}$ and range in $\dot{\mathbf{V}}_{\mathrm{Y}}$; therefore, since each $\mathbf{W}_{\mathrm{I}}^{+} \cap \mathbf{L}$ is the stable manifold at $\boldsymbol{\Sigma}_{c, \mathrm{I}}$ of both $\mathbf{X}_{c}$ and $\mathbf{Y}_{c}$, which have $m_{1}$-th order contact along it, (17) is true.

Let us make the induction hypothesis that so is $(17)_{j}$ for some $j \leqslant \ell$. Using the Extension Lemma (2.2), we shall prove that $d_{0} \cup \mathscr{W}_{j+1}$ is a neighbourhood of $\mathscr{W}_{j} \cup \Sigma_{c, I_{j}}$ in $\dot{\mathrm{L}}$ and that, near $\boldsymbol{\Sigma}_{c, \mathrm{I}_{j}}, H_{0}$ has $m_{j+1}$-th order contact with Id along $\mathbf{W}_{c, \mathrm{I}_{j}}^{-}$. Since $\mathbf{W}_{c, \mathrm{I}_{j}}^{-} \cap \mathbf{L}$ is the global unstable manifold at $\boldsymbol{\Sigma}_{c, \mathrm{I}_{j}}$ of both $\mathbf{X}_{c}$ and $\mathbf{Y}_{c}$, which have $m_{j+1}$-th order contact along it, this will imply ( $\left.I_{7}\right)_{j+1}$, as $\mathrm{H}_{0}$ is a conjugacy between $\mathbf{X}_{c}$ and $\mathbf{Y}_{c}$ on and off $\mathscr{V}$.

Let $\varphi_{\mathrm{X}, \mathrm{j}}:\left(\mathbf{Q}_{c, \mathrm{I}_{j}} \times \mathbf{R}, \boldsymbol{\Sigma}_{c, \mathrm{I}_{j}} \times \mathbf{R}\right) \rightarrow \mathrm{E}_{c, \mathrm{I}_{j}}$ be the local diffeomorphism $(x, s) \mapsto f_{\mathbf{X}}^{s}(x)$, let $\mathscr{C}_{\mathrm{X}, \mathrm{j}}:\left(\mathbf{Q}_{c}, \boldsymbol{\Sigma}_{c, \mathrm{I}_{j}}\right) \rightarrow\left(\mathbf{Q}_{c, \mathrm{I}_{j}}, \boldsymbol{\Sigma}_{c, \mathrm{I}_{j}}\right)$ be the local "chart" such that $q_{\mathrm{I}_{j}} \circ \varphi_{\mathrm{X}, j}^{-1} \circ \varphi_{\mathrm{X}}=\mathscr{C}_{\mathrm{X}, j} \circ q$, and let $\mathscr{C}_{\mathrm{Y}, j}$ be associated to Y in the same fashion. By the perturbed version of Proposition 3 in the Complex Isolating Block Lemma (iv), $\mathbf{B}_{c, \mathrm{I}_{j}}$ lies in the interior of the images of $\mathscr{C}_{\mathrm{X}, \mathrm{j}}$ and $\mathscr{C}_{\mathrm{Y}, j}$; moreover, $\mathbf{X}_{c, j}=\left.\left(\mathscr{C}_{\mathrm{X}, \mathrm{j}}\right)_{*} \mathbf{X}_{c}\right|_{\mathbf{B}, \mathrm{I}_{j}}$ and $\mathbf{Y}_{c, j}=\left.\left(\mathscr{C}_{\mathbf{Y}, j}\right)_{*} \mathbf{Y}_{c}\right|_{\mathbf{B} c, \mathrm{I}_{j}}$ are $\mathbf{C}^{1}$-small perturbations of $\left.\boldsymbol{\xi}_{c, \mathrm{I}_{j}}\right|_{\mathbf{B}_{c, 1},}$, having $m_{j}$-th order contact along the image of $\mathbf{W}_{c, \mathrm{I}_{j}}^{+} \cup \mathbf{W}_{c, \mathrm{I}_{j}}^{-}$by $\mathscr{C}_{\mathrm{X}, j}$ and $\mathscr{C}_{\mathrm{Y}, j}$, i.e. along $\boldsymbol{\Sigma}_{c, \mathrm{I}_{j}} \times\left(\mathrm{E}_{\mathrm{r}_{j}^{+}} \cup \mathrm{E}_{\mathrm{I}_{j}^{-}}\right)$. Therefore, we can apply part (ii) of the Extension Lemma (2.2) with

$$
(\mathbf{B}, \boldsymbol{\xi}, \mathbf{X}, \mathrm{Y}, k, m, \omega, h)=\left(\mathbf{B}_{c, \mathrm{I}_{j}}, \boldsymbol{\xi}_{c, \mathrm{I}_{j}}, \mathbf{X}_{c, \mathrm{j}}, \mathbf{Y}_{c, j}, m_{j+1}, m_{j}, \boldsymbol{\omega}_{j}, \mathbf{h}_{\mathrm{j}}\right)
$$

where $\mathbf{h}_{j}=\left.\mathscr{C}_{\mathbf{Y}, j} \circ\left(\left.\mathrm{H}_{\mathbf{0}}\right|_{d_{0} \cup \boldsymbol{w}_{j}}\right) \circ \mathscr{C}_{\mathbf{X}, j}^{-1}\right|_{\omega_{j}}$ and $\boldsymbol{\omega}_{\mathrm{j}}$ is defined as follows: given $\mathrm{C}_{\mathrm{X}, \mathrm{I}_{j}}$ and $\mathrm{C}_{\mathrm{Y}, \mathrm{I}_{j}}$ as in Step 3 of the proof of the Complex Isolating Block Lemma, First Part, $\boldsymbol{\omega}_{j}$ is the ( $\mathbf{X}_{c, j}$-saturated) set of those $\mathbf{x}$ in the interior of $\mathbf{C}_{\mathrm{X}, \mathrm{I}_{j}}$ such that $\mathbf{h}_{j}(\mathbf{x})$ lies in the interior of $\mathbf{C}_{\mathbf{Y}, \mathrm{I}_{j}}$. As the set $\Omega$ provided by the Extension Lemma in this situation clearly satisfies $\mathscr{C}_{\mathrm{X}, j}^{-1}(\Omega) \subset \mathrm{D}_{0}$, we conclude that $\mathbf{D}_{0}$ and $\mathrm{H}_{0}$ have the required properties near $\boldsymbol{\Sigma}_{c, \mathrm{I}_{j}}$.

Corollary 7. - Theorem 2 is true.
Proof. - Let $k$ be a positive integer, and let $p \geqslant k$ be as in the Complex Extension Lemma. By the Complex Preparation Lemma (2.3), we just have to prove the fol-
lowing: if an S-vector field has $p$-th order contact with an S-normal form $\mathrm{S}+\mathrm{N}$ along $\mathscr{V}$, then Z is $\mathrm{C}^{k}$-conjugate to $\mathrm{S}+\mathrm{N}$. This will be a consequence of the Complex Extension Lemma if we can prove that there exists $A \in G L(n, \mathbf{C})$, with $A^{*} S=S$, such that-given a representative $\widetilde{\mathbf{Z}}$ of $\mathrm{Z}-\left.a \mathrm{~A}^{*} \widetilde{\mathrm{Z}}\right|_{\mathrm{B}}$ and $\left.a \mathrm{~A}^{*}(\mathrm{~S}+\mathrm{N})\right|_{\mathrm{B}}$ are well-defined and lie in $\mathscr{N}^{\prime \prime} \cap \mathscr{N}_{0}^{\prime}$.

Now, this is quite easy: for each $\lambda$ in the spectrum $\Lambda$ of S , let $d(\lambda)$ be the dimension of the corresponding eigenspace $\mathrm{E}(\lambda)$ of S , and let $\left(x_{\lambda}^{j}\right)_{1 \leqslant j \leqslant d(\lambda)}$ be a system of complex linear coordinates on $E(\lambda)$ in which the nilpotent endomorphism $\left.\mathrm{N}^{1}\right|_{E(\lambda)}$ has an upper triangular matrix. If $d=\max d(\lambda)$, define $\mathrm{A}_{\varepsilon}, \varepsilon>0$, by

$$
\mathrm{A}_{\varepsilon} \mathrm{E}(\lambda)=\mathrm{E}(\lambda) \quad \text { and }\left.\quad x_{\lambda}^{j} \circ \mathrm{~A}\right|_{\mathrm{E}(\lambda)}=\varepsilon^{j+d} x_{\lambda}^{j}, \quad \mathrm{I} \leqslant j \leqslant d(\lambda), \quad \lambda \in \Lambda .
$$

For every small enough $\varepsilon, A=A_{\varepsilon}$ fulfils our requirements. Indeed, $A_{\varepsilon}$ preserves $S$ and (therefore) each one of its s.i.m.s', and both $A_{\varepsilon}^{*} \widetilde{Z}$ and $A_{\varepsilon}^{*}(S+N)$ tend to $S$ on $B$ in the $\mathrm{C}^{1}$-topology when $\varepsilon$ tends to $o$.

## 5. PROOF OF THEOREM

## (5.1) Algebraic background: normal forms and their formal flows

Hypotheses and notations. - Given a (not necessarily weakly hyperbolic) S, we let $\sigma$ and $c_{j}, x_{j}, \mathrm{I} \leqslant j \leqslant n$, be as in (3.1), and $\sigma^{u}(v)=\sigma(u, v)$. For each $p \in \mathbf{N}^{n}$, we denote by $p_{1}, \ldots, p_{n}$ its coordinates, and let $|p|=\Sigma p_{j}$ and $x^{p}=x_{1}^{p_{1}} \ldots x_{n}^{p_{n}}$. For each $X \in \mathbf{d}$, we denote by $\hat{\mathbf{R}}_{\mathrm{X}}$ the Taylor expansion of its complex flow at ( $o, o$ ), viewed as a convergent power series in the variables $u, x_{1}, \ldots, x_{n}$, where $u$ is the (complex) parameter of the flow; we let dom $\widehat{\mathbf{R}}_{\mathrm{x}}$ be the strict convergence domain of $\widehat{\mathbf{R}}_{\mathrm{x}}$-if $\hat{\mathbf{R}}_{\mathrm{X}}=\Sigma a_{m, p} u^{m} x^{p}$, recall that $\operatorname{dom} \widehat{\mathbf{R}}_{\mathrm{X}}$ is the (open) set of those $\left(u^{\prime}, v^{\prime}\right) \in \mathbf{C} \times \mathbf{C}^{n}$ such that, for some $(u, v) \in \mathbf{C} \times \mathbf{C}^{n}$ with $|u|>\left|u^{\prime}\right|$ and $\left|x_{j}(v)\right|>\left|x_{j}\left(v^{\prime}\right)\right|$ for every $j$, the set $\left\{a_{m, p} u^{m} x^{p}(v)\right\}$ is bounded; for each $(u, v) \in \operatorname{dom} \widehat{\mathbf{R}}_{\mathrm{x}}$, we shall write $\hat{\mathbf{R}}_{\mathbf{X}}(u, v)=\hat{\mathbf{R}}_{\mathrm{X}}\left(u, x_{\mathbf{1}}(v), \ldots, x_{n}(v)\right)=\Sigma a_{m, p} u^{m} x^{p}(v)$-which is well-defined, by Abel's lemma.

Proposition 10. - For every S -normal form $\mathrm{S}+\mathrm{N}$,
(i) the polynomial vector field $\mathbf{N}-\mathbf{N}^{1}$ is a linear combination of the monomials $x^{p} \frac{\partial}{\partial x_{j}}$,
$p \in \mathrm{P}_{j}=\left\{p \in \mathbf{N}^{n}:|p|>\mathrm{I}\right.$ and $\left.\Sigma p_{k} c_{k}=c_{j}\right\}, \quad \mathrm{I} \leqslant j \leqslant n$,
(ii) if we view $\hat{\mathrm{R}}_{\mathrm{N}}$ and $\widehat{\mathrm{R}}_{\mathrm{N}^{1}}$ as elements of $\mathbf{C}\{u\}\left[\left[x_{1}, \ldots, x_{n}\right]\right]^{n}$, then

$$
x_{j} \circ\left(\hat{\mathbf{R}}_{\mathrm{N}}-\hat{\mathbf{R}}_{\mathrm{N}^{2}}\right)=\sum_{p \in \mathrm{P}_{j}} a_{\mathrm{N}, p}(u) x^{p} \quad \text { for } \mathrm{I} \leqslant j \leqslant n,
$$

where each $a_{\mathrm{N}, p}: \mathbf{C} 2$ is polynomial; moreover, seeting $\hat{\mathbf{R}}_{\mathrm{N}}^{u}(v)=\hat{\mathbf{R}}_{\mathrm{N}}(u, v)$, we have

$$
\hat{\mathbf{R}}_{\mathrm{S}+\mathrm{N}}^{u}=\sigma^{u} \circ \hat{\mathbf{R}}_{\mathrm{N}}^{u}=\hat{\mathbf{R}}_{\mathrm{N}}^{u} \circ \sigma^{u} \quad \text { for each } u \text {, and }
$$

(iii) if $\mathbf{N}^{\mathbf{1}}=\mathbf{o}$, then the degree of each $a_{\mathrm{N}, p}$ is less than $|p|$.

Proof. - Since $\left[\mathrm{S}, \mathrm{N}-\mathrm{N}^{1}\right]=\mathrm{o}$ and $\mathrm{S}=\Sigma c_{j} x_{j} \frac{\partial}{\partial x_{j}}$, (i) is clear. To prove the rest, we shall use a linear representation: denote by $\mathscr{E}$ the complex algebra of all germs at $\quad o \in \mathbf{C}^{n}$ of holomorphic complex functions, by $m$ its maximal ideal $\{f \in \mathscr{E}: f(\mathrm{o})=0\}$ and by $\mathscr{E}_{k}$ the finite dimensional algebra $\mathscr{E} / \mathrm{m}^{k+1}$ for each nonnegative integer $k$. Each $\mathrm{X} \in \mathbf{d}$ defines the derivation $\mathrm{X}^{*}: f \mapsto \mathrm{~L}_{\mathrm{X}} f$ of $\mathscr{E}$, hence a derivation $\mathrm{X}_{k}^{*}$ of $\mathscr{E}_{k}$. Denoting by $\mathrm{J}^{k}$ the canonical projection of $\mathscr{E}$ onto $\mathscr{E}_{k}$, we clearly have that $\mathrm{J}^{k}\left(x_{j} \circ \hat{\mathbf{R}}_{\mathrm{N}}^{u}\right)=e^{u \mathrm{~N}_{k}^{*}}\left(\mathrm{~J}^{k} x_{\mathrm{j}}\right)$ for every $(j, k, u)\left({ }^{\mathbf{1}}\right)$. Now, as $\mathrm{N}^{1}$ is nilpotent, so is $\mathbf{N}_{k}^{*}$ for each $k$; therefore, each $u \mapsto \mathrm{~J}^{k} x_{j} \circ \widehat{\mathbf{R}}_{\mathrm{N}}^{u}$ is polynomial, and its degree is less than $k$ if $\mathbf{N}^{1}=\mathbf{o}$ (because this means that $\mathbf{N}^{*} \mathfrak{m}^{q} \subset \mathfrak{m}^{q+1}$ for every positive integer $q$ ). From this and the fact that each $\hat{\mathrm{R}}_{\mathrm{N}}^{u}$ preserves S , assertions (ii)-(iii) follow at once.

The following obvious result ([Ch 86], (4.3.2), Lemme 4) will be useful:
Proposition 11. - Let $\mathbf{N}^{n}$ be equipped with the ordering $\leqslant$ defined by " $p \leqslant q$ if and only if $p_{j} \leqslant q_{j}$ for every $j ’$; then, for each $\mathrm{A} \subset \mathbf{N}^{n}$, the minimal set min A of A is finite, and every $p \in \mathrm{~A}$ satisfies $p \geqslant q$ for some $q \in \min \mathrm{~A}$. Therefore, if $\mathbf{P}_{0}=\left\{p \in \mathbf{N}^{n} \backslash\{0\}: \Sigma p_{k} c_{k}=0\right\}$ and if $\mathrm{P}_{1}, \ldots, \mathrm{P}_{n}$ are as in Proposition 10, then, for $\mathrm{o} \leqslant j \leqslant n$, each element of $\mathrm{P}_{j}$ can be written-in a non-unique fashion in general-as the sum of one element of $\min \mathrm{P}_{j}$ and finitely many elements of $\min \mathbf{P}_{\mathbf{0}}$.

These two results yield the easy part of the Poincare-Dulac theorem:
Corollary 8. - If $\mathrm{P}_{0}$ is empty-which of course is the case when S is in the Poincaré domain-then
(i) each $\mathrm{P}_{j}$ is finite, and, for every $\mathrm{X} \in \mathbf{d}$ with $[\mathrm{S}, \mathrm{X}]=0, \mathrm{X}-\mathrm{X}^{\mathbf{1}}$ is a linear combination of the monomials $x^{p} \frac{\partial}{\partial x_{j}}, \quad p \in \mathrm{P}_{j}, \quad \mathrm{I} \leqslant j \leqslant n$, and
(ii) every S -normal form $\mathrm{S}+\mathrm{N}$ generates a global holomorphic action of $\mathbf{C}$ on $\mathbf{C}^{n}$; more precisely, by Proposition 10 (ii), the complex flow of $\mathbf{N}$ is an algebraic $\mathbf{C}$-action on $\mathbf{C}^{n}$.

We shall now see what happens if $\mathrm{P}_{\mathbf{0}}$ is nonempty. For simplicity, we consider only special S-normal forms, in the following sense: call an S-vector field special if $\mathbf{X}^{\mathbf{1}}=\mathrm{S}$ (of course, such an $X$ has nothing special in the " generic " case when $S$ has only simple eigenvalues!).

Corollary 9. - There exists a positive integer $k$ such that, if $z: \mathbf{C}^{n} \rightarrow \mathbf{C}^{\min \mathrm{P}_{0}}$ and $w: \mathbf{C} \times \mathbf{C}^{n} \rightarrow \mathbf{C}^{\{1, \ldots, k\} \times \min \mathbf{P}_{0}}$ are given by

$$
\left\{\begin{array}{l}
z(v)=\left(z_{p}(v)\right)_{p \in \min \mathrm{P}_{0}} \quad \text { and } \quad z_{p}=x^{p}, \quad p \in \min \mathrm{P}_{0} \\
w(u, v)=\left(w_{m, p}(u, v)\right)_{1 \leqslant m \leqslant k, p \in \min \mathrm{P}_{0}}, \quad \text { where } w_{m, p}=u^{m} z_{p}
\end{array}\right.
$$

[^3]the following hold: for each special S -normal form $\mathrm{S}+\mathrm{N}$, there are polynomial functions $\mathrm{M}_{\mathrm{N}, p}: \mathbf{C}$ 2 and convergent power series $\mathrm{R}_{\mathrm{N}, p}$ at $\mathrm{o} \in \mathbf{C}^{\min \mathrm{P}_{0}} \times \mathbf{C}^{\{1, \ldots, k\} \times \min \mathrm{P}_{0}}, p \in \mathrm{U}_{j>0} \min \mathrm{P}_{j}$, such that
\[

$$
\begin{equation*}
x_{j}\left(\hat{\mathbb{R}}_{\mathrm{N}}(u, v)-v\right)=\sum_{p \in \min \mathrm{P}_{j}}\left(u \mathrm{M}_{\mathrm{N}, p}(u)+\mathrm{R}_{\mathrm{N}, p}(z(v), w(u, v))\right) x^{p}(v), \tag{18}
\end{equation*}
$$

\]

$$
\mathrm{I} \leqslant j \leqslant n
$$

and

$$
\begin{equation*}
\mathrm{R}_{\mathrm{N}, p}=\mathrm{o} \quad \text { on } \mathbf{C}^{\min \mathrm{P}_{0}} \times\{\mathrm{o}\} \tag{19}
\end{equation*}
$$

Identity (18) means in particular that we have $(u, v) \in \operatorname{dom} \hat{\mathbf{R}}_{\mathbf{N}}$ if and only if $(z(v), w(u, v))$ lies in the strict convergence domain $\operatorname{dom} \mathrm{R}_{\mathrm{N}, p}$ for every $p$.

Proof. - By Proposition ıo (ii)-(iii), $x_{j} \circ \hat{\mathbf{R}}_{\mathrm{N}}=x_{j}+\sum_{p \in \mathrm{P}_{j}, 0<m<|p|} a_{m, p} u^{m} x^{p}, \quad \mathrm{I} \leqslant j \leqslant n$. Now, by Proposition 11 , there does exist a positive integer $k$ such that each $u^{m} x^{p}$, $p \in \bigcup_{j>0} \mathrm{P}_{\mathrm{j}} \backslash \min \mathrm{P}_{j}$ and $0<m<|p|$, can be written $z^{\mathrm{K}(m, p)} w^{\mathrm{L}(m, p)} x^{\mathbf{Q}(p)}$, with $\mathrm{Q}(p) \in \bigcup_{j>0} \min \mathrm{P}_{j}$ and $|\mathrm{L}(m, p)|>0$. We shall stick to the following

Rule. - Choose exactly one such mapping $(m, p) \xrightarrow{\mathbf{J}}(\mathbf{K}(m, p), \mathrm{L}(m, p), \mathrm{Q}(p))$.
Clearly, J is injective. Therefore, if we set $b_{r, \ell, q}=a_{m, p}$ if $(r, \ell, q)=\mathrm{J}(m, p)$ and $b_{r, \ell, q}=0$ in the remaining case, our corollary follows at the formal level, with $u \mathrm{M}_{\mathrm{N}, p}(u)=\sum_{0<m<|p|} a_{m, p} u^{m}$ and $\mathrm{R}_{\mathrm{N}, p}(z, w)=\sum_{r, \ell} b_{r, \ell, p} z^{r} w^{\ell}$.

Now, given $(u, v)$ and ( $u^{\prime}, v^{\prime}$ ) in $\mathbf{C} \times \mathbf{C}^{n}$,
a) the set $\left\{a_{m, p} u^{m} x^{p}(v)\right\}$ is bounded if and only if so is $\left\{b_{r, \ell, q} z^{r}(v) w^{\ell}(u, v) x^{q}(v)\right\}$, and
b) the inequalities $\left|u^{\prime}\right|<|u|$ and $\left|x_{j}\left(v^{\prime}\right)\right|<\left|x_{j}(v)\right|$ for every $j$ imply $\left|z_{p}\left(v^{\prime}\right)\right|<\left|z_{p}(v)\right|$ and $\left|w_{m, p}\left(u^{\prime}, v^{\prime}\right)\right|<\left|w_{m, p}(u, v)\right|$ for every ( $m, p$; this and $a$ ) show that $\left(u^{\prime}, v^{\prime}\right) \operatorname{dom} \hat{\mathbf{R}}_{\mathrm{N}}$ implies $\left(z\left(v^{\prime}\right), w\left(u^{\prime}, v^{\prime}\right)\right) \in \operatorname{dom} \mathbf{R}_{\mathrm{N}, p}$ for every $p$;
c) conversely, if we have $\left|z_{p}\left(v^{\prime}\right)\right|<\left|z_{p}(v)\right|$ and $\left|w_{m, p}\left(u^{\prime}, v^{\prime}\right)\right|<\left|w_{m, p}(u, v)\right|$ for every ( $m, p$ ), then, assuming that $\left\{b_{r, \ell, q} z^{r}(v) w^{\ell}(u, v)\right\}$ is bounded for every $q$, we get the following: as $(u, v) \mapsto(z(v), w(u, v))$ is continuous, there exists $\quad\left(u^{\prime \prime}, v^{\prime \prime}\right) \in \mathbf{C} \times \mathbf{C}^{n}$ with $\left|u^{\prime \prime}\right|>\left|u^{\prime}\right|$ and $\left|x_{j}\left(v^{\prime \prime}\right)\right|>\left|x_{j}\left(v^{\prime}\right)\right|$ for every $j$, satisfying $\left|z_{p}\left(v^{\prime \prime}\right)\right|<\left|z_{p}(v)\right|$ and $\left|w_{m, p}\left(u^{\prime \prime}, v^{\prime \prime}\right)\right|<\left|w_{m, p}(u, v)\right|$ for every $(m, p)$. In other words, by $\left.a\right)$, $\left(z\left(v^{\prime}\right), w\left(u^{\prime}, v^{\prime}\right)\right) \in \operatorname{dom} \mathrm{R}_{\mathrm{N}, p}$ for every $p$ implies $\left(u^{\prime}, v^{\prime}\right) \in \operatorname{dom} \hat{\mathbf{R}}_{\mathrm{N}}$.

Corollary 10. - If $\mathrm{S}+\mathrm{N}$ and $\mathrm{S}+\mathrm{N}^{\prime}$ are two special S -normal forms, then, there are polynomial functions $\left.\mathrm{Q}_{\mathrm{N}, \mathrm{N}^{\prime}, p}: \mathbf{C}\right)$ and convergent power series $\mathrm{T}_{\mathrm{N}, \mathrm{N}^{\prime}, p}$ at $\mathrm{o} \in \mathbf{C}^{\min \mathrm{P}_{0}} \times \mathbf{C}^{\{1, \ldots, k\} \times \min \mathrm{P}_{0}}, \quad p \in \bigcup_{j>0} \min \mathbf{P}_{j}$, such that, for $\mathrm{I} \leqslant j \leqslant n$,

$$
\begin{align*}
& x_{j}\left(\hat{\mathbf{R}}_{\mathrm{S}+\mathrm{N}^{\prime}}^{u} \circ \hat{\mathbf{R}}_{\mathrm{S}+\mathrm{N}}^{-u}(v)\right)=x_{j}\left(\hat{\mathbf{R}}_{\mathrm{N}^{\prime}}^{u} \circ \hat{\mathbf{R}}_{\mathrm{N}}^{-u}(v)\right)  \tag{20}\\
& =x_{j}(v)+\sum_{p \in \min \mathrm{P}_{j}}\left(u \mathrm{Q}_{\mathrm{N}, \mathrm{~N}^{\prime}, p}(u)+\mathrm{T}_{\mathrm{N}, \mathrm{~N}^{\prime}, p}(z(v), w(u, v))\right) x^{p}(v)
\end{align*}
$$

and
(2I)

$$
\mathrm{T}_{\mathrm{N}, \mathrm{~N}^{\prime}, p}=\mathrm{o} \quad \text { on } \mathbf{C}^{\min \mathrm{P}_{0}} \times\{\mathrm{o}\}
$$

Identity (20) means in particular that, if we have $(-u, v) \in \operatorname{dom} \widehat{\mathbf{R}}_{\mathrm{N}}$ and $\left(u, \hat{\mathbf{R}}_{\mathrm{N}}{ }^{u}(v)\right) \in \operatorname{dom} \hat{\mathbf{R}}_{\mathrm{N}^{\prime}}$, then $(z(v), w(u, v))$ belongs to dom $\mathrm{T}_{\mathrm{N}, \mathrm{N}^{\prime}, p}$ for each $p$.

Proof. - Since $N$ and $N^{\prime}$ lie in the centraliser of $S, \hat{\mathbf{R}}_{\mathbf{S}+\mathbf{N}^{\prime}}^{u} \circ \hat{\mathbf{R}}_{\mathbf{S}+\mathrm{N}}^{-u}=\hat{\mathbf{R}}_{\mathbf{N}^{\prime}}^{u} \circ \widehat{\mathbf{R}}_{\mathbf{N}}{ }^{u}$.
Now, Proposition 10 (ii)-(iii) and straightforward calculations yield

$$
x_{j} \circ \hat{\mathbf{R}}_{\mathbf{N}^{\prime}}^{u} \circ \hat{\mathbf{R}}_{\mathrm{N}}^{-u}=x_{j}+\sum_{p \in \mathrm{P}_{j}} b_{\mathrm{N}, \mathrm{~N}^{\prime}, p}(u) x^{p}
$$

where each $\left.b_{\mathrm{N}, \mathrm{N}^{\prime}, p}: \mathbf{C}\right)$ is polynomial, of degree less than $|\boldsymbol{p}|$, and vanishes at 0 . We conclude as in the proof of Corollary 9.

## (5.2) A generalisation of Theorem $I$

If $S$ is weakly hyperbolic, call an S-normal form $S+N$ reduced if $N$ is a linear combination of monomials $x^{p} \frac{\partial}{\partial x_{j}}, \quad p \in \mathrm{P}_{j}, \quad \mathrm{I} \leqslant j \leqslant n$, such that every $c_{k}$ with $p_{k} \neq 0$ lies in $\mathbf{R} c_{j}$. In particular, if $S$ is hyperbolic, the only reduced S-normal form is $S$ itself-and, of course, every S-normal form is special. Therefore, Theorem $I$ is a particular case of

Theorem 3. - If S is weakly hyperbolic, then every special S -vector field is $\mathrm{C}^{0}$-conjugate to a reduced special S-normal form.

Before proving this result, let us explain why it provides " good" normal forms:
Proposition 12. - If S is weakly hyperbolic, then every reduced S-normal form $\mathrm{S}+\mathrm{N}$ generates a holomorphic $\mathbf{C}$-action $\mathbf{R}_{\mathrm{S}+\mathrm{N}}$ of the form $\mathbf{R}_{\mathrm{S}+\mathrm{N}}^{u}=\mathbf{R}_{\mathrm{N}}^{u} \circ \sigma^{u}=\sigma^{u} \circ \mathbf{R}_{\mathrm{N}}^{u}$, where $\mathbf{R}_{\mathrm{N}}$ is the algebraic $\mathbf{C}$-action $(u, v) \mapsto \mathbf{R}_{\mathrm{N}}^{u}(v), \quad u \in \mathbf{C}, \quad v \in \mathbf{C}^{n}$, generated by $\mathbf{N}$. More precisely, denoting by $v_{\mathrm{J}}$ the canonical projection of $v \in \mathbf{C}^{n}=\bigoplus_{\mathrm{I} \in \mathscr{K}} \mathrm{E}_{\mathrm{I}}$ onto $\mathrm{E}_{\mathrm{J}}, \mathrm{J} \in \mathscr{K}$, the hypothesis that $\mathrm{S}+\mathrm{N}$ is reduced means that $\mathrm{N}(v)=\sum_{\mathrm{I} \in \mathscr{K}} \mathrm{N}\left(v_{\mathrm{I}}\right)$ for each $v \in \mathbf{C}^{n}$; for every $\mathrm{I}, \mathrm{N}_{\mathrm{I}}=\left.\mathrm{N}\right|_{\mathrm{E}_{\mathrm{I}}}$ and $\mathrm{S}_{\mathrm{I}}=\left.\mathrm{S}\right|_{\mathrm{E}_{\mathrm{I}}}$ are vector fields on $\mathrm{E}_{\mathrm{I}}$, and $\mathrm{S}_{\mathrm{I}}$ is in the Poincaré domain. Therefore, by Corollary 8 , the flow $\mathrm{R}_{\mathrm{N}_{\mathrm{I}}}$ of each $\mathrm{N}_{\mathrm{I}}$ is an algebraic action, and the actions $\mathrm{R}_{\mathrm{N}}$ and $\mathrm{R}_{\mathrm{S}+\mathrm{N}}$ " split" as follows: for each $v \in \mathbf{C}^{n}$,
hence

$$
\begin{aligned}
& \mathbf{R}_{\mathrm{N}}(u, v)=\sum_{\mathrm{I} \in \mathscr{K}} \mathrm{R}_{\mathrm{N}_{\mathrm{I}}}\left(u, v_{\mathrm{I}}\right), \\
& \mathbf{R}_{\mathrm{S}+\mathrm{N}}^{u}(v)=\sum_{\mathrm{I} \in \mathscr{K}} \mathrm{R}_{\mathrm{S}_{\mathrm{I}}}^{u} \circ \mathrm{R}_{\mathrm{N}_{\mathrm{I}}}^{u}\left(v_{\mathrm{I}}\right)=\sum_{\mathrm{I} \in \mathscr{K}} \mathrm{R}_{\mathrm{N}_{\mathrm{I}}}^{u} \circ \mathrm{R}_{\mathrm{S}_{\mathrm{I}}}^{u}\left(v_{\mathrm{I}}\right)
\end{aligned}
$$

Proof. - For each $\mathrm{I} \in \mathscr{K}, \mathrm{S}$ and N are tangent to each of the two s.i.m. $\mathrm{E}_{\mathrm{r}_{0}}$ and $\mathrm{E}_{\mathrm{I}_{0}^{-}}$of S , hence to their intersection $\mathrm{E}_{\mathrm{I}}$. As every $\mathrm{S}_{\mathrm{I}}$ is in the Poincaré domain by weak
hyperbolicity, one just has to check our characterisation of reduced S-normal forms, which is easy.

Remark. - When S in only weakly hyperbolic, a reduced special S-normal form $\mathrm{S}+\mathrm{N}$ may be $\mathrm{C}^{0}$-unequivalent to S : for example, if $n=2, c_{1}=\mathrm{I}, c_{2}=2$ and $\mathrm{N}=x_{1}^{2} \frac{\partial}{\partial x_{2}}$, then every leaf of the foliation defined by S (see the beginning of ( I .3 )) is a punctured complex line with puncture at $o \in \mathbf{C}^{2}$, whereas the leaves of the foliation defined by $\mathrm{S}+\mathrm{N}$ are injectively immersed complex lines, with the sole exception of $\mathrm{O} y \backslash\{0\}$.

## (5.3) Proof of Theorem 3

Still assuming S weakly hyperbolic, we shall establish
Lemma 3. - For each regular value bof $\mathbf{F}$, every special S -normal form $\mathrm{S}+\mathrm{N}_{0}$ is $\mathrm{C}^{0}$-conjugate to the (special) S-normal form $\mathrm{S}+\mathbf{N}_{\mathbf{1}}$ obtained from $\mathrm{S}+\mathrm{N}_{\mathbf{0}}$ by cancelling the coefficient of every monomial $x^{p} \frac{\partial}{\partial x_{j}}, \quad p \in \mathbf{P}_{j}$, such that $b \in \sum_{p_{k} \neq 0} \mathbf{R}_{+} c_{k}$.

This implies Theorem 3: by Theorem 2, every special S-vector field is C1-conjugate to a special S-normal form $S+N$ (to see that it is indeed special, just notice that the conjugacies constructed in the proofs of the Complex Preparation Lemma [Ch 86a] and of Corollary 7 are tangent to the identity at $\mathbf{o}$ ). Now, $\mathbf{N}$ is a linear combination of monomials $x^{p} \frac{\partial}{\partial x_{j}}, p \in \mathrm{P}_{j}, \mathrm{I} \leqslant j \leqslant n$, and, for each such monomial, Proposition 7 yields the following: either $\sum_{p_{k} \neq 0} \mathbf{R}_{+} c_{k}$ is a half-line, or it contains a regular value of $\mathbf{F}$. Therefore, applying Lemma 3 finitely many times, we obtain that $\mathrm{S}+\mathrm{N}$ is $\mathrm{C}^{0}$-conjugate to $\mathrm{S}+\mathrm{N}^{\prime}$, where $\mathrm{N}^{\prime}$ is obtained from N by cancelling the coefficients of those $x^{p} \frac{\partial}{\partial x_{j}}$ such that $\sum_{p_{k} \neq 0} \mathbf{R}_{+} c_{k}$ is not a half-line, hence Theorem 3 by weak hyperbolicity.

Proof of Lemma 3. - We may assume $b \in \Sigma \mathbf{R} c_{j}$, as $\mathrm{N}_{0}=\mathrm{N}_{1}$ if this is not the case. Our other hypotheses and notations will be those of (4.1).

Step 1. - For each $p \in \mathbf{N}^{n}$, we have $b \in \sum_{p_{k} \neq 0} \mathbf{R}_{+} c_{k}$ if and only if $x^{p}$ vanishes identically on $\hat{\mathscr{V}}_{b}$. Indeed, a s.i.m. $\mathrm{E}_{\mathrm{I}}$ of $\mathrm{S}, \mathrm{I} \subset\{\mathrm{I}, \ldots, n\}$, is contained in $\hat{\mathscr{V}}_{b}$ if and only if $b \notin \sum_{j \in \mathrm{I}} \mathrm{R}_{+} c_{j}$. Therefore, we have $b \in \sum_{p_{k} \neq 0} \mathrm{R}_{+} c_{k}$ if and only if, for every such I, there exists $k \notin \mathrm{I}$ with $p_{k} \neq \mathrm{o}$, hence our result.

Hypotheses and notation. - For each $\varepsilon>0$, let $\widetilde{\varepsilon}=\varepsilon \operatorname{Id} \in \operatorname{GL}(n, \mathbf{C})$. In the sequel, we let $\mathrm{N}=\widetilde{\varepsilon}^{*} \mathrm{~N}_{0}, \quad \mathrm{~N}^{\prime}=\widetilde{\varepsilon}^{*} \mathrm{~N}_{\mathbf{1}}, \quad \mathrm{X}=\left.a \widetilde{\varepsilon}^{*}\left(\mathrm{~S}+\mathrm{N}_{0}\right)\right|_{\mathrm{B}}=\left.a(\mathrm{~S}+\mathrm{N})\right|_{\mathrm{B}}$, $\mathrm{Y}=\left.a \widetilde{\varepsilon}^{*}\left(\mathrm{~S}+\mathrm{N}_{1}\right)\right|_{\mathrm{B}}=\left.a\left(\mathrm{~S}+\mathrm{N}^{\prime}\right)\right|_{\mathrm{B}}$, and the locution "for every small enough $\varepsilon$ " is implicit. When $\varepsilon$ tends to $o$, so do N and $\mathrm{N}^{\prime}$ in the $\mathrm{C}^{1}$-topology. Therefore, we may (and shall) assume that X and Y belong to the set $\mathscr{N}_{0}^{\prime}$ defined in the Complex Isolating Block Lemma. To simplify notations, we assume $a=\mathrm{I}$, which of course is no restriction.

We shall prove that the mapping $h$ associated to $\mathrm{X}, \mathrm{Y}$ by Proposition 9 (ii) fulfils conditions (i)-(ii) of the Complex Extension Lemma with $k=\mathbf{o}$.

This extension problem will be solved using Corollaries 9-10, which will allow us to obtain our conjugacy as the sum of a convergent power series in variables which are continuous functions. More precisely, we shall see that the mapping obtained by replacing $\mathrm{R}_{\mathrm{X}}$ and $\mathrm{R}_{\mathrm{Y}}$ by $\hat{\mathrm{R}}_{\mathrm{X}}$ : dom $\hat{\mathrm{R}}_{\mathrm{X}} \rightarrow \mathbf{C}^{n}$ and $\hat{\mathrm{R}}_{\mathrm{Y}}: \operatorname{dom} \hat{\mathrm{R}}_{\mathrm{Y}} \rightarrow \mathbf{C}^{n}$ in Proposition 9 (I3) can be extended continuously by Id on $\hat{\mathscr{V}}_{b} \cap \dot{\mathrm{~V}}_{\mathrm{X}}$ and coincides with $h$ near $\mathscr{V} \cap \dot{\mathrm{V}}_{\mathrm{x}}$.

Step 2. - For each $(m, p) \in \mathbf{N} \times \mathbf{N}^{n}$, if $x^{p}$ vanishes identically on $\hat{\mathscr{V}}_{b}$, then
(i) the mapping $\tilde{\widetilde{V}}_{\mathrm{X}} \ni v \mapsto \begin{cases}\gamma_{\mathrm{X}, v}(s)^{m} x^{p}(v) & \text { if } v \notin \hat{\mathscr{V}}_{b} \\ 0 & \text { if } v \in \hat{\mathscr{V}}_{b}\end{cases}$
is uniformly bounded with respect to $s$ and (small enough) $\varepsilon$, and
(ii) the mapping $v \mapsto \begin{cases}r_{\mathrm{X}}(v)^{m} x^{p}(v) & \text { if } v \notin \hat{\mathscr{V}}_{b} \\ 0 & \text { if } v \in \hat{\mathscr{V}}_{b}\end{cases}$
is continuous on $\widetilde{\widetilde{V}}_{\mathrm{x}}$.
Indeed, for every $v \in \tilde{\widetilde{V}}_{\mathrm{X}} \backslash \hat{\boldsymbol{V}}_{b}$, our hypothesis on $p$ yields $\left|x^{p}(v)\right| \leqslant d\left(v, \hat{\boldsymbol{\gamma}}_{b}\right)$, hence (i) and-since $r_{\mathrm{x}}$ is continuous-(ii), by the Complex Isolating Block Lemma (vii).

Step 3.- (i) The mapping $\widetilde{\widetilde{V}}_{\mathrm{X}} \ni v \mapsto \begin{cases}\left(z(v), w\left(r_{\mathrm{X}}(v), v\right)\right) & \text { if } v \notin \hat{\mathscr{V}}_{b} \\ (\mathrm{o}, \mathrm{o}) & \text { if } v \in \hat{\mathscr{V}}_{b}\end{cases}$ is continuous.
(ii) We have $\left(-r_{\mathrm{X}}(v), v\right) \in \operatorname{dom} \hat{\mathrm{R}}_{\mathrm{N}}$ and $\hat{\mathrm{R}}_{\mathrm{s}+\mathrm{N}}^{-r_{\mathrm{x}}(v)}(v)=\mathrm{R}_{\mathrm{X}}^{-r_{\mathrm{X}}(v)}(v)$ for each $v \in \widetilde{\widetilde{\mathrm{~V}}}_{\mathrm{X}} \backslash \hat{\boldsymbol{V}}_{b}$.

Indeed, by Step I, each $x^{p}$ with $p \in \mathrm{P}_{0}$ vanishes identically on $\hat{\mathscr{V}}_{b}$ (as $\Sigma p_{j} c_{j}=0$, the origin lies in the convex hull of $\left\{c_{j}: p_{j} \neq 0\right\}$, hence in its interior by weak hyperbolicity), and thus Step 2 (ii) yields (i). Moreover, by Step 2 (i), the mapping $v \mapsto\left(z(v), w\left(-\gamma_{\mathrm{x}, v}(s), v\right)\right)$ is uniformly bounded on $\tilde{\widetilde{V}}_{\mathrm{X}} \backslash \hat{\boldsymbol{V}}_{b}$ with respect to $s$ and (small enough) $\varepsilon$. Now, we have $\operatorname{dom} \mathrm{R}_{\mathrm{N}, p} \supset \varepsilon^{-1} \operatorname{dom} \mathrm{R}_{\mathrm{N}_{0}, p}$ for each $p \in \bigcup_{j>0} \min \mathrm{P}_{j}$;
therefore, if $\varepsilon$ is small enough, Corollary 9 implies $\left(-\gamma_{X, v}(s), v\right) \in \operatorname{dom} \hat{\mathrm{R}}_{\mathrm{N}}$ for every $v \in \widetilde{\mathrm{~V}}_{\mathrm{X}} \backslash \hat{\boldsymbol{V}}_{b}$ and every $s$. This proves (ii), for both $\mathrm{R}_{\mathrm{X}}^{-r_{\mathrm{X}}(v)}(v)$ and $\hat{\mathrm{R}}_{\mathrm{X}}^{-r_{\mathrm{X}}(v)}(v)$ are obtained by " integrating $S+N$ along the path $-\gamma_{X, v}$ ".

Step 4. - (i) For each $v \in \widetilde{\widetilde{V}}_{\mathrm{X}} \backslash \hat{\boldsymbol{V}}_{b}$, we have $\left(r_{\mathrm{X}}(v), \mathrm{R}_{\mathrm{X}}^{-r_{\mathrm{x}}(v)}(v)\right) \in \operatorname{dom} \hat{\mathrm{R}}_{\mathrm{N}}$.
(ii) The mapping $\hat{\mathrm{H}}: \dot{\mathrm{V}}_{\mathrm{X}} \rightarrow \mathbf{C}^{n}$, equal to Id on $\hat{\mathscr{V}}_{b}$ and to $v \mapsto \hat{\mathbf{R}}_{\mathbf{S}^{r(v)}+\mathrm{N}^{\prime}} \circ \mathrm{R}_{\mathrm{X}}^{-r_{\mathrm{x}}(v)}(v)$ off $\hat{\mathscr{r}}_{b}$, is continuous.

Indeed, for each $q \in \min \mathrm{P}_{0}$ and each $v \in \tilde{\widetilde{\mathrm{~V}}}_{\mathrm{x}} \backslash \hat{\boldsymbol{V}}_{b}$,

$$
\left.z_{q}\left(\mathbf{R}^{-\mathbf{x}^{-r}(v)}(v)\right)=z_{q}\left(\hat{\mathbf{R}}_{\mathbf{S}+\mathrm{N}}^{-\tau_{\mathrm{N}}^{(v)}(v)}\right)\right)=z_{q}\left(\hat{\mathbf{R}}_{\mathbf{N}}^{-r_{\mathbf{x}}(v)}(v)\right)
$$

by Step (ii) and Proposition io (ii); now, Corollary 9 (18) implies that $z_{q}\left(\hat{\mathbf{R}}_{\mathbf{N}}{ }^{u}(v)\right)$ is a polynomial in the variables $\mathrm{R}_{\mathrm{N}, p}(z(v), w(u, v))$ and $u \mathrm{M}_{\mathrm{N}, p}(u)$, all of whose coefficients are of the form $z_{p^{\prime}}(v), p^{\prime} \in \mathrm{P}_{0}$; moreover, the coefficients of each $\mathrm{M}_{\mathrm{N}, p}$ tend to o with $\varepsilon$, and so does each $\sup \left|\mathrm{R}_{\mathrm{N}, \mathrm{p}}\left(z(v), w\left(-r_{\mathrm{x}}(v), v\right)\right)\right|$. Therefore, we obtain (i) as Step 3 (ii).

By Step 3 (ii) and Proposition 10 (ii),
off $\hat{\mathscr{V}}_{b}$, hence, for $1 \leqslant j \leqslant n$,

$$
\begin{equation*}
x_{j} \circ \hat{\mathbf{H}}(v)=x_{j}(v)+\sum_{p \in \min \mathrm{P}_{j}}\left(r_{\mathrm{X}}(v) \mathrm{Q}_{\mathrm{N}, \mathrm{~N}^{\prime}, p}\left(r_{\mathrm{X}}(v)\right)+\mathrm{T}_{\mathrm{N}, \mathrm{~N}^{\prime}, p}\left(z(v), w\left(r_{\mathrm{X}}(v), v\right)\right)\right) x^{p}(v) \tag{22}
\end{equation*}
$$

by Corollary 1o (20). Therefore, by Step 2 (ii) and Corollary 10 (21), we just have to prove that $\mathrm{Q}_{\mathrm{N}, \mathrm{N}^{\prime}, p}=\mathrm{o}$ if $x^{p}$ does not vanish identically on $\hat{\mathscr{V}}_{b}$, which is clear: for each s.i.m. W $\subset \hat{\mathscr{V}}_{b}$ of $S$, we have indeed
for $1 \leqslant j \leqslant n$ and $v \in \mathrm{~W}$, since $\mathrm{N}=\mathrm{N}^{\prime}$ on W .
Step 5. - The connected component $\mathrm{U}_{0}$ of o in $\hat{\mathrm{H}}^{-1}\left(\dot{\mathbf{V}}_{\mathrm{Y}}\right)$ contains $\hat{\mathscr{V}}_{b} \cap \dot{\dot{\mathrm{~V}}}_{\mathrm{X}}$ and is contained in U (notation of the Complex Extension Lemma). For every $v \in \mathrm{U}_{0}\left(\hat{\mathcal{V}}_{b}\right.$, $r_{\mathrm{Y}}(\hat{\mathrm{H}}(v))=r_{\mathrm{X}}(v)$, and $\hat{\mathrm{H}}(v)=h(v) \quad$ (notation of Proposition 9).

Indeed, $\quad \hat{\mathscr{V}}_{b} \cap \dot{\mathrm{~V}}_{\mathrm{X}}=\hat{\mathscr{V}}_{b} \cap \dot{\mathrm{~V}}_{\mathrm{Y}}=\hat{\mathscr{V}}_{b} \cap \dot{\mathrm{~L}}, \quad$ hence $\quad\left(\right.$ as $\mathrm{X}=\mathrm{Y}$ on $\hat{\mathscr{V}}_{b}$ ) $\hat{\mathscr{V}}_{b} \cap \dot{\dot{V}}_{\mathrm{X}}=\hat{\mathscr{V}}_{b} \cap \dot{\mathrm{~V}}_{\mathrm{Y}}$; since this is a connected set containing o, on which $\hat{\mathrm{H}}=\mathrm{Id}$, it lies in $\mathrm{U}_{0}$.

By the same argument as in the proof of Step 4 (i), (22) yields

$$
\left(-r_{\mathrm{X}}(v), \hat{\mathbf{H}}(v)\right) \in \operatorname{dom} \hat{\mathbf{R}}_{\mathbf{N}^{\prime}} \quad \text { and } \quad \hat{\mathbf{R}}_{\mathbf{Y}}^{-r_{\mathrm{X}}(v)}(\hat{\mathbf{H}}(v))=\mathbf{R}_{\mathbf{X}}^{-r_{\mathbf{X}}(v)}(v), \quad v \in \mathrm{U}_{0} \backslash \hat{\boldsymbol{V}}_{b} .
$$

Moreover, replacing X by Y in Step 3, we get

$$
\begin{array}{r}
\left(-r_{\mathbf{Y}}(\hat{\mathbf{H}}(v)), \hat{\mathbf{H}}(v)\right) \in \operatorname{dom} \hat{\mathbf{R}}_{\mathbf{N}^{\prime}} \quad \text { and } \quad \hat{\mathbf{R}}_{\overline{\mathbf{Y}}^{-r_{\mathbf{Y}}}(\hat{\mathbf{H}}(v))}(\hat{\mathbf{H}}(v))=\mathbf{R}_{\mathbf{Y}}^{-r_{\mathbf{Y}}(\hat{\mathbf{H}}(v)\rangle}(\hat{\mathbf{H}}(v)) \\
v \in \mathrm{U}_{0} \backslash \hat{\mathscr{V}}_{b}
\end{array}
$$

Thus, if we can show that $r_{\mathrm{X}}=r_{\mathrm{Y}} \circ \hat{\mathrm{H}}$ in $\mathrm{U}_{0} \backslash \hat{\boldsymbol{V}}_{b}$, we shall have that $\mathrm{R}_{\mathrm{Y}}^{-r_{\mathrm{X}}(v)}(\hat{\mathrm{H}}(v))=\mathrm{R}_{\mathrm{X}}^{-r_{\mathrm{x}}(v)}(v)$, hence $\hat{\mathrm{H}}(v)=\mathrm{R}_{\mathrm{Y}}^{r_{\mathrm{Y}}(v)} \circ \mathrm{R}_{\mathrm{X}}^{r_{\mathrm{r}}(v)}(v)$, for each $v \in \mathrm{U}_{0} \backslash \hat{\mathscr{V}}_{b}$. By Proposition 9, this will prove our result. Now, both $r_{\mathrm{X}}$ and $r_{\mathrm{Y}} \circ \hat{\mathrm{H}}$ are analytic in the connected set $U_{0} \backslash \hat{\mathscr{V}}_{b}$, and they clearly coincide-because so do $h$ and $\hat{H}$-in a neighbourhood of V (the inclusion $\mathrm{V} \subset \mathrm{U}_{0}$ comes from the fact V and $\dot{\mathrm{V}}_{\mathrm{x}} \cap \mathscr{V}=\dot{\mathrm{V}}_{\mathrm{Y}} \cap \mathscr{V}$ are connected); therefore, $r_{\mathrm{X}}=r_{\mathrm{Y}} \circ \hat{\mathrm{H}}$ in $\mathrm{U}_{0} \backslash \hat{\boldsymbol{V}}_{b}$.

With the notation of the Complex Extension Lemma, we have proved that $U$ is an open neighbourhood of $\hat{\mathscr{V}}_{b} \backslash \dot{\mathbf{V}}_{\mathrm{X}}$ in $\dot{\mathrm{V}}_{\mathrm{X}}$ and that the mapping $\mathrm{H}: \mathrm{U} \rightarrow \mathbf{C}^{n}$, equal to Id on $\hat{\mathscr{V}}_{b}$ and to $h$ off $\hat{\mathscr{V}}_{h}$, is continuous, hence a homeomorphism onto its image. Now, we have that $\mathrm{H}^{*}(z \mathrm{Y})=z \mathrm{X}$ for every $z \in \mathbf{C}$, on and off $\hat{\mathscr{V}}_{b}$ : this is true off $\hat{\mathscr{V}}_{b}$ by Proposition 9 , and on $\hat{\mathscr{V}}_{b}$ because $\left.\mathrm{X}\right|_{\mathrm{W}}=\left.\mathrm{Y}\right|_{\mathrm{w}}$ is tangent to W for every s.i.m. $W \subset \hat{\mathscr{V}}_{b}$ of S , and $\left.\mathrm{H}\right|_{\mathrm{w}}=\left.\mathrm{Id}\right|_{\mathrm{w}}$. Therefore, the germ of H at o is a $\mathrm{C}^{0}$-conjugacy between X and Y , hence Lemma 3 .

## 6. CONCLUDING REMARKS

## (6. 1) $\mathbf{C}^{k}$-linearisations

The following result is stated as Theorem B in [DR]:
Theorem 4. - For each positive integer $k$, there is an open and dense subset $\mathrm{V}_{k}$ of $\mathfrak{g l}(n, \mathbf{C})$, the complementary subset of which has codimension one, such that every S -vector field with $\mathrm{S} \in \mathrm{V}_{k}$ is $\mathrm{C}^{k}$-conjugate to S .

Proof. - If $\mathrm{S}_{\mathbf{0}} \in \mathfrak{g l}(n, \mathbf{C})$ is weakly hyperbolic and has only simple eigenvalues, then ([Ch 86a]), for every $k \in \mathbf{N}$, there exist an integer $\ell \geqslant k$ and a neighbourhood $\mathrm{U}_{k}$ of $\mathrm{S}_{0}$ in $\mathfrak{g l}(n, \mathbf{C})$ such that, for every $\mathrm{S} \in \mathrm{U}_{k}$, the S -normal forms in the Complex Preparation Lemma (2.3) can be chosen of degree $\ell$; now, the set $\mathrm{U}_{k}^{\prime}$ of those $\mathrm{S} \in \mathrm{U}_{k}$ such that the only S-normal form of degree $\ell$ is S itself has a closed, one-codimensional complementary subset in $\mathrm{U}_{\boldsymbol{k}}$-see (5. I).

Given a positive integer $k$, let $a \in \mathbf{C}$ and $p \geqslant k$ be as in the Complex Extension Lemma (4.4) with $S=S_{0}$. As the eigenvalues, eigenspaces and maximal s.i.m.'s of $\mathrm{T} \in \mathfrak{g l}(n, \mathbf{C})$ depend analytically on T near $\mathrm{S}_{0}$, there is an open neighbourhood $\mathrm{U}^{\prime \prime}$ of $\mathrm{S}_{0}$ in $\mathfrak{g l}(n, \mathbf{C})$ such that, for each $\mathrm{T} \in \mathrm{U}^{\prime \prime} \cap \mathrm{U}_{p}^{\prime}$ and each T -vector field Z , having $p$-th order contact with $T$ along its s.i.m.'s we have the following: given a representative $\widetilde{Z}$
of $Z$, there exists an $A \in G L(n, \mathbf{C})$ such that the hypotheses of the Complex Extension Lemma are fulfiled by $\mathrm{Y}=\left.a \mathrm{~A}^{*} \widetilde{Z}\right|_{\mathrm{B}}$ and $\mathrm{X}=\left.a \mathrm{~A}^{*} \mathrm{~T}\right|_{\mathrm{B}}$; thus, every T-vector field with $\mathrm{T} \in \mathrm{U}^{\prime \prime} \cap \mathrm{U}_{p}^{\prime}$ is $\mathrm{C}^{k}$-conjugate to T , hence Theorem 4 .

Although the set $\mathrm{V}_{k}$ we obtain in this fashion is larger than that of [DR], it is still rather "small", and its definition is not really simple; the purpose of [Ch $86 c$ ] is to construct a much more reasonable $\mathrm{V}_{1}$.

## (6.2) Families

Given a diagonalisable $\mathrm{S} \in \operatorname{gl}(n, \mathbf{C})$, an $(\mathrm{S}, \boldsymbol{p})$-unfolding is a germ X at $(o, o) \in \mathbf{C}^{n} \times \mathbf{C}^{p}$ of a holomorphic vector field of the form $(x, u) \mapsto\left(\mathrm{X}_{u}(x), o\right)$, where each $X_{u}$ is a local holomorphic vector field on $\mathbf{C}^{n}$, such that $X_{0}(0)=0$ and that $S$ is the semi-simple part of $d \mathrm{X}_{0}(\mathrm{o})$. If $o$ is not an eigenvalue of S , an ( $\mathrm{S}, \mathrm{p}$ )-normal form is an (S, p)-unfolding having a representative $(x, u) \mapsto\left(\mathrm{S} x+\mathrm{N}_{u}(x), o\right)$ such that each $\mathrm{N}_{u}$ is polynomial and commutes with $S$, and $S+N_{0}$ is an S-normal form. An (S, p)unfolding and a ( $\mathrm{T}, \boldsymbol{p}$ )-unfolding are $\mathrm{C}^{k}$-conjugate, $k \in \mathbf{N}$, when there exists a $\mathrm{C}^{k}$-conjugacy $h:\left(\mathbf{C}^{n} \times \mathbf{C}^{p},(0,0)\right) 2$ between them (in the sense of (1.3)) admitting a representative of the form $(x, u) \mapsto\left(h_{u}(x) u,\right)$. The above methods yield the following result [Ch 86b]:

Theorem 5. - (i) If S is weakly hyperbolic, then, for each pair $(k, p)$ of positive integers, every ( $\mathrm{S}, \mathrm{p}$ )-unfolding is $\mathrm{C}^{k}$-conjugate to an $(\mathrm{S}, \mathrm{p})$-normal form.
(ii) Let $\left(z_{1}, \ldots, z_{n}\right)$ be the canonical coordinate system on $\mathbf{C}^{n}$. If S is hyperbolic, then, for each $p \in \mathbf{N}$, every ( $\mathbf{S}, \boldsymbol{p}$ )-unfolding is $\mathbf{C}^{\mathbf{0}}$-conjugate to a normal form represented by $(x, u) \mapsto\left(\left(c_{j}(u) z_{j}(x)\right)_{1 \leqslant j \leqslant n}, \mathbf{o}\right)$, where $c_{1}, \ldots, c_{n}:\left(\mathbf{C}^{p}, \mathbf{o}\right) \rightarrow \mathbf{C}$ are local holomorphic functions.
(iii) For each pair $(k, p)$ of positive integers and each S in the set $\mathrm{V}_{k}$ of Theorem 4, every ( $\mathrm{S}, \mathrm{p}$ )-unfolding is $\mathrm{C}^{k}$-conjugate to a normal form of the same type as in (ii).

Of course, (i), (ii) and (iii) are generalisations of Theorems 2, i and 4 respectively. In [Ch $86 c$ ], we prove (iii) for $k=1$ and a reasonable $\mathrm{V}_{1}$; as the germs of $c_{1}, \ldots, c_{n}$ are $\mathrm{C}^{1}$-conjugacy invariants, this provides universal unfoldings ([A]) for $\mathrm{C}^{1}$-conjugacy. Noticing that, as in the proof of Theorem 4, the degree of the ( $\mathrm{S}, \mathrm{p}$ ) -normal forms in (i) with respect to the $\mathbf{C}^{n}$-variable has a bound which depends only on S , we obtain versal unfoldings for $\mathrm{C}^{k}$-conjugacy, under the sole weak hyperbolicity hypothesis. The problem of finding universal unfoldings in this general case seems very difficult—see [Ch 86c].

If we replace $\mathrm{C}^{k}$-conjugacy by (holomorphic) conjugacy, (i) and (iii) are true in the Poincaré domain, where the set which corresponds to $V_{k}$ is simply the set of those $S$ which have only simple eigenvalues and for which $\mathrm{P}_{1}, \ldots, \mathrm{P}_{n}$ are empty-see [A], $\S 36$, C. In the Siegel domain, there is no hope for such results-which are false even at the formal level; this is why Theorem 5 is interesting.

## (6.3) Historical and technical comments

The proof of Theorem 2 originated in an attempt to understand and generalise the excellent-and curiously underrated-work of Dumortier and Roussarie ([DR]). My main contributions are the Complex Preparation Lemma-which would have simplified their proof of Theorem 4-the weakening of Hyperbolicity, and the statement and proof of a general normal form theorem instead of a linearisation result. The " Lyapunov function " $\mathbf{F}$, introduced in [C], is not really necessary here (it is not used in [DR]), but-besides being crucial in the proof of Theorem 1 -it makes everything twork just as well for general smooth germs of $\mathbf{Z}^{k} \times \mathbf{R}^{m}$-actions, yielding the generalisation of [DR] I was aiming for (see Chapter 3 of [Ch 86] and [Ch 86d]).

The Complex Preparation Lemma (2.3) and the Extension Lemma (2.2) came from a geometric reading of Nelson's nice (almost) proof of Sternberg's theorem ([N]). A difference with [ N ] is that I localise everything in "isolating blocks", which makes the situation geometrically clearer-but technically worse, due to problems of definition domains; this formulation allowed the direct study of holomorphic flows in Secion 4 -otherwise, they should have been extended to global smooth $\mathbf{R}^{2}$-actions (as in Section 6 of [Ch 86]).

The author is entirely responsible for the rather simple-minded proof of Theorem I . In [Ch $86 c$ ], similarly, $\mathrm{C}^{1}$-conjugacies are constructed as the sums of convergent power series in variables which are functions of class $\mathrm{C}^{\mathbf{1}}$.

MAIN NOTATION

| Notation | Section | Page | Notation | Section | Page |
| :---: | :---: | :---: | :---: | :---: | :---: |
| d, S, $\mathrm{X}^{1}$ (linear part of X ) | (1.1) | 144 | V, $\mathrm{B}_{\mathrm{c}, \mathrm{I}}$ | (4.3) | 163 |
| $f_{\mathbf{X}}, f_{\mathbf{X}}^{t}$ (real flow of $\mathbf{X}$ ) | (2.2) | 150 | $\mathrm{X}_{c}, \tilde{\mathrm{~V}}_{\mathrm{X}}, \tilde{\tilde{V}}_{\mathbf{X}}, \mathscr{C}_{\mathrm{X}, \mathrm{I}}, \mathrm{X}_{\mathrm{I}}$ | - | 163 |
| $\boldsymbol{r}$ | (2.3) | 154 | $\mathrm{L}_{\mathbf{X}}$ (Lie derivative), $\mathrm{C}_{\mathbf{X}}, \mathrm{J}$ | - | ${ }_{165}$ |
| F | ) | 155 |  |  | ${ }^{163}$ |
| $c_{j}, x_{j}(1 \leqslant j \leqslant n)$ | (3.1) | 155 | $\mathscr{N}^{\prime}, \mathscr{N}_{0}$ | - | $1{ }_{165}$ |
| $\sigma, a, \sigma_{1}, \mathrm{~F}_{1}$ | - | 155 | $\mathrm{C}, \mathrm{R}_{\mathbf{x}}$ | -- | 165 |
| $\mathrm{W}_{0}^{ \pm}, \mathrm{O}_{x_{j}}, \mathrm{Q}_{c}, \mathrm{E}_{c}$ | (3.2) | 156 | $\varphi_{\mathrm{X}}, \Omega_{\mathbf{X}}, g_{\mathbf{X}}, r_{\mathbf{X}}, \gamma_{\mathbf{X}, v}, s_{\mathbf{X}}$ | - | 166 |
| $\varphi_{c}, \Phi_{c}^{t}, g_{c}, \xi_{c}$ | (3.2) | ${ }_{1} 56$ | $t_{\mathbf{X}}$ | - | 166 |
| $\left.\begin{array}{l} \|v\|, \mathrm{E}_{\mathrm{I}}, \mathscr{K}, g_{\mathrm{I}}, \mathrm{I}_{0}^{ \pm}, \mathscr{I}_{c} \\ \Sigma_{c}, \mathrm{I}, \mathrm{E}_{c, \mathrm{I}}, \mathrm{Q}_{c, \mathrm{I}}, v_{\mathrm{I}}, \mathrm{I}^{ \pm} \end{array}\right\}$ | (3.3) | 156-7 | $\begin{aligned} & \dot{\mathrm{V}}_{\mathbf{X}},{\dot{\mathrm{V}_{\mathbf{X}}}}^{\omega_{\mathbf{X}}, \Phi_{\mathbf{X}}, h, \mathscr{N}^{\prime \prime}, \mathrm{U}, \mathrm{U}^{\prime}, \mathrm{H}} \end{aligned}$ | $\begin{aligned} & (4 \cdot 4) \\ & (4 \cdot 4) \end{aligned}$ | 168 169 |
| $\mathscr{C}_{c}, \mathrm{I}, \mathrm{W}_{c}^{ \pm}, \mathrm{I}, \xi_{c, \mathrm{I}}, \mathrm{f}_{c}, \Sigma_{c}$ | - | 157-8 | $\widehat{\mathbf{R}}_{\mathbf{X}}$, dom $\mathbf{T}$ (strict convergence |  |  |
| $Q^{\prime}{ }_{b}, \hat{\mathscr{V}}_{b}$ | - | I59 | domain of a power series T ), $\mathrm{P}_{\boldsymbol{j}}$ | (5.1) | 172 |
| $\mathrm{E}_{\boldsymbol{b}}$ | - | 160 | $\mathrm{P}_{0}, \min \mathrm{P}_{j}, z, w, \mathrm{M}_{\mathrm{N}, p}$, |  |  |
| B, L | (4. 1) | 161 | $\mathrm{R}_{\mathrm{N}, p}$ ( ${ }^{\text {r }}$ | - | 173 |
| V, $\widetilde{\mathrm{V}}$ | (4.2) | 16I | $Q_{N, ~} \mathbf{N}^{\prime}, p, \mathrm{~T}_{\mathrm{N}, \mathrm{N}^{\prime}, \boldsymbol{p}}$ | - | 174 |

## DEFINITIONS AND MAIN RESULTS

| Definitions | Section | Page | Definitions | Section | Page |
| :---: | :---: | :---: | :---: | :---: | :---: |
| conjugate | (1.1) | 144 | Proposition 6-7 | (3.3) | 159 |
| $\mathrm{C}^{k}$-conjugate | (1.3) | 145 | 8 | (4.2) | 161 |
| Corollary 1 | - | 146 | 9 | (4.4) | 169 |
| - 2 | (2.2) | 153 | 10 | (5.1) | 172 |
| 3 | (3.2) | 156 | - 11 | - | 173 |
| 4 | - | 156 | 12 | (5.2) | 175 |
| - 5 | (3.3) | 159 | quotient | (2.2) | 153 |
| 6 | - | 160 | reduced S-normal form | (5.2) | 175 |
| 7 | (4.4) | 171 | S-vector field | (I. 1 ) |  |
| - 8-9 | (5.1) | 173 | S-normal form | (1.1) | 144 |
| 10 | - | 174 | X-saturated | (2.2) | 150 |
| $\mathrm{C}^{k}$-equivalent | (1.3) | 145 | Siegel domain | (1.2) | 145 |
| (weakly) hyperbolic | - | 146 | special S-vector field | (5.1) | 173 |
| Lemmas: |  |  | stable subspace | (2.1) | 147 |
| Preparation Lemma | (2.1) | 147 | strict convergence domain | (5.1) | 172 |
| Isolating Block - | (2.2) | 148 | strongly invariant manifold | (2.3) | 154 |
| Extension - | - | 150 | Theorems: |  |  |
| Complex Preparation - | (2.3) | 154 | Poincaré-Dulac Theorem | (1.2) | 144 |
| Complex Isolating Block - | (4.3) | \{ 163 | Siegel linearisation - | (1.2) | 145 |
| Complex Isolating Block - | (4.3) | $1{ }^{6} 5$ | Theorems 1-2 | (1.3) | 146 |
| Complex Extension - | (4.4) | 169 | Sternberg's Theorem | (2.1) | 147 |
| Poincaré domain | (1.2) | 144 | Theorem 3 | (5.2) | 175 |
| Proposition I | (3.2) | 156 | 4 | (6.1) | 179 |
| 2 | (3.3) | 156 | - 5 | (6.2) | 180 |
| 3-4-5 | - | 158 | unstable subspace | (2.1) | 147 |

## REFERENCES

[A] V. I. Arnol'd, Chapitres supplémentaires de la théorie des équations différentielles ordinaires, Moscou, Mir, 1980.
[C] C. Самacho, On $\mathbf{R}^{k} \times \mathbf{Z}^{l}$-actions, in Differentiable Dynamical Systems, M. M. Perxoto ed., IMPA, Academic Press, 1973.
[GKP] G. Camacho, N. Kuiper, J. Palis, The topology of holomorphic flows with singularity, Publ. Math. I.H.E.S., 48 (1978), 5-38.
[Ch 8o] M. Chaperon, Propriétés génériques des germes d'actions différentiables de groupes de Lie commutatifs élémentaires, thèse, Université Paris 7, 1980.
[Ch 85] M. Chaperon, Differential geometry and dynamics: two examples, in Singularities and Dynamical Systems, S. N. Pnevmatikos ed., Elsevier B. V. (North Holland), 1985, 187-207.
[Ch 86] M. Chaperon, Géométrie différentielle et singularités de systèmes dynamiques, Astérisque, 138-1 39 (1986).
[Ch 86a] M. Chaperon, Invariant manifolds and a preparation lemma for complex flows near a singularity, Preprint, École, Polytechnique, 1986.
[Ch 86b] M. Chaperon, $\mathrm{C}^{k}$-versal unfoldings of holomorphic fows near their singularities, Preprint, École Polytechnique 1986.
[Ch 86c] M. Chaperon, Ć-linearisable local holomorphic flows, to appear.
[Ch 86d] M. Chaperon, Smooth conjugacy results for germs of smooth $\mathbf{Z}^{k} \times \mathbf{R}^{m}$-actions preserving various structures, to appear.
[CZ] C. C. Conley, E. Zehnder, The Birkhoff-Lewis fixed point theorem and a conjecture of V. I. Arnol'd, Invent. Math., 73 (1983), 33-49.
[DR] F. Dumortier, R. Roussarie, Smooth linearisation of germs of $\mathbf{R}^{2}$-actions and holomorphic vector fields, Ann. Inst. Fourier, 30 (1) (1980), 31-64.
[G] J. Guckenheimer, Hartman's theorem for complex flows in the Poincaré domain, Composition Math., 24 (1) (1972).
[I] Ju. S. Il'Iašenko, Global and local aspects of the theory of complex differential equations, Proceedings of the International Congress of Mathematicians, Helsinki, 1978, 821-826.
[LM] S. López de Medrano, Topology of the intersection of quadrics in $\mathbf{R}^{n}$, Preprint, Universidad Nacional Autónoma de México, 1986.
[MR] J. Martinet, J. P. Ramis, Problèmes de modules pour des équations différentielles non linéaires du premier ordre, Publ. Math. I.H.E.S., 55 (1982), 63-164.
[N] E. Nelson, Topics in dynamics, Part I, Flows, Princeton, 1970.

Centre de mathématiques, U.A. no ${ }^{0} 69$ du C.N.R.S.
Ecole polytechnique,
91128 Palaiseau Cedex (France).

Manuscrit reçu le 10 octobre 1986.


[^0]:    (1) The degree of which admits an upper bound depending only (and nicely) on S and $k$.

[^1]:    ( ${ }^{1}$ ) Our results can be extended to the case when $Q$ is the direct sum of two arbitrary riemannian vector bundles $\mathrm{W}^{+}$and $\mathrm{W}^{-}$over M .

[^2]:    ${ }^{1}$ ) Restricted to the complementary subset of the unstable subspace of $a \mathrm{~S}$

[^3]:    ${ }^{(1)}$ Viewing $\hat{\mathbf{R}}_{\mathrm{N}}^{u}$ as a germ at $\mathrm{o} \in \mathbf{C}^{n}$ of a holomorphic diffeomorphism.

