# Subelliptic wave equations are never observable 

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# Subelliptic wave equations are never observable 

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#### Abstract

It is well-known that observability (and, by duality, controllability) of the elliptic wave equation, i.e., with a Riemannian Laplacian, in time $T_{0}$ is almost equivalent to the Geometric Control Condition (GCC), which stipulates that any geodesic ray meets the control set within time $T_{0}$. We show that in the subelliptic setting, GCC is never verified, and that subelliptic wave equations are never observable in finite time. More precisely, given any subelliptic Laplacian $\Delta=-\sum_{i=1}^{m} X_{i}^{*} X_{i}$ on a manifold $M$ such that $\operatorname{Lie}\left(X_{1}, \ldots, X_{m}\right)=T M$ but $\operatorname{Span}\left(X_{1}, \ldots, X_{m}\right) \subsetneq T M$, we show that for any $T_{0}>0$ and any measurable subset $\omega \subset M$ such that $M \backslash \omega$ has nonempty interior, the wave equation with subelliptic Laplacian $\Delta$ is not observable on $\omega$ in time $T_{0}$. The proof is based on the construction of sequences of solutions of the wave equation concentrating on spiraling geodesics (for the associated subRiemannian distance) spending a long time in $M \backslash \omega$. As a counterpart, we prove a positive result of observability for the wave equation in the Heisenberg group, where the observation set is a well-chosen part of the phase space.


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## 1 Introduction

### 1.1 Setting

This article focuses on the wave equation in sub-Riemannian manifolds, i.e., on subelliptic wave equations. Let $n \in \mathbb{N}^{*}$ and let $M$ be a smooth connected compact manifold of dimension $n$ with a non-empty boundary $\partial M$. We consider a smooth horizontal distribution $\mathcal{D}$ on $M$, i.e., a smooth assignment $M \ni x \mapsto \mathcal{D}_{x} \subset T_{x} M$ (possibly with non-constant rank), and a Riemannian metric $g$ on $\mathcal{D}$. We also assume that $\mathcal{D}$ satisfies the Hörmander condition

$$
\begin{equation*}
\operatorname{Lie}(\mathcal{D})=T M \tag{1}
\end{equation*}
$$

(see Mon02]). The triple $(M, \mathcal{D}, g)$ is called a sub-Riemannian structure. Additionally, we make the important assumption that the set of all $x \in M$ such that $\mathcal{D}_{x} \neq T_{x} M$ is dense in $M$; in other words, $(M, \mathcal{D}, g)$ is nowhere Riemannian. Finally, we assume that $M$ is endowed with a smooth volume $\mu$.

We consider the sub-Riemannian Laplacian $\Delta_{g, \mu}$ on $L^{2}(M, \mu)$, which only depends on $g$ and $\mu$, defined by

$$
\Delta_{g, \mu}=-\sum_{i=1}^{m} X_{i}^{*} X_{i}=\sum_{i=1}^{m} X_{i}^{2}+\operatorname{div}_{\mu}\left(X_{i}\right) X_{i}
$$

where $\left(X_{i}\right)_{1 \leqslant i \leqslant m}$ denotes a local $g$-orthonormal frame such that $\mathcal{D}=\operatorname{Span}\left(X_{1}, \ldots, X_{m}\right)$ and the star designates the transpose in $L^{2}(M, \mu)$. The divergence with respect to $\mu$ is defined by $L_{X} \mu=\left(\operatorname{div}_{\mu} X\right) \mu$, where $L_{X}$ stands for the Lie derivative. Then $\Delta_{g, \mu}$ is hypoelliptic (see [Hör67]). In order to simplify notations, we set $\Delta=\Delta_{g, \mu}$ in the sequel, since $g$ and $\mu$ are fixed once for all.

We consider $\Delta$ with Dirichlet boundary conditions and the domain $D(\Delta)$ which is the completion in $L^{2}(M, \mu)$ of the set of all $u \in C_{c}^{\infty}(M)$ for the norm $\|(\operatorname{Id}-\Delta) u\|_{L^{2}}$.

Consider the wave equation

$$
\left\{\begin{array}{l}
\partial_{t t}^{2} u-\Delta u=0 \text { in }(0, T) \times M  \tag{2}\\
u=0 \text { on }(0, T) \times \partial M \\
\left(u_{\mid t=0}, \partial_{t} u_{\mid t=0}\right)=\left(u_{0}, u_{1}\right)
\end{array}\right.
$$

where $T>0$, and the initial data $\left(u_{0}, u_{1}\right)$ are in an appropriate energy space. The natural energy of a solution is

$$
E(u(t, \cdot))=\frac{1}{2} \int_{M}\left(\left|\partial_{t} u(t, x)\right|^{2}+\left|\nabla^{\mathrm{sR}} u(t, x)\right|^{2}\right) d \mu(x)
$$

where, for any $\phi \in C^{\infty}(M)$,

$$
\nabla^{s R} \phi=\sum_{i=1}^{m}\left(X_{i} \phi\right) X_{i}
$$

is the horizontal gradient. Note that $\nabla^{\mathrm{sR}}$ is the formal adjoint of $\left(-\operatorname{div}_{\mu}\right)$ in $L^{2}(M, \mu)$.
We denote by $\mathcal{H}(M)$ the completion of $C_{c}^{\infty}(M)$ for the norm

$$
\|v\|_{\mathcal{H}}=\left(\int_{M}\left|\nabla^{\mathrm{sR}} v(x)\right|^{2} d \mu(x)\right)^{\frac{1}{2}} .
$$

and the space of initial data of (2) is naturally endowed with the norm

$$
\left\|\left(u_{0}, u_{1}\right)\right\|_{\mathcal{H} \times L^{2}}=\left(\left\|u_{0}\right\|_{\mathcal{H}}^{2}+\left\|u_{1}\right\|_{L^{2}(M, \mu)}^{2}\right)^{\frac{1}{2}}
$$

It is well-known (see for example GR15, Theorem 2.1], EN99, Section II.6]) that for any $\left(u_{0}, u_{1}\right) \in \mathcal{H}(M) \times L^{2}(M)$, there exists a unique solution

$$
u \in C^{0}(0, T ; \mathcal{H}(M)) \cap C^{1}\left(0, T ; L^{2}(M)\right)
$$

to (22) (in a mild sense). Moreover, if $u$ is a solution of (2), then

$$
\frac{d}{d t} E(u(t, \cdot))=0 .
$$

In this paper, we investigate exact observability for the wave equation (2).
Definition 1. Let $T_{0}>0$ and $\omega \subset M$ be a $\mu$-measurable subset. The subelliptic wave equation (21) is exactly observable on $\omega$ in time $T_{0}$ if there exists a constant $C_{T_{0}}(\omega)>0$ such that, for any $\left(u_{0}, u_{1}\right) \in \mathcal{H}(M) \times L^{2}(M)$, the solution $u$ of (2) satisfies

$$
\begin{equation*}
\int_{0}^{T_{0}} \int_{\omega}\left|\partial_{t} u(t, x)\right|^{2} d \mu(x) d t \geqslant C_{T_{0}}(\omega)\left\|\left(u_{0}, u_{1}\right)\right\|_{\mathcal{H} \times L^{2}}^{2} \tag{3}
\end{equation*}
$$

### 1.2 Main result

Our main result is the following.
Theorem 1. Let $T_{0}>0$ and let $\omega \subset M$ be a measurable subset such that $M \backslash \omega$ has nonempty interior. Then the subelliptic wave equation (2) is not exactly observable on $\omega$ in time $T_{0}$.

Consequently, using a duality argument (see Section 4.2), we obtain that exact controllability does not hold either in any finite time.

Definition 2. Let $T_{0}>0$ and $\omega \subset M$ be a measurable subset. The subelliptic wave equation (2) is exactly controllable on $\omega$ in time $T_{0}$ if for any $\left(u_{0}, u_{1}\right) \in \mathcal{H}(M) \times L^{2}(M)$, there exists $g \in L^{2}\left(\left(0, T_{0}\right) \times M\right)$ such that the solution $u$ of

$$
\left\{\begin{array}{l}
\partial_{t t}^{2} u-\Delta u=\mathbf{1}_{\omega} g \quad \text { in }\left(0, T_{0}\right) \times M  \tag{4}\\
u=0 \text { on }\left(0, T_{0}\right) \times \partial M, \\
\left(u_{\mid t=0}, \partial_{t} u_{\mid t=0}\right)=\left(u_{0}, u_{1}\right)
\end{array}\right.
$$

satisfies $u\left(T_{0}, \cdot\right)=0$.
Corollary 1. Let $T_{0}>0$ and let $\omega \subset M$ be a measurable subset such that $M \backslash \omega$ has nonempty interior. Then the subelliptic wave equation (2) is not exactly controllable on $\omega$ in time $T_{0}$.

Remark 3. Theorem 1 holds under the two assumptions that $\mathcal{D}$ satisfies the Hörmander condition (1) and that the set of $x \in M$ such that $\mathcal{D}_{x} \neq T_{x} M$ is dense in $M$. However, inspecting the proof, we see that the conclusion of Theorem 1 also holds under the weaker assumption that the set of $x \in M$ such that $\mathcal{D}_{x} \subsetneq \mathcal{D}_{x}+\left[\mathcal{D}_{x}, \mathcal{D}_{x}\right]$ is dense in $M$.

In the statement of Theorem [1, we assumed that the sub-Riemannian structure ( $M, \mathcal{D}, g$ ) verifies $\mathcal{D}_{x} \neq T_{x} M$ for a dense set of $x \in M$. Let us explain how to adapt this result to the case of almost-Riemannian structures, i.e., sub-Riemannian structures which do not necessarily verify this assumption. A typical example is the Grushin case, for which $X_{1}=\partial_{x_{1}}$ and $X_{2}=x_{1} \partial_{x_{2}}$ are vector fields on $(-1,1)_{x_{1}} \times \mathbb{T}_{x_{2}}$. Then $\operatorname{rank}\left(\mathcal{D}_{x}\right)$ is equal to 1 for $x_{1}=0$ and to 2 otherwise.

Theorem 2. Let $T_{0}>0$ and let $\omega \subset M$ be a measurable set such that $M \backslash \omega$ has an interior which is non-empty and which moreover contains a point $x$ such that $\mathcal{D}_{x} \neq T_{x} M$. Then the subelliptic wave equation (2) is not exactly observable on $\omega$ in time $T_{0}$.

Remark 4. In the Grushin case, the corresponding Laplacian is elliptic outside of the singular submanifold $S=\left\{x_{1}=0\right\}$. According to the above result, in the Grushin case the subelliptic wave equation is observable on any subset containing $S$ (with some finite minimal time of observability, see [BLR92]], but is not observable in any finite time on any subset $\omega$ such that the interior of $M \backslash \omega$ has a non-empty intersection with $S$.
Remark 5. The assumption of compactness on $M$ is not necessary: we may remove it, and just require that the subelliptic wave equation (21) in $M$ is well-posed. It is for example the case if $M$ is complete for the sub-Riemannian distance induced by $g$ since $\Delta$ is then essentially self-adjoint ([Str86]).
Remark 6. Theorem 1 remains true if $M$ has no boundary. In this case, since non-zero constant functions on $M$ are solutions of (22), one usually requires for the observability inequality (3) to hold only for solutions with initial data ( $u_{0}, u_{1}$ ) verifying $\int_{M} u_{0} d \mu=0$ (i.e., we quotient by constant functions). Then, Theorem 1 remains true: anticipating the proof, we notice that the spiraling geodesics of Proposition 12 still exist (since their construction is purely local), and we subtract to the initial datum $u_{0}^{k}$ of the localized solutions constructed in Proposition 11 their spatial average $\int_{M} u_{0}^{k} d \mu$.

### 1.3 Ideas of the proof

The proof of Theorem 1 mainly requires two ingredients:

1. There exist solutions of the free subelliptic wave equation (2) whose energy concentrates along any given (normal) geodesic of $(M, \mathcal{D}, g)$;
2. There exist normal geodesics of $(M, \mathcal{D}, g)$ which "spiral" around curves transverse to $\mathcal{D}$, and which therefore remain arbitrarily close to their starting point on arbitrarily large-time intervals.
Combining these two facts, the proof of Theorem 1 is straightforward (see Section 4.1). Note that the first point follows from the general theory of propagation of complex Lagrangian spaces, while the second point is the main novelty of this paper.

Since our construction is purely local (meaning that it does not "feel" the boundary and only relies on the local structure of the vector fields), we can focus on the case where there is a (small) open neighborhood $V$ of the origin such that $V \subset M \backslash \omega$. In the sequel, we assume it is the case.

Let us give an example of sub-Riemannian structure where the spiraling geodesics used in the proof of Theorem 1 are particularly simple. We consider the three-dimensional manifold with boundary $M_{1}=(-1,1)_{x_{1}} \times \mathbb{T}_{x_{2}} \times \mathbb{T}_{x_{3}}$, where $\mathbb{T}=\mathbb{R} / \mathbb{Z} \approx(-1,1)$ is the 1 D torus. We endow $M_{1}$ with the vector fields $X_{1}=\partial_{x_{1}}$ and $X_{2}=\partial_{x_{2}}-x_{1} \partial_{x_{3}}$ and we set $\mathcal{D}_{1}=$ $\operatorname{Span}\left(X_{1}, X_{2}\right)$, with the metric $g_{1}$ being defined by the fact that $\left(X_{1}, X_{2}\right)$ is a $g_{1}$-orthonormal frame of $\mathcal{D}_{1}$. Then, $\left(M_{1}, \mathcal{D}_{1}, g_{1}\right)$ is a sub-Riemannian structure, which we will call in the sequel the "Heisenberg manifold with boundary". We endow it with an arbitrary smooth volume $\mu$. The geodesics we consider are given by

$$
\begin{aligned}
& x_{1}(t)=\varepsilon \sin (t / \varepsilon) \\
& x_{2}(t)=\varepsilon \cos (t / \varepsilon)-\varepsilon \\
& x_{3}(t)=\varepsilon(t / 2-\varepsilon \sin (2 t / \varepsilon) / 4) .
\end{aligned}
$$

They spiral around the $x_{3}$ axis $x_{1}=x_{2}=0$.
Here, one should think of $\varepsilon$ as a small parameter. In the sequel, we denote by $x_{\varepsilon}$ the geodesic with parameter $\varepsilon$.

Clearly, given any $T_{0}>0$, for $\varepsilon$ sufficiently small, we have $x_{\varepsilon}(t) \in V$ for every $t \in\left(0, T_{0}\right)$. Our objective is to construct solutions $u^{k}$ of the subelliptic wave equation (2) such that $\left\|\left(u_{0}^{k}, u_{1}^{k}\right)\right\|_{\mathcal{H} \times L^{2}}=1$ and the energy of $u^{k}(t, \cdot)$ outside of a ball $B_{g_{1}}\left(x_{\varepsilon}(t), r_{k}\right)$ centered at $x_{\varepsilon}(t)$ and with small radius $r_{k}>0$

$$
\int_{M_{1} \backslash B_{g_{1}}\left(x_{\varepsilon}(t), r_{k}\right)}\left(\left|\partial_{t} u^{k}(t, x)\right|^{2}+\left|\nabla^{\mathrm{sR}} u^{k}(t, x)\right|^{2}\right) d \mu(x)
$$

tends to 0 as $k \rightarrow+\infty$ uniformly with respect to $t \in\left(0, T_{0}\right)$. As a consequence, the observability inequality (3) fails.

The construction of solutions of the free wave equation whose energy concentrates on geodesics is classical in the elliptic (or Riemannian) case: these are the so-called Gaussian beams, for which a construction may be found for example in Ral82. Here, we adapt this construction to our subelliptic (sub-Riemannian) setting, which does not raise any problem since the geodesics we consider stay in the elliptic part of the operator $\Delta$. It may also be directly justified with the theory of propagation of complex Lagrangian spaces (see Section (2).

In the general case where $(M, \mathcal{D}, g)$ is not necessarily the Heisenberg manifold without boundary, the existence of spiraling geodesics also has to be justified. For that purpose, we first approximate $(M, \mathcal{D}, g)$ by its nilpotent approximation, and we then prove that in the latter, it is possible to identify a "Heisenberg sub-structure", which gives the desired spiraling geodesics.

### 1.4 Sub-Riemannian geodesics

In this section, we recall a few basic facts about sub-Riemannian geodesics. In this paper, we just need to focus on normal geodesics, which are the natural extension of Riemannian geodesics since they are projections of bicharacteristics. Recall that there may also exist abnormal geodesics (see Mon94), but we did not address the problem of constructing solutions of (2) concentrating on these geodesics since it is not useful for our purpose.

We denote by $S_{\mathrm{phg}}^{m}\left(T^{*}((0, T) \times M)\right)$ the set of polyhomogeneous symbols of order $m$ with compact support and by $\Psi_{\mathrm{phg}}^{m}((0, T) \times M)$ the set of associated polyhomogeneous pseudodifferential operators of order $m$ whose distribution kernel has compact support in $(0, T) \times M$ (see Appendix A).

We set $P=\partial_{t t}^{2}-\Delta \in \Psi_{\mathrm{phg}}^{2}((0, T) \times M)$, whose principal symbol is

$$
p_{2}(t, \tau, x, \xi)=-\tau^{2}+g^{*}(x, \xi)
$$

with $\tau$ the dual variable of $t$ and $g^{*}$ the principal symbol of $-\Delta$. For $\xi \in T^{*} M$, we have (see Appendix (A)

$$
g^{*}(x, \xi)=\sum_{i=1}^{m} h_{X_{i}}^{2}
$$

Here, given any smooth vector field $X$ on $M$, we denoted by $h_{X}$ the Hamiltonian function (momentum map) on $T^{*} M$ associated with $X$ defined in local $(x, \xi)$-coordinates by $h_{X}(x, \xi)=\xi(X(x))$. Then $g^{*}$ is both the principal symbol of $\Delta$, and also the cometric associated with $g$. Equivalently, $g^{*}(x, \xi)=\sup \left\{\xi(v)^{2} \mid v \in T_{x} M, g_{x}(v)=1\right\}$.

In $T^{*}(\mathbb{R} \times M)$, the Hamiltonian vector field $\vec{H}_{p_{2}}$ associated with $p_{2}$ is given by $\vec{H}_{p_{2}} f=$ $\left\{p_{2}, f\right\}$. Since $\vec{H}_{p_{2}} p_{2}=0$, we get that $p_{2}$ is constant along the integral curves of $\vec{H}_{p_{2}}$. Thus, the characteristic set $\mathcal{C}\left(p_{2}\right)=\left\{p_{2}=0\right\}$ is preserved by the flow of $\vec{H}_{p_{2}}$. Null-bicharacteristics are then defined as the maximal integral curves of $\vec{H}_{p_{2}}$ which live in $\mathcal{C}\left(p_{2}\right)$. In other words, the null-bicharacteristics are the maximal solutions of

$$
\left\{\begin{array}{l}
\dot{t}(s)=-2 \tau(s)  \tag{5}\\
\dot{x}(s)=\nabla_{\xi} g^{*}(x(s), \xi(s)) \\
\dot{\tau}(s)=0 \\
\dot{\xi}(s)=-\nabla_{x} g^{*}(x(s), \xi(s)) \\
\tau^{2}(0)=g^{*}(x(0), \xi(0))
\end{array}\right.
$$

This definition needs to be adapted when the null-bicharacteristic meets the boundary $\partial M$, but in the sequel, we only consider solutions of (5) on time intervals where $x(t)$ does not reach $\partial M$.

In the sequel, we take $\tau=-1 / 2$, which gives $g^{*}(x(s), \xi(s))=1 / 4$. This also implies that $t(s)=s+t_{0}$ and, taking $t$ as a time parameter, we are led to solve

$$
\left\{\begin{array}{l}
\dot{x}(t)=\nabla_{\xi} g^{*}(x(t), \xi(t)),  \tag{6}\\
\dot{\xi}(t)=-\nabla_{x} g^{*}(x(t), \xi(t)), \\
g^{*}(x(0), \xi(0))=\frac{1}{4}
\end{array}\right.
$$

Remark 7. Note that in the subelliptic setting, the co-sphere bundle $S^{*} M$ may be decomposed as $S^{*} M=U^{*} M \cup S \Sigma$, where $U^{*} M=\left\{g^{*}=1\right\}$ is a cylinder bundle, $\Sigma=\left\{g^{*}=0\right\}$ is the characteristic cone and $S \Sigma$ is a twofold covering of $M$. As explained in CdVHT18, Section 1], each fiber of $S^{*} M$ is obtained by compactifying a cylinder with two points at infinity.

We denote by $\phi_{t}: S^{*} M \rightarrow S^{*} M$ the (normal) geodesic flow defined by $\phi_{t}\left(x_{0}, \xi_{0}\right)=$ $(x(t), \xi(t))$, where $(x(t), \xi(t))$ is a solution of the system given by the first two lines of (6) and initial conditions $\left(x_{0}, \xi_{0}\right)$. Note that any point in $S \Sigma$ is a fixed point of $\phi_{t}$.

The curves $x(t)$ which solve (6) are the geodesics for the sub-Riemannian metric $g$. In other words, the projections of the null-bicharacteristics onto $M$, using the variable $t$ as a parameter, coincide with the geodesics on $M$ associated with the metric sub-Riemannian $g$ traveled at speed one.

### 1.5 Observability in some regions of phase-space

We have explained in Section 1.3 that the existence of solutions of the subelliptic wave equation (2) concentrated on spiraling geodesics is an obstruction to observability in Theorem 1. Our goal in this section is to state a result ensuring observability if one "removes" in some sense these geodesics.

For this result, we focus on a version of the Heisenberg manifold described in Section 1.3 which has no boundary. This technical assumption avoids us using boundary microlocal defect measures in the proof, which, in this sub-Riemannian setting, are difficult to handle. As a counterpart, the observability inequality is written for functions with null initial average, since otherwise constant solutions would not be observable.

We consider the Heisenberg group $G$, that is $\mathbb{R}^{3}$ with the composition law

$$
\left(x_{1}, x_{2}, x_{3}\right) \star\left(x_{1}^{\prime}, x_{2}^{\prime}, x_{3}^{\prime}\right)=\left(x_{1}+x_{1}^{\prime}, x_{2}+x_{2}^{\prime}, x_{3}+x_{3}^{\prime}-x_{1} x_{2}^{\prime}\right)
$$

Then $X_{1}=\partial_{x_{1}}$ and $X_{2}=\partial_{x_{2}}-x_{1} \partial_{x_{3}}$ are left invariant vector fields on $G$. Since $\Gamma=$ $\sqrt{2 \pi} \mathbb{Z} \times \sqrt{2 \pi} \mathbb{Z} \times 2 \pi \mathbb{Z}$ is a co-compact subgroup of $\mathbb{R}^{3}$, the left quotient $M_{H}=\Gamma \backslash G$ is a compact three dimensional manifold and, moreover, $X_{1}$ and $X_{2}$ are well-defined as vector fields on the quotient. Finally, we define the Heisenberg Laplacian $\Delta_{H}=X_{1}^{2}+X_{2}^{2}$ on $M_{H}$. Since $\left[X_{1}, X_{2}\right]=-\partial_{x_{3}}$, it is a hypoelliptic operator. We set $\mathcal{D}_{H}=\operatorname{Span}\left(X_{1}, X_{2}\right)$, with the metric $g_{H}$ being defined by the fact that $\left(X_{1}, X_{2}\right)$ is a $g_{H}$-orthonormal frame of $\mathcal{D}_{H}$. Then, $\left(M_{H}, \mathcal{D}_{H}, g_{H}\right)$ is a sub-Riemannian structure, which we call the "Heisenberg manifold without boundary". We endow ( $M_{H}, \mathcal{D}_{H}, g_{H}$ ) with an arbitrary smooth volume $\mu$.

We note that $g^{*}(x, \xi)=\xi_{1}^{2}+\left(\xi_{2}-x_{1} \xi_{3}\right)^{2}$ and hence the null-bicharacteristics are solutions of

$$
\begin{aligned}
& \dot{x}_{1}(t)=2 \xi_{1} \\
& \dot{x}_{2}(t)=2\left(\xi_{2}-x_{1} \xi_{3}\right), \\
& \dot{x}_{3}(t)=-2 x_{1}\left(\xi_{2}-x_{1} \xi_{3}\right),
\end{aligned}
$$

$$
\dot{\xi}_{1}(t)=2 \xi_{3}\left(\xi_{2}-x_{1} \xi_{3}\right)
$$

$$
\dot{\xi}_{2}(t)=0
$$

$$
\dot{\xi}_{3}(t)=0
$$

The spiraling geodesics described in Section 1.3 correspond to $\xi_{1}=\cos (t / \varepsilon) / 2, \xi_{2}=0$ and $\xi_{3}=1 /(2 \varepsilon)$. In particular, the constant $\xi_{3}$ is a kind of rounding number reflecting the fact that the geodesic spirals at a certain speed around the $x_{3}$ axis. Moreover, $\xi_{3}$ is preserved by the flow (somehow, the Heisenberg flow is completely integrable), and this property plays a key role in the proof of Theorem 3 below and justifies that we state it only for the Heisenberg manifold (without boundary).

As said above, geodesics corresponding to a large momentum $\xi_{3}$ are precisely the ones used to contradict observability in Theorem 1 . We expect to be able to establish observability if we consider only solutions of (21) whose $\xi_{3}$ (in a certain sense) is not too large. This is the purpose of our second main result.

Set

$$
V_{\varepsilon}=\left\{(x, \xi) \in T^{*} M_{H}:\left|\xi_{3}\right|>\frac{1}{\varepsilon}\left(g_{x}^{*}(\xi)\right)^{1 / 2}\right\}
$$

Note that since $\xi_{3}$ is constant along null-bicharacteristics, the complementary $V_{\varepsilon}^{c}$ of $V_{\varepsilon}$, is invariant under the geodesic flow, i.e., $\phi_{t}\left(V_{\varepsilon}^{c}\right) \subset V_{\varepsilon}^{c}$ (and in some sense $V_{\varepsilon}$ is also invariant, but one should take care that the geodesic flow is not defined in $S \Sigma$, see Remark (7).

Theorem 3. Let $B \subset M_{H}$ be an open sub-Riemannian ball and suppose that $B$ is sufficiently small, so that $\omega=M_{H} \backslash B$ contains an horizontal strip. Let $a \in S_{p h g}^{0}\left(T^{*} M_{H}\right)$, $a \geqslant 0$, such that, denoting by $j: T^{*} \omega \rightarrow T^{*} M_{H}$ the canonical injection,

$$
j\left(T^{*} \omega\right) \cup V_{\varepsilon} \subset S u p p(a) \subset T^{*} M_{H}
$$

and in particular a does not depend on time. There exists $\kappa>0$ such that for any $\varepsilon>0$ and any $T \geqslant \kappa \varepsilon^{-1}$, there holds

$$
\begin{equation*}
\int_{0}^{T}\left|\left(O p(a) \partial_{t} u, \partial_{t} u\right)_{L^{2}\left(M_{H}, \mu\right)}\right| d t \geqslant C\left\|\left(u(0), \partial_{t} u(0)\right)\right\|_{\mathcal{H} \times L^{2}}^{2} \tag{7}
\end{equation*}
$$

for some $C=C(\varepsilon, T)>0$ and for any solution $u$ of (22) which satisfies $\int_{M_{H}} u(0) d \mu=0$.

### 1.6 Comments on the existing literature

The recent article [BS19] deals with the controllability of the Grushin Schrödinger equation $i \partial_{t} u-\Delta_{G} u=0$, where $u \in L^{2}\left((0, T) \times M_{G}\right), M_{G}=(-1,1)_{x} \times \mathbb{T}_{y}$ and $\Delta_{G}=\partial_{x}^{2}+x^{2} \partial_{y}^{2}$ is the Grushin Laplacian. Given a control set of the form $\omega=(-1,1)_{x} \times \omega_{y}$, where $\omega_{y}$ is an open subset of $\mathbb{T}$, the authors prove the existence of a minimal time of control $\mathcal{L}(\omega)$ related to the maximal height of a horizontal strip contained in $M_{G} \backslash \omega$. The intuition is that there are solutions of the Grushin Schrödinger equation which travel along the degenerate line $x=0$ at a finite speed: in some sense, along this line, the Schrödinger equation behaves like a classical (half)-wave equation. What we want here is to explain in a few words why there is a minimal time of observability for the Schrödinger equation, while the wave equation is never observable in finite time as shown by Theorem 1 .

The plane $\mathbb{R}_{x, y}^{2}$ endowed with the vector fields $\partial_{x}$ and $x \partial_{y}$ also admits geodesics similar to the 1-parameter family $q_{\varepsilon}$, namely, for $\varepsilon>0$,

$$
\begin{aligned}
x(t) & =\varepsilon \sin (t / \varepsilon) \\
y(t) & =\varepsilon(t / 2-\varepsilon \sin (2 t / \varepsilon) / 4)
\end{aligned}
$$

These geodesics, denoted by $\gamma_{\varepsilon}$, also "spiral" around the line $x=0$ more and more quickly as $\varepsilon \rightarrow 0$, and so we might expect to construct solutions of the Grushin Schrödinger equation with energy concentrated along $\gamma_{\varepsilon}$, which would contradict observability when $\varepsilon \rightarrow 0$ as above for the Heisenberg wave equation.

However, we can convince ourselves that it is not possible to construct such solutions: in some sense, the dispersion phenomena of the Schrödinger equation exactly compensate the lengthening of the geodesics $\gamma_{\varepsilon}$ as $\varepsilon \rightarrow 0$ and explain that even these Gaussian beams may be observed in $\omega$ from a certain minimal time $\mathcal{L}(\omega)>0$ which is uniform in $\varepsilon$.

To put this argument into a more formal form, we consider the solutions of the bichar-
acteristic equations for the Grushin Schrödinger equation $i \partial_{t} u-\Delta_{G} u=0$ given by

$$
\begin{aligned}
x(t) & =\varepsilon \sin \left(\xi_{y} t\right) \\
y(t) & =\varepsilon^{2} \xi_{y}\left(\frac{t}{2}-\frac{\sin \left(2 \xi_{y} t\right)}{4 \xi_{y}}\right) \\
\xi_{x}(t) & =\varepsilon \xi_{y} \cos \left(\xi_{y} t\right) \\
\xi_{y}(t) & =\xi_{y}
\end{aligned}
$$

It follows from the hypoellipticity of $\Delta_{G}$ (see [BS19, Section 3] for a proof) that

$$
\left|\xi_{y}\right|^{1 / 2} \lesssim \sqrt{-\Delta_{G}}=\left(\left|\xi_{x}\right|^{2}+x^{2}\left|\xi_{y}\right|^{2}\right)^{1 / 2}=\varepsilon\left|\xi_{y}\right| .
$$

Therefore $\varepsilon^{2}\left|\xi_{y}\right| \gtrsim 1$, and hence $|y(t)| \gtrsim t$, independently from $\varepsilon$ and $\xi_{y}$. This heuristic gives the intuition that a minimal time $\mathcal{L}(\omega)$ is required to detect all solutions of the Grushin Schödinger equation from $\omega$, but that for $T_{0}>\mathcal{L}(\omega)$, no solution is localized enough to stay in $M \backslash \omega$ during the time interval $\left(0, T_{0}\right)$. Roughly speaking, the frequencies of order $\xi_{y}$ travel at speed $\sim \xi_{y}$, which is typical for a dispersion phenomenon. This picture is very different from the one for the wave equation (which we consider in this paper) for which no dispersion occurs.

Let us add some general bibliographical comments. Control of subelliptic PDEs has attracted much attention in the last decade. Most results in the literature deal with subelliptic parabolic equations, either the Grushin heat equation (DK20, Koe17) or the heat equation in the Heisenberg group ( $\overline{\mathrm{BC} 17}$, see also references therein). The paper BS19] is the first to deal with a subelliptic Schrödinger equation and the present work is the first to handle exact controllability of subelliptic wave equations.

Recall that the exact controllability of the elliptic wave equation is known to be almost equivalent to the so-called Geometric Control Condition (GCC) (see [BLR92]): this condition says that any geodesic enters the control set $\omega$ within time $T$. In some sense, our main result is that GCC is not verified in the subelliptic setting, as soon as $M \backslash \omega$ has non-empty interior.

A slightly different problem is the approximate controllability of hypoelliptic PDEs, which has been studied in [L20] for hypoelliptic wave and heat equations. Approximate controllability is weaker than exact controllability, and it amounts to proving "quantitative" unique continuation results for hypoelliptic operators. For the hypoelliptic wave equation, it is proved in LL20 that for $T>2 \sup _{x \in M}(\operatorname{dist}(x, \omega))$ (here, dist is the sub-Riemannian distance), the observation of the solution on $(0, T) \times \omega$ determines the initial data, and therefore the whole solution.

### 1.7 Organization of the paper

In Section 2] we construct exact solutions of the subelliptic wave equation (2) concentrating on any given normal sub-Riemannian geodesic. First, in Section 2.1 we show that, given any normal sub-Riemannian geodesic $t \mapsto x(t)$ of $(M, \mathcal{D}, g)$ (i.e., a projection of a nullbicharacteristic of the associated Hamiltonian system) which does not hit $\partial M$ in the time interval $(0, T)$, it is possible to construct a sequence $\left(v_{k}\right)_{k \in \mathbb{N}}$ of approximate solutions of (2) whose energy concentrates along $t \mapsto x(t)$ during the time interval $(0, T)$ as $k \rightarrow+\infty$. By "approximate", we mean here that $\partial_{t t}^{2} v_{k}-\Delta v_{k}$ is small, but not necessarily exactly equal to 0 . In Section 2.1, we provide a first proof for this construction using the classical propagation of complex Lagrangian spaces. An other proof using a Gaussian beam approach is provided in Appendix B Then, in Section [2.2, using this sequence $\left(v_{k}\right)_{k \in \mathbb{N}}$, we explain how to construct a sequence $\left(u_{k}\right)_{k \in \mathbb{N}}$ of exact solutions of $\left(\partial_{t t}^{2}-\Delta\right) u=0$ in $M$ with the same concentration property along the geodesic $t \mapsto x(t)$.

In Section3, we prove the existence of geodesics which spiral in $M$, spending an arbitrarily large time in $M \backslash \omega$. These geodesics generalize the example described in Section 1.3 for the Heisenberg manifold with boundary. The proof proceeds in two steps, first proving the
result in the so-called "nilpotent case" (Section 3.2), and then extending it to the general case (Section 3.3).

In Section 4.1, we use the results of Section 2 and Section 3 to conclude the proof of Theorem 1 and to prove Theorem [2 In Section 4.2 we deduce Corollary 11 by a duality argument. Finally, in Section 4.3, we prove Theorem 3,

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## 2 Gaussian beams along normal sub-Riemannian geodesics

### 2.1 Construction of sequences of approximate solutions

We consider a solution $(x(t), \xi(t))_{t \in[0, T]}$ of (6]) on $M$. We shall describe the construction of solutions of

$$
\begin{equation*}
\partial_{t t}^{2} u-\Delta u=0 \tag{8}
\end{equation*}
$$

on $[0, T] \times M$ with energy

$$
E(u(t, \cdot)):=\frac{1}{2} \int_{M}\left(\left|\partial_{t} u(t, x)\right|^{2}+\left|\nabla^{\mathrm{sR}} u(t, x)\right|^{2}\right) d \mu(x)
$$

concentrated along $x(t)$ for $t \in[0, T]$. The following proposition, which is inspired by Ral82] and MZ02], shows that it is possible, at least for approximate solutions of (8).
Proposition 8. Fix $T>0$ and let $(x(t), \xi(t))_{t \in[0, T]}$ be a solution of (6) which does not hit the boundary $\partial M$ in the time-interval $(0, T)$. Then there exist $a_{0}, \psi \in C^{2}((0, T) \times M)$ such that, setting, for $k \in \mathbb{N}$,

$$
v_{k}(t, x)=k^{\frac{n}{4}-1} a_{0}(t, x) e^{i k \psi(t, x)}
$$

the following properties hold:

- $v_{k}$ is an approximate solution of (8), meaning that

$$
\begin{equation*}
\left\|\partial_{t t}^{2} v_{k}-\Delta v_{k}\right\|_{L^{1}\left((0, T) ; L^{2}(M)\right)} \leqslant C k^{-\frac{1}{2}} \tag{9}
\end{equation*}
$$

- The energy of $v_{k}$ is bounded below with respect to $k$ and $t \in[0, T]$ :

$$
\begin{equation*}
\exists A>0, \forall t \in[0, T], \quad \liminf _{k \rightarrow+\infty} E\left(v_{k}(t, \cdot)\right) \geqslant A \tag{10}
\end{equation*}
$$

- The energy of $v_{k}$ is small off $x(t)$ : for any $t \in[0, T]$, we fix $V_{t}$ an open subset of $M$ for the initial topology of $M$, containing $x(t)$, so that the mapping $t \mapsto V_{t}$ is continuous ( $V_{t}$ is chosen sufficiently small so that this makes sense in a chart). Then

$$
\begin{equation*}
\sup _{t \in[0, T]} \int_{M \backslash V_{t}}\left(\left|\partial_{t} v_{k}(t, x)\right|^{2}+\left|\nabla^{s R} v_{k}(t, x)\right|^{2}\right) d \mu(x) \underset{k \rightarrow+\infty}{\rightarrow} 0 \tag{11}
\end{equation*}
$$

Remark 9. The construction of approximate solutions such as the ones provided by Proposition 8 is usually done for strictly hyperbolic operators, that is operators with a principal symbol $p_{m}$ of order $m$ such that the polynomial $f(s)=p_{m}(t, q, s, \xi)$ has $m$ distinct real roots when $\xi \neq 0$ (see for example [Ral82]). The operator $\partial_{t t}^{2}-\Delta$ is not strictly hyperbolic because $g^{*}$ is degenerate, but our proof shows that the same construction may be adapted without difficulty to this operator along normal bicharacteristics. It was already noted by [Ral82] that the construction of Gaussian beams could be done for more general operators than strictly hyperbolic ones, and that the differences between the strictly hyperbolic case and more general cases arise while dealing with propagation of singularities.

Hereafter we provide two proofs of Proposition 8. The first proof is short and is actually quite straightforward for readers acquainted with the theory of propagation of complex Lagrangian spaces, once one has noticed that the solutions of (6) which we consider live in the elliptic part of the principal symbol of $\partial_{t t}^{2}-\Delta$. For the sake of completeness, and because this also has its own interest, we provide in Appendix B a second proof, more elementary, using the concept of Gaussian beams in the subelliptic context.

First proof of Proposition 8. The usual construction of Gaussian beams, or more generally of a WKB approximation, is related to the transport of complex Lagrangian spaces along bicharacteristics, as reported for example in Hör07, Chapter 24.2] and [vr19, Chapter 1.2].

The usual way to solve (at least approximately) evolution equations of the form

$$
\begin{equation*}
P u=0 \tag{12}
\end{equation*}
$$

where $P$ is a hyperbolic second order differential operator with real principal symbol $p_{2}$ and $C^{\infty}$ coefficients is to search for oscillatory solutions $v_{k}(x)=k^{\frac{n}{4}-1} a_{0}(x) e^{i k \psi(x)}$. In this expression as in the whole proof, we suppress the time variable $t$. Thus we use $x=$ $\left(x_{0}, x_{1}, \ldots, x_{n}\right)$ where $x_{0}=t$ in our earlier notations. We also set $x^{\prime}=\left(x_{1}, \ldots, x_{n}\right)$ and without lost of generality, we only consider bicharacteristics starting from 0 at time 0 .

Plugging this Ansatz into (12), the leading term in $P v_{k}$ is

$$
\begin{equation*}
p_{2}(x, \nabla \psi(x))=0 \tag{13}
\end{equation*}
$$

which we solve for initial conditions $\psi\left(0, x^{\prime}\right)=\psi_{0}\left(x^{\prime}\right), \nabla \psi_{0}(0)=\xi(0)$ and $\psi_{0}(0)=0$. We use the decomposition of $P$ into two half-waves (see Hör07, Lemma 23.2.8])

$$
P(x, D)=\left(D_{0}-\Gamma_{+}\left(x, D^{\prime}\right)\right)\left(D_{0}-\Gamma_{-}\left(x, D^{\prime}\right)\right)+\omega\left(x, D^{\prime}\right)
$$

Here, $\omega \in S^{-\infty}$, the principal symbol $p_{2}$ is written in coordinates as $p_{2}\left(x_{0}, x^{\prime}, \xi_{0}, \xi^{\prime}\right)=$ $\xi_{0}^{2}-r\left(x, \xi^{\prime}\right)$, and $\Gamma_{ \pm}$has principal symbol $\gamma_{ \pm}= \pm r^{1 / 2}$ in a conic neighborhood of $(0, \xi(0))$. This decomposition is justified because $p_{2}$ is elliptic in the direction $\xi(0)$.

Considering separately each half-wave, for example the one with the sign + , we are led to solve $\partial_{t} \psi-\gamma_{+}(x, \nabla \psi)=0$, which is a Hamilton-Jacobi (or eikonal) equation. Its solution $\psi_{t}$ is given by the transport of the values of $\psi_{0}$ by $\Phi_{t}$, where $\left(\Phi_{t}\right)_{t \geqslant 0}$ is a family of symplectomorphisms on the phase space $T^{*} M$, and we have $\Phi_{t}=e^{t \vec{H}_{\gamma_{+}}}$. The trouble is that the solution to this eikonal equation is only local in time, and blow-up (through caustics for example) may happen when $x^{\prime} \mapsto \pi\left(\Phi_{t}\left(x^{\prime}, \nabla \psi_{0}\left(x^{\prime}\right)\right)\right)$ ceases to be a diffeomorphism (conjugate point), where $\pi: T^{*} M \rightarrow M$ is the canonical projection. In the language of Lagrangian spaces, $\Lambda_{0}=\left\{\left(x^{\prime}, \nabla \psi_{0}\left(x^{\prime}\right)\right)\right\} \subset T^{*} M$ is a Lagrangian subspace and, since $\Phi_{t}$ is a symplectomorphism, $\Lambda_{t}=\Phi_{t}\left(\Lambda_{0}\right)$ is Lagrangian as well. If $\pi_{\mid \Lambda_{t}}$ is a local diffeomorphism, one can locally describe $\Lambda_{t}$ by $\Lambda_{t}=\left\{\left(x^{\prime}, \nabla \psi_{t}\left(x^{\prime}\right)\right)\right\} \subset T^{*} M$ for some function $\psi_{t}$, but blowup happens when $\operatorname{rank}\left(d \pi_{\mid \Lambda_{t}}\right)<n$ (classical conjugate point theory), and such a $\psi_{t}$ may not exist.

However, if the phase $\psi$ is complex and satisfies the condition $\operatorname{Im}\left(D^{2} \psi\right)>0$, where $D^{2} \psi$ denotes the Hessian, no blow-up happens as we now explain. Indeed, $\Lambda_{0}=\left\{\left(x^{\prime}, \nabla \psi_{0}\left(x^{\prime}\right)\right)\right\}$ then lives in the complexification of the tangent space $T^{*} M$, which may be thought of as $\mathbb{C}^{2 n}$. Because of the condition $\operatorname{Im}\left(D^{2} \psi_{0}\right)>0, \Lambda_{0}$ is called a "strictly positive Lagrangian space" (see [Hör07, Definition 21.5.5]), meaning that $i \sigma_{\mathbb{C}}(v, v)>0$ for $v$ in the tangent space to $\Lambda_{0}$ ( $\sigma_{\mathbb{C}}$ is defined just below). As in the real case, $\Lambda_{0}$ is transported by a "real dynamics", which we wrote in (36) with coordinates $(y, \eta)$ on $\mathbb{R}^{2 n}$ or $\mathbb{C}^{2 n}$. The fact that the symplectic forms $\sigma$ and $\sigma_{\mathbb{C}}$, defined by $\sigma=\sum d y_{j} \wedge d \eta_{j}$ and $\sigma_{\mathbb{C}}=\sum d y_{j} \wedge \overline{d \eta_{j}}$, are invariant under the dynamics (36) implies that $\sigma=0$ on the tangent space to $\Lambda_{t}$, and that $i \sigma_{\mathbb{C}}(v, v)>0$ for $v$ tangent to $\Lambda_{t}$. It precisely means that $\Lambda_{t}$ is also a strictly positive Lagrangian space. Then, by HS13, Proposition 1.2.5] (or Hör07, Chapter 21]), we know that there exists $\psi_{t}$ with $\operatorname{Im}\left(D^{2} \psi_{t}\right)>0$ such that $\Lambda_{t}=\left\{\left(x^{\prime}, \nabla \psi_{t}\left(x^{\prime}\right)\right)\right\}$. This choice ensures that the terms in the expansion of $P v_{k}$ with a power $k^{\frac{n}{4}+1}$ vanish along the bicharacteristic. To cancel the terms with a power $k^{\frac{n}{4}}$,
it is sufficient to solve linear transport ODEs for $a_{0}$ along the bicharacteristic (see Hör07, Equation (24.2.9)]).

Since we solved (12) at a sufficiently large order along the bicharacteristic starting from $(0, \xi(0))$, the bound (9) then follows from Lemma 20 in Appendix B The bounds (10) and (11) follow from the facts that $v_{k}(x)=k^{\frac{n}{4}-1} a_{0}(x) e^{i k \psi(x)}$ and $\operatorname{Im}\left(D^{2} \psi_{t}\right)>0$.

Remark 10. An interesting question would be to understand the delocalization properties of the Gaussian beams constructed along normal sub-Riemannian geodesics in Proposition 8. Compared with the usual Riemannian case done for example in [Ral82], there is a new phenomenon in the sub-Riemannian case since the geodesic $x(t)$ (or, more precisely, the associated momentum $\xi$ ) may approach the characteristic manifold $\Sigma=\left\{g^{*}=0\right\}$ which is the set of directions in which $\Delta$ is not elliptic. In finite time $T$ as in our case, the geodesic remains far from $\Sigma$, but it may happen as $T \rightarrow+\infty$ that it goes closer and closer to $\Sigma$. The question is then to understand the link between the delocalization properties of the Gaussian beams constructed along such a geodesic, and notably the interplay between the time $T$ and the semi-classical parameter $1 / k$.

### 2.2 Construction of sequences of exact solutions in $M$

In this section, using the approximate solutions of Proposition 2.1. we construct exact solutions of (8) whose energy concentrates along a given normal geodesic of $M$ which does not meet the boundary $\partial M$ during the time interval $[0, T]$.
Proposition 11. Let $(x(t), \xi(t))_{t \in[0, T]}$ be a solution of (6) in $M$ which does not meet $\partial M$ and $\theta \in C_{c}^{\infty}([0, T] \times M)$ with $\theta(t, \cdot) \equiv 1$ in a neighborhood of $x(t)$ and such that the support of $\theta(t, \cdot)$ stays at positive distance of $\partial M$.

Suppose $\left(v_{k}\right)_{k \in \mathbb{N}}$ is constructed along $x(t)$ as in Proposition 8 and $u_{k}$ is the solution of the Cauchy problem

$$
\left\{\begin{array}{l}
\left(\partial_{t t}^{2}-\Delta\right) u_{k}=0 \text { in }(0, T) \times M \\
u_{k}=0 \text { in }(0, T) \times \partial M \\
u_{k \mid t=0}=\left(\theta v_{k}\right)_{\mid t=0}, \partial_{t} u_{k \mid t=0}=\left[\partial_{t}\left(\theta v_{k}\right)\right]_{\mid t=0}
\end{array}\right.
$$

Then:

- The energy of $u_{k}$ is bounded below with respect to $k$ and $t \in[0, T]$ :

$$
\begin{equation*}
\exists A>0, \forall t \in[0, T], \quad \liminf _{k \rightarrow+\infty} E\left(u_{k}(t, \cdot)\right) \geqslant A . \tag{14}
\end{equation*}
$$

- The energy of $u_{k}$ is small off $x(t)$ : for any $t \in[0, T]$, we fix $V_{t}$ an open subset of $M$ for the initial topology of $M$, containing $x(t)$, so that the mapping $t \mapsto V_{t}$ is continuous ( $V_{t}$ is chosen sufficiently small so that this makes sense in a chart). Then

$$
\begin{equation*}
\sup _{t \in[0, T]} \int_{M \backslash V_{t}}\left(\left|\partial_{t} u_{k}(t, x)\right|^{2}+\left|\nabla^{s R} u_{k}(t, x)\right|^{2}\right) d \mu(x) \underset{k \rightarrow+\infty}{\rightarrow} 0 . \tag{15}
\end{equation*}
$$

Proof of Proposition 11. Set $h_{k}=\left(\partial_{t t}^{2}-\Delta\right)\left(\theta v_{k}\right)$. We consider $w_{k}$ the solution of the Cauchy problem

$$
\left\{\begin{array}{l}
\left(\partial_{t t}^{2}-\Delta\right) w_{k}=h_{k} \text { in }(0, T) \times M,  \tag{16}\\
w_{k}=0 \text { in }(0, T) \times \partial M, \\
\left(w_{k \mid t=0}, \partial_{t} w_{k \mid t=0}\right)=(0,0) .
\end{array}\right.
$$

Differentiating $E\left(w_{k}(t, \cdot)\right)$ and using Gronwall's lemma, we get the energy inequality

$$
\sup _{t \in[0, T]} E\left(w_{k}(t, \cdot)\right) \leqslant C\left(E\left(w_{k}(0, \cdot)\right)+\left\|h_{k}\right\|_{L^{1}\left(0, T ; L^{2}(M)\right)}\right) .
$$

Therefore, using (9), we get $\sup _{t \in[0, T]} E\left(w_{k}(t, \cdot)\right) \leqslant C k^{-1}$. Since $u_{k}=\theta v_{k}-w_{k}$, we obtain that

$$
\lim _{k \rightarrow+\infty} E\left(u_{k}(t, \cdot)\right)=\lim _{k \rightarrow+\infty} E\left(\left(\theta v_{k}\right)(t, \cdot)\right)=\lim _{k \rightarrow+\infty} E\left(v_{k}(t, \cdot)\right)
$$

for every $t \in[0, T]$ where the last equality comes from the fact that $\theta$ and its derivatives are bounded and $\left\|v_{k}\right\|_{L^{2}} \leqslant C k^{-1}$ when $k \rightarrow+\infty$. Using (10), we conclude that (14) holds.

To prove (15), we observe similarly that

$$
\begin{aligned}
& \sup _{t \in[0, T]} \int_{M \backslash V_{t}}\left(\left|\partial_{t} u_{k}(t, x)\right|^{2}+\left|\nabla^{\mathrm{sR}} u_{k}(t, x)\right|^{2}\right) d \mu(x) \\
\leqslant & C \sup _{t \in[0, T]}\left(\int_{M \backslash V_{t}}\left(\left|\partial_{t} v_{k}(t, x)\right|^{2}+\left|\nabla^{\mathrm{sR}} v_{k}(t, x)\right|^{2}\right) d \mu(x)\right)+C k^{-\frac{1}{2}} \\
\rightarrow & 0
\end{aligned}
$$

as $k \rightarrow+\infty$, according to (11). It concludes the proof of Proposition 11 .

## 3 Existence of spiraling geodesics

The goal of this section is to prove the following proposition, which is the second building block of the proof of Theorem 1 after the construction of localized solutions of the subelliptic wave equation (2) done in Section (2,

Proposition 12. For any $T_{0}>0$, any $x \in M$ and any open neighborhood $V$ of $x$ in $M$ (with the initial topology on $M$ ), there exists a geodesic $t \mapsto x(t)$ of $(M, \mathcal{D}, g)$ such that $x(t) \in V$ for any $t \in\left[0, T_{0}\right]$.

In Section 3.1, we define the so-called nilpotent approximation $\left(\widehat{M}^{q}, \widehat{\mathcal{D}}^{q}, \widehat{g}^{q}\right)$ of $(M, \mathcal{D}, g)$ at a point $q \in M$, a sub-Riemannian structure which is a first-order approximation of $(M, \mathcal{D}, g)$ at point $q \in M$ whose associated Lie algebra $\operatorname{Lie}\left(\widehat{\mathcal{D}}{ }^{q}\right)$ is nilpotent. The proof of Proposition 12 then splits into two steps, first proving the result for the nilpotent approximation (in Section (3.2) and then showing that the geodesics of $(M, \mathcal{D}, g)$ are well approached by the geodesics of $\left(\widehat{M}^{q}, \widehat{\mathcal{D}}^{q}, \widehat{g}^{q}\right)$ (in Section (3.3) which is sufficient to conclude.

### 3.1 Nilpotent approximation

Given a sub-Riemannian structure $(M, \mathcal{D}, g)$ and a point $q \in M$, its tangent space at $q$ in the Gromov-Hausdorff sense is the sub-Riemannian structure $\left(\widehat{M}^{q}, \widehat{\mathcal{D}}^{q}, \widehat{g}^{q}\right)$, also called nilpotent approximation. It is defined intrinsically (meaning that it does not depend on a choice of coordinates or of local frame) as an equivalence class under the action of sub-Riemannian isometries (see [Bel96], Jea14).

Sub-Riemannian flag. We define the sub-Riemannian flag of $(M, \mathcal{D}, g)$ as follows: we set $\mathcal{D}^{0}=\{0\}, \mathcal{D}^{1}=\mathcal{D}$, and, for any $j \geqslant 1, \mathcal{D}^{j+1}=\mathcal{D}^{j}+\left[\mathcal{D}, \mathcal{D}^{j}\right]$. For any point $q \in M$, it defines a flag

$$
\{0\}=\mathcal{D}_{q}^{0} \subset \mathcal{D}_{q}^{1} \subset \ldots \subset \mathcal{D}_{q}^{r-1} \nsubseteq \mathcal{D}_{q}^{r(q)}=T_{q} M
$$

The integer $r(q)$ is called the non-holonomic order of $\mathcal{D}$ at $q$, and it is equal to 2 in the Heisenberg manifold for example. For $0 \leqslant i \leqslant r(q)$, we set $n_{i}(q)=\operatorname{dim} \mathcal{D}_{q}^{i}$, and the sequence $\left(n_{i}(q)\right)_{0 \leqslant i \leqslant r(q)}$ is called the growth vector at point $q$. We set $\mathcal{Q}(q)=\sum_{i=1}^{r(q)} i\left(n_{i}(q)-n_{i-1}(q)\right)$. Finally, we define the non-decreasing sequence of weights $w_{i}(q)$ for $1 \leqslant i \leqslant n$ as follows. Given any $1 \leqslant i \leqslant n$, there exists a unique $1 \leqslant j \leqslant n$ such that $n_{j-1}(q)+1 \leqslant i \leqslant n_{j}(q)$. We set $w_{i}(q)=j$.

Regular and singular points. We say that $q \in M$ is regular if the growth vector $\left(n_{i}\left(q^{\prime}\right)\right)_{0 \leqslant i \leqslant r\left(q^{\prime}\right)}$ at $q^{\prime}$ is constant for $q^{\prime}$ in a neighborhood of $q$. Otherwise, $q$ is said to be singular. If any point $q \in M$ is regular, we say that the structure is equiregular.

Non-holonomic orders. The non-holonomic order of a smooth germ of function is given by the formula

$$
\operatorname{ord}_{q}(f)=\min \left\{s \in \mathbb{N}: \exists i_{1}, \ldots, i_{s} \in\{1, \ldots, k\} \text { such that }\left(X_{i_{1}} \ldots X_{i_{s}} f\right)(q) \neq 0\right\}
$$

where we adopt the convention that $\min \emptyset=+\infty$. The non-holonomic order of a smooth germ of vector field $X$ at $q$, denoted by $\operatorname{ord}_{q}(X)$, is the real number defined by

$$
\operatorname{ord}_{q}(X)=\sup \left\{\sigma \in \mathbb{R}: \operatorname{ord}_{q}(X f) \geqslant \sigma+\operatorname{ord}_{q}(f), \quad \forall f \in C^{\infty}(q)\right\}
$$

For example, every $X$ which has the property that $X\left(q^{\prime}\right) \in \mathcal{D}_{q^{\prime}}^{i}$ for any $q^{\prime}$ in a neighborhood of $q$ is of non-holonomic order $\geqslant-i$.

Privileged coordinates. Locally around $q \in M$, it is possible to define a set of so-called "privileged coordinates" of $M$ (see Bel96).

A family $\left(Z_{1}, \ldots, Z_{n}\right)$ of $n$ vector fields is said to be adapted to the sub-Riemannian flag of $(M, \mathcal{D}, g)$ at $q$ if it is a frame of $T_{q} M$ at $q$ and if $Z_{i}(q) \in \mathcal{D}_{q}^{w_{i}(q)}$ for any $i \in\{1, \ldots, n\}$.

A system of privileged coordinates ar $q$ is a system of local coordinates $\left(x_{1}, \ldots, x_{n}\right)$ such that $\operatorname{ord}_{q}\left(x_{i}\right)=w_{i}$ for $1 \leqslant i \leqslant n$. In particular, for privileged coordinates, we have $\partial_{x_{i}} \in \mathcal{D}_{q}^{w_{i}(q)} \backslash \mathcal{D}_{q}^{w_{i}(q)-1}$ at $q$, meaning that privileged coordinates are adapted to the flag.

Exponential coordinates of the second kind. Choose an adapted frame $\left(Z_{1}, \ldots, Z_{n}\right)$ at $q$. It is proved in [Jea14, Appendix B] that the inverse of the local diffeomorphism

$$
\left(x_{1}, \ldots, x_{n}\right) \mapsto \exp \left(x_{1} Z_{1}\right) \circ \cdots \circ \exp \left(x_{n} Z_{n}\right)(q)
$$

defines privileged coordinates at $q$, called exponential coordinates of the second kind.
Dilations. We consider a chart of privileged at $q$ given by a smooth mapping $\psi_{q}: U \rightarrow$ $\mathbb{R}^{n}$, where $U$ is a neighborhood of $q$ in $M$, with $\psi_{q}(q)=0$. For every $\varepsilon \in \mathbb{R} \backslash\{0\}$, we consider the dilation $\delta_{\varepsilon}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ defined by

$$
\delta_{\varepsilon}(x)=\left(\varepsilon^{w_{i}(q)} x_{1}, \ldots, \varepsilon^{w_{n}(q)} x_{n}\right)
$$

for every $x=\left(x_{1}, \ldots, x_{n}\right)$. A dilation $\delta_{\varepsilon}$ acts also on functions and vector fields on $\mathbb{R}^{n}$ by pull-back: $\delta_{\varepsilon}^{*} f=f \circ \delta_{\varepsilon}$ and $\delta_{\varepsilon}^{*} X$ is the vector field such that $\left(\delta_{\varepsilon}^{*} X\right)\left(\delta_{\varepsilon}^{*} f\right)=\delta_{\varepsilon}^{*}(X f)$.

Nilpotent approximation. Fix a system of privileged coordinates $\left(x_{1}, \ldots, x_{n}\right)$ at $q$. Given a sequence of integers $\alpha=\left(\alpha_{1}, \ldots, \alpha_{n}\right)$, we define the weighted degree of $x^{\alpha}=$ $x_{1}^{\alpha_{1}} \ldots x_{n}^{\alpha_{n}}$ to be $w(\alpha)=w_{1}(q) \alpha_{1}+\ldots+w_{n}(q) \alpha_{n}$. Every vector field $X_{i}$ is of non-holonomic order $\geqslant-1$, hence it has a Taylor expansion

$$
X_{i}(x) \sim \sum_{\alpha, j} a_{\alpha, j} x^{\alpha} \partial_{x_{j}}
$$

where $w(\alpha) \geqslant w_{j}(q)-1$ if $a_{\alpha, j} \neq 0$. Therefore, we may write $X_{i}$ as a formal series

$$
X_{i}=X_{i}^{(-1)}+X_{i}^{(0)}+X_{i}^{(1)}+\ldots
$$

where $X_{i}^{(s)}$ is a homogeneous vector field of degree $s$. We set $\widehat{X}_{i}^{q}=X_{i}^{(-1)}$ for $1 \leqslant i \leqslant m$. Each $\widehat{X}_{i}^{q}$ may be seen as a vector field on $\mathbb{R}^{n}$ thanks to the coordinates $\left(x_{1}, \ldots, x_{n}\right)$. Moreover,

$$
\widehat{X}_{i}^{q}=\lim _{\varepsilon \rightarrow 0} \varepsilon \delta_{\varepsilon}^{*}\left(\psi_{q}\right)_{*} X_{i}
$$

in $C^{\infty}$ topology: all derivatives uniformly converge on compact subsets. For $\varepsilon>0$ small enough we have

$$
X_{i}^{\varepsilon}:=\varepsilon \delta_{\varepsilon}^{*}\left(\psi_{q}\right)_{*} X_{i}=\widehat{X}_{i}^{q}+\varepsilon R_{i}^{\varepsilon}
$$

where $R_{i}^{\varepsilon}$ depends smoothly on $\varepsilon$ for the $C^{\infty}$ topology. Note also that $\widehat{X}_{i}^{q}$ is homogeneous of degree -1 with respect to dilations, i.e., $\lambda \delta_{\lambda}^{*} \widehat{X}_{i}^{q}=\widehat{X}_{i}^{q}$ for any $\lambda \neq 0$. An important property is that $\left(\widehat{X}_{1}^{q}, \ldots, \widehat{X}_{m}^{q}\right)$ generates a nilpotent Lie algebra of step $r=w_{n}$ (see Jea14, Proposition 2.3]).

The nilpotent approximation $\left(\widehat{M}^{q}, \widehat{\mathcal{D}}^{q}, \widehat{g}^{q}\right)$ of $(M, \mathcal{D}, g)$ at $q$ is then defined as $\widehat{M} \simeq$ $\mathbb{R}^{n}$ endowed with the sub-Riemannian distribution $\widehat{\mathcal{D}}^{q}=\operatorname{Span}\left(\widehat{X}_{1}^{q}, \ldots, \widehat{X}_{k}^{q}\right)$ and the subRiemannian metric $\widehat{g}^{q}$ on $\widehat{\mathcal{D}}^{q}$ verifying $\widehat{g}^{q}\left(\widehat{X}_{i}^{q}, \widehat{X}_{j}^{q}\right)=g_{q}\left(X_{i}, X_{j}\right)$. Note that the nilpotent sub-Riemannian structure $\left(\widehat{M}^{q}, \widehat{\mathcal{D}}^{q}, \widehat{g}^{q}\right)$ is homogeneous with respect to the dilations $\delta_{\varepsilon}$.

### 3.2 The nilpotent case

We first prove Proposition 12 when $(M, \mathcal{D}, g)$ is nilpotent, i.e., Lie $(\mathcal{D})$ is a nilpotent Lie algebra. The nilpotent Hamiltonian is denoted by

$$
\begin{equation*}
\widehat{H}=\sum_{i=1}^{m} h_{\widehat{X}_{i}^{q}}^{2} \tag{17}
\end{equation*}
$$

As explained in Section 1.3 the proof consists in identifying a "Heisenberg sub-structure" in $(M, \mathcal{D}, g)$, for which we know that there exist spiraling geodesics.
Lemma 13. Let us assume that $(M, \mathcal{D}, g)$ is nilpotent. Then, for any $T>0$, any $x \in M$ and any open neighborhood $V$ of $x$ in $M$, there exists $\lambda_{0} \in T^{*} M$ such that $\pi\left(e^{t \vec{H}} \lambda_{0}\right) \in V$ for any $t \in\left[0, T_{0}\right]$.

Proof of Lemma 13. For this proof, since all considered sub-Riemannian structures are nilpotent, we simplify notations by replacing the vector fields $\widehat{X}_{i}$ by the notation $X_{i}$.

The main idea of the proof is to use a desingularization $(\tilde{M}, \tilde{\mathcal{D}}, \tilde{g})$ of $(M, \mathcal{D}, g)$ around $x$ whose Lie algebra is free up to some step $r$ (see Jea14, Definition 2.13]) and nilpotent. In this desingularized sub-Riemannian structure, we isolate a "Heisenberg sub-structure", and projecting its spiraling geodesics onto $(M, \mathcal{D}, g)$, we get the result. By Heisenberg substructure of $(\tilde{M}, \tilde{\mathcal{D}}, \tilde{g})$, we mean two vector fields $\tilde{X}_{1}, \tilde{X}_{2}$ which may be completed into a $\tilde{g}$-orthonormal family, such that $\left[\tilde{X}_{1}, \tilde{X}_{2}\right] \notin \operatorname{Span}\left(\tilde{X}_{1}, \tilde{X}_{2}\right)$ and the geodesics in $\tilde{M}$ of the Hamiltonian $h_{\tilde{X}_{1}}^{2}+h_{\tilde{X}_{2}}^{2}$ are also geodesics of $(\tilde{M}, \tilde{\mathcal{D}}, \tilde{g})$.

Let us observe that, given a $g$-orthonormal frame $\left(X_{1}, \ldots, X_{n}\right)$ of $(M, \mathcal{D}, g)$, we cannot isolate a Heisenberg sub-structure by simply taking the quotient of $\operatorname{Lie}\left(X_{1}, \ldots, X_{m}\right)$ by its ideal $I\left(X_{3}, \ldots, X_{m}\right)$, because $I\left(X_{3}, \ldots, X_{m}\right)$ may be very large (even equal to $\left.\operatorname{Lie}\left(X_{1}, \ldots, X_{m}\right)\right)$ if there are relations such as $\left[X_{1}, X_{3}\right]=X_{1}$. This is why we introduce the desingularization $(\tilde{M}, \tilde{\mathcal{D}}, \tilde{g})$ : its Lie algebra is free up to a step $r$, and therefore such relations cannot hold.

Let us first note that there exist $q_{0} \in V$ and $X_{1}, X_{2} \in \mathcal{D}$ such that $\left[X_{1}, X_{2}\right]\left(q_{0}\right) \notin \mathcal{D}_{q_{0}}$. This is because otherwise $[\mathcal{D}, \mathcal{D}] \subset \mathcal{D}$ in $V$, and therefore the rank condition (11) cannot hold in $V$. We denote by $r$ the non-holonomic order of $\mathcal{D}$ at $q_{0}$.

We now introduce some notations related to free Lie algebras. Let $\mathcal{L}=\mathcal{L}(1, \ldots, m)$ be the free Lie algebra generated by $\{1, \ldots, m\}$. We denote by $\mathcal{L}^{s}$ the subspace generated by elements of $\mathcal{L}$ of length $\leqslant s$, and by $\tilde{n}_{s}$ the dimension of $\mathcal{L}^{s}$.
Lemma 14 ([Bel96], Jea14 (Lemma 2.5, Theorem 2.9 and Remark 2.9)). Let $\tilde{M}=M \times$ $\mathbb{R}^{\tilde{n}_{r}-n}$. Then there exist a neighborhood $\tilde{U} \subset \tilde{M}$ of $\left(q_{0}, 0\right)$, a neighborhood $U \subset M$ of $q_{0}$ with $U \times\{0\} \subset \tilde{U}$, local coordinates $(x, y)$ on $\tilde{U}$, and smooth vector fields $\tilde{X}_{1}, \ldots, \tilde{X}_{m}$ on $\tilde{U}$, such that

- Every $\tilde{q}$ in $\tilde{U}$ is regular;
- For any $1 \leqslant i \leqslant m$, we have $d \pi\left(\tilde{X}_{i}\right)=X_{i}$, where $\pi: \tilde{M} \rightarrow M$ is the canonical projection;
- Lie $\left(\tilde{X}_{1}, \ldots, \tilde{X}_{m}\right)$ is a free nilpotent Lie algebra of step r;
- The image by the projection $\pi$ of any geodesic (traveled at speed 1) of the non-holonomic system in $\tilde{U}$ defined by $\tilde{X}_{1}, \ldots, \tilde{X}_{m}$ is a geodesic of $(M, \mathcal{D}, g)$ (also traveled at speed 1 ).
We set $\tilde{\mathcal{D}}=\operatorname{Span}\left(\tilde{X}_{1}, \ldots, \tilde{X}_{m}\right)$ and we denote by $\tilde{g}$ the Riemannian metric on $\tilde{\mathcal{D}}$ such that $\left(\tilde{X}_{1}, \ldots, \tilde{X}_{m}\right)$ is $\tilde{g}$-orthonormal. We denote by $\mathcal{I}$ the ideal of $\operatorname{Lie}\left(\tilde{X}_{1}, \ldots, \tilde{X}_{m}\right)$ generated by $\tilde{X}_{3}, \ldots, \tilde{X}_{m}$ and all brackets of $\tilde{X}_{1}, \ldots, \tilde{X}_{m}$ of length $\geqslant 3$. Then $\operatorname{Lie}\left(\tilde{X}_{1}, \ldots, \tilde{X}_{m}\right) / \mathcal{I}$ has dimension 3 .

We denote by $\tilde{X}_{1}^{H}$ and $\tilde{X}_{2}^{H}$ the images of $\tilde{X}_{1}$ and $\tilde{X}_{2}$ through this quotient. The Lie group $G$ generated by $\exp \left(t \tilde{X}_{1}^{H}\right)$ and $\exp \left(t \tilde{X}_{2}^{H}\right)$ is nilpotent and free up to step 2 , and therefore there are local coordinates $\left(x_{1}, x_{2}, x_{3}\right)$ on $G$ such that $\tilde{X}_{1}^{H}=\partial_{x_{1}}$ and $\tilde{X}_{2}^{H}=\partial_{x_{2}}+x_{1} \partial_{x_{3}}$. Then, the geodesics exhibited in Section 1.3 are spiraling geodesics in $G$. Since $\operatorname{Lie}\left(\tilde{X}_{1}, \ldots, \tilde{X}_{m}\right)$ is free, it gives spiraling geodesics in $(\tilde{M}, \tilde{\mathcal{D}}, \tilde{g})$ (traveled at speed 1, i.e., $\tilde{g}^{*}=1$ ). Their
images by the canonical projection $\pi: \tilde{M} \rightarrow M$ are spiraling geodesics in ( $M, \mathcal{D}, g$ ) (with $g^{*}=1$ ).

Remark 15. The geodesics constructed in Lemma 13 lose their optimality quickly, in the sense that their first conjugate point and their cut-point are close to $q_{0}$. In local coordinates, they are closer to $q_{0}$ as the norm of $\lambda_{0}$ is larger.

### 3.3 The general case

In this section, we conclude the proof of Proposition 12, We do not assume anymore that $(M, \mathcal{D}, g)$ is nilpotent. We show that on finite time intervals, the geodesics of $(M, \mathcal{D}, g)$ remain close to those of $\left(\widehat{M^{q}}, \widehat{\mathcal{D}}, \widehat{g}\right)$. We set $Y_{i}=\left(\psi_{q}\right)_{*} X_{i}$ and $X_{i}^{\varepsilon}=\varepsilon \delta_{\varepsilon}^{*} Y_{i}$ which is a vector field on $\mathbb{R}^{n}$. Recall that

$$
X_{i}^{\varepsilon}=\widehat{X}_{i}^{q}+\varepsilon R_{i}^{\varepsilon}
$$

where $R_{i}^{\varepsilon}$ depends smoothly on $\varepsilon$ for the $C^{\infty}$ topology. Therefore, using the homogeneity of $\widehat{X}_{i}^{q}$, we get

$$
\begin{equation*}
Y_{i}=\frac{1}{\varepsilon}\left(\delta_{\varepsilon}\right)_{*} X_{i}^{\varepsilon}=\frac{1}{\varepsilon}\left(\delta_{\varepsilon}\right)_{*}\left(\widehat{X}_{i}^{q}+\varepsilon R_{i}^{\varepsilon}\right)=\widehat{X}_{i}^{q}+\left(\delta_{\varepsilon}\right)_{*} R_{i}^{\varepsilon} . \tag{18}
\end{equation*}
$$

Recall that the Hamiltonian $\widehat{H}$ associated to the vector fields $\widehat{X}_{i}^{q}$ is given by (17). Similarly, we set

$$
H=\sum_{i=1}^{m} h_{Y_{i}}^{2}
$$

where we used the notation $h_{Z}$ for vector fields $Z$ which was introduced in Section 1.4 We note that (18) gives

$$
h_{Y_{i}}=h_{\widehat{X}_{i}^{q}}+h_{\left(\delta_{\varepsilon}\right)_{*} R_{i}^{\varepsilon}} .
$$

Hence

$$
\vec{H}=2 \sum_{i=1}^{m} h_{Y_{i}} \vec{h}_{Y_{i}}=\overrightarrow{\hat{H}}+\vec{\Theta}
$$

where $\vec{\Theta}$ is a smooth vector field on $T^{*} \mathbb{R}^{n}$ such that $\|\vec{\Theta}(x, \xi)\| \rightarrow 0$ when $\|x\| \rightarrow 0$ (uniformly for $\|\xi\| \leqslant C$ ). This last point comes from the fact that $\vec{\Theta}$ has higher non-holonomic order than $\overrightarrow{\hat{H}}$. In other words, it follows from the smooth dependence of $R_{i}^{\varepsilon}$ on $\varepsilon$ for the $C^{\infty}$ topology (uniform convergence of all derivatives on compact subsets of $\mathbb{R}^{n}$ ). Together with Lemma (13) and using that $X_{i}=\psi_{q}^{*} Y_{i}$, it concludes the proof of Proposition 12

## 4 Proofs

### 4.1 Proofs of Theorem 1 and Theorem 2

In this section, we conclude the proof of Theorem 1 and we prove Theorem 2
Proof of Theorem 1. Fix a point $x \in M$ and an open neighborhood $V$ of $x$ in $M$ such that $V \subset M \backslash \omega$. Such $x$ and $V$ exist since $M \backslash \omega$ has non-empty interior. Fix $V^{\prime}$ an open neighborhood of $x$ in $M$ such that $\overline{V^{\prime}} \subset V$, and fix also $T_{0}>0$.

As already explained in Section 1.3 , to conclude the proof of Theorem 11, we use Proposition 11 applied to the particular geodesics constructed in Proposition 12

By Proposition 12, we know that there exists a normal geodesic $t \mapsto x(t)$ such that $x(t) \in V^{\prime}$ for any $t \in\left(0, T_{0}\right)$. We denote by $\left(u_{k}\right)_{k \in \mathbb{N}}$ a sequence of solutions of (8) as in Proposition 11 whose energy at time $t$ concentrates on $x(t)$ for $t \in\left(0, T_{0}\right)$. Because of (14), we know that

$$
\left\|\left(u_{k}(0), \partial_{t} u_{k}(0)\right)\right\|_{\mathcal{H} \times L^{2}} \geqslant c>0
$$

uniformly in $k$.

Therefore, in order to establish Theorem 1 it is sufficient to show that

$$
\begin{equation*}
\int_{0}^{T_{0}} \int_{\omega}\left|\partial_{t} u_{k}(t, x)\right|^{2} d \mu(x) d t \underset{k \rightarrow+\infty}{\rightarrow} 0 \tag{19}
\end{equation*}
$$

Since $x(t) \in V^{\prime}$ for any $t \in\left(0, T_{0}\right)$, we get that for $V_{t}$ chosen sufficiently small for any $t \in\left(0, T_{0}\right)$, the inclusion $V_{t} \subset V$ holds. Combining this last remark with (15), we get (19), which concludes the proof of Theorem 1

Proof of Theorem 2. We first observe that the growth vector (defined in Section 3.1) of the nilpotent approximation $\left(\widehat{M}^{x}, \widehat{\mathcal{D}}^{x}, \widehat{g}^{x}\right)$ at 0 is equal to the growth vector of $(M, \mathcal{D}, g)$ at $x$ (see Jea14, Lemma 2.1]). Hence, ( $\left.\widehat{M}^{x}, \widehat{\mathcal{D}}^{x}, \widehat{g}^{x}\right)$ is "strictly sub-Riemannian" at 0 , in the sense that $\widehat{\mathcal{D}}_{0}^{x} \neq T_{0} \widehat{M}^{x}$. The result then follows from Lemma 13 and the computations of Section 3.3 .

### 4.2 Proof of Corollary 1

We endow the topological dual $\mathcal{H}(M)^{\prime}$ with the norm $\|v\|_{\mathcal{H}(M)^{\prime}}=\left\|(-\Delta)^{-1 / 2} v\right\|_{L^{2}(M)}$.
The following proposition is standard (see, e.g., TW09, [LRLTT17).
Lemma 16. Let $T_{0}>0$, and $\omega \subset M$ be a measurable set. Then the following two observability properties are equivalent:
(P1): There exists $C_{T_{0}}$ such that for any $\left(v_{0}, v_{1}\right) \in \mathcal{H}(M) \times L^{2}(M)$, the solution $v \in$ $C^{0}\left(0, T_{0} ; \mathcal{H}(M)\right) \cap C^{1}\left(0, T_{0} ; L^{2}(M)\right)$ of (2) satisfies

$$
\begin{equation*}
\int_{0}^{T_{0}} \int_{\omega}\left|\partial_{t} v(t, q)\right|^{2} d \mu(q) d t \geqslant C_{T_{0}}\left\|\left(v_{0}, v_{1}\right)\right\|_{\mathcal{H}(M) \times L^{2}(M)} \tag{20}
\end{equation*}
$$

(P2): There exists $C_{T_{0}}$ such that for any $\left(v_{0}, v_{1}\right) \in L^{2}(M) \times \mathcal{H}(M)^{\prime}$, the solution $v \in$ $C^{0}\left(0, T_{0} ; L^{2}(M)\right) \cap C^{1}\left(0, T_{0} ; \mathcal{H}(M)^{\prime}\right)$ of (2) satisfies

$$
\begin{equation*}
\int_{0}^{T_{0}} \int_{\omega}|v(t, q)|^{2} d \mu(q) d t \geqslant C_{T_{0}}\left\|\left(v_{0}, v_{1}\right)\right\|_{L^{2} \times \mathcal{H}(M)^{\prime}}^{2} \tag{21}
\end{equation*}
$$

Proof. Let us assume that (P2) holds. Let $u$ be a solution of (2) with initial conditions $\left(u_{0}, u_{1}\right) \in \mathcal{H}(M) \times L^{2}(M)$. We set $v=\partial_{t} u$, which is a solution of (21) with initial data $v_{\mid t=0}=u_{1} \in L^{2}(M)$ and $\partial_{t} v_{\mid t=0}=\Delta u_{0} \in \mathcal{H}(M)^{\prime}$. Since $\left\|\left(v_{0}, v_{1}\right)\right\|_{L^{2} \times \mathcal{H}(M)^{\prime}}=$ $\left\|\left(u_{1}, \Delta u_{0}\right)\right\|_{L^{2} \times \mathcal{H}(M)^{\prime}}=\left\|\left(u_{0}, u_{1}\right)\right\|_{\mathcal{H}(M) \times L^{2}}$, applying the observability inequality (21) to $v=\partial_{t} u$, we obtain (20). The proof of the other implication is similar.

Finally, using Theorem 1. Lemma 16 and the standard HUM method (Lio88), we get Corollary 1

### 4.3 Proof of Theorem 3

Let us first define spaces adapted to the fact that $M_{H}$ has no boundary. We set

$$
\mathcal{H}_{0}=\left\{u_{0} \in \mathcal{H}, \int_{M_{H}} u_{0} d \mu=0\right\}
$$

and we denote its topological dual by $\mathcal{H}_{0}^{\prime}$.
Lemma 17. The embeddings $\mathcal{H}_{0} \hookrightarrow L^{2}$ and $L^{2} \hookrightarrow \mathcal{H}_{0}^{\prime}$ are compact.
Proof. By duality, we only need to prove that $\mathcal{H}_{0} \hookrightarrow L^{2}$ is compact. Using a classical subelliptic estimate (see Hör67] and [RS76. Theorem 17]), we know that there exists $C>0$ such that for any $u \in C^{\infty}\left(M_{H}\right)$,

$$
\|u\|_{H^{\frac{1}{2}}\left(M_{H}\right)} \leqslant C\left(\|u\|_{L^{2}\left(M_{H}\right)}+\|u\|_{\mathcal{H}\left(M_{H}\right)}\right)
$$

Moreover, by the Poincaré inequality, we also have for any $u \in C^{\infty}\left(M_{H}\right)$ with $\int_{M_{H}} u d \mu=0$ that $\|u\|_{L^{2}\left(M_{H}\right)} \leqslant C\left\|\partial_{x} u\right\|_{L^{2}\left(M_{H}\right)} \leqslant C\|u\|_{\mathcal{H}\left(M_{H}\right)}$. Therefore, $\mathcal{H}_{0} \hookrightarrow H^{\frac{1}{2}}\left(M_{H}\right)$ with continuous embedding. The result then follows from the fact that $H^{\frac{1}{2}}\left(M_{H}\right)$ is compactly embedded in $L^{2}\left(M_{H}\right)$.

We then prove a weak observability inequality.
Lemma 18. There exists $\kappa>0$ such that for any $\varepsilon>0$ and any $T \geqslant \kappa \varepsilon^{-1}$, there holds

$$
\begin{equation*}
C\left\|\left(u_{0}, u_{1}\right)\right\|_{\mathcal{H} \times L^{2}}^{2} \leqslant \int_{0}^{T}\left|\left(O p(a) \partial_{t} u, \partial_{t} u\right)_{L^{2}\left(M_{H}, \mu\right)}\right| d t+\left\|\left(u_{0}, u_{1}\right)\right\|_{L^{2} \times \mathcal{H}^{\prime}}^{2} \tag{22}
\end{equation*}
$$

for some $C(\varepsilon, T)>0$ and any solution $u$ of (2) such that $\int_{M_{H}} u_{0} d \mu=0$.
Proof of Lemma 18. In this proof, we use the notation $P=\partial_{t t}^{2}-\Delta_{H}$. For the sake of a contradiction, suppose that there exists a sequence $\left(u^{k}\right)_{k \in \mathbb{N}}$ of solutions of the wave equation such that $\left\|\left(u_{0}^{k}, u_{1}^{k}\right)\right\|_{\mathcal{H} \times L^{2}}=1$ for any $k \in \mathbb{N}$ and

$$
\begin{equation*}
\left\|\left(u_{0}^{k}, u_{1}^{k}\right)\right\|_{L^{2} \times \mathcal{H}^{\prime}} \rightarrow 0, \quad \int_{0}^{T}\left|\left(\mathrm{Op}(a) \partial_{t} u^{k}, \partial_{t} u^{k}\right)_{L^{2}\left(M_{H}, \mu\right)}\right| d t \rightarrow 0 \tag{23}
\end{equation*}
$$

as $k \rightarrow \infty$. Following the strategy of Tar90 and Gér91, our goal is to associate a defect measure to the sequence $\left(u^{k}\right)_{k \in \mathbb{N}}$. Since the functional spaces involved in our result are unusual, we give the argument in detail.

We consider the space of functions $u \in C^{\infty}\left([0, T] \times M_{H}\right)$ such that $\int_{M_{H}} u(0, \cdot) d \mu=0$, and we denote by $\mathcal{H}_{T}$ its the completion for the norm $\|\cdot\|_{\mathcal{H}_{T}}$ induced by the scalar product

$$
(u, v)_{\mathcal{H}_{T}}=\int_{0}^{T} \int_{M_{H}}\left(\partial_{t} u \partial_{t} v+\left(\nabla^{\mathrm{sR}} u\right) \cdot\left(\nabla^{\mathrm{sR}} v\right)\right) d \mu(q) d t
$$

By conservation of energy, the quantity $\left\|u^{k}\right\|_{\mathcal{H}_{T}}$ is bounded independently of $k$. Together with the first convergence in (23), it implies that $u^{k} \rightharpoonup 0$ for the weak topology of $\mathcal{H}_{T}$.

Fix $B \in \Psi_{\mathrm{phg}}^{0}\left((0, T) \times M_{H}\right)$. We have

$$
\begin{align*}
\left(B u^{k}, u^{k}\right)_{\mathcal{H}_{T}}= & \int_{0}^{T} \int_{M_{H}}\left(\partial_{t}\left(B u^{k}\right) \partial_{t} u^{k}+\left(\nabla^{\mathrm{sR}}\left(B u^{k}\right)\right) \cdot\left(\nabla^{\mathrm{sR}} u^{k}\right)\right) d \mu(q) d t \\
= & \int_{0}^{T} \int_{M_{H}}\left(\left(\left[\partial_{t}, B\right] u^{k}\right) \partial_{t} u^{k}+\left(\left[\nabla^{\mathrm{sR}}, B\right] u^{k}\right) \cdot\left(\nabla^{\mathrm{sR}} u^{k}\right)\right) d \mu(q) d t \\
& +\int_{0}^{T} \int_{M_{H}}\left(\left(B \partial_{t} u^{k}\right)\left(\partial_{t} u^{k}\right)+\left(B \nabla^{\mathrm{sR}} u^{k}\right) \cdot\left(\nabla^{\mathrm{sR}} u^{k}\right)\right) d \mu(q) d t \tag{24}
\end{align*}
$$

Since $\left[\partial_{t}, B\right] \in \Psi_{\mathrm{phg}}^{0}\left((0, T) \times M_{H}\right),\left[\nabla^{\mathrm{sR}}, B\right] \in \Psi_{\mathrm{phg}}^{0}\left((0, T) \times M_{H}\right)$ and $u^{k} \rightarrow 0$ strongly in $L^{2}\left((0, T) \times M_{H}\right)$, the first one of the two lines in (24) converges to 0 as $k \rightarrow+\infty$. Moreover, the last line is bounded uniformly in $k$ since $B \in \Psi_{\mathrm{phg}}^{0}\left((0, T) \times M_{H}\right)$. Hence $\left(B u^{k}, u^{k}\right)_{\mathcal{H}_{T}}$ is uniformly bounded. By a standard diagonal extraction argument (see Gér91] for example), there exists a subsequence, which we still denote by $\left(u^{k}\right)_{k \in \mathbb{N}}$ such that $\left(B u^{k}, u^{k}\right)$ converges for any $B$ of principal symbol $b$ in a countable dense subset of $C_{c}^{\infty}\left((0, T) \times M_{H}\right)$. Moreover, the limit only depends on the principal symbol $b$, and not on the full symbol.

Let us now prove that

$$
\begin{equation*}
\liminf _{k \rightarrow+\infty}\left(B u^{k}, u^{k}\right)_{\mathcal{H}_{T}} \geqslant 0 \tag{25}
\end{equation*}
$$

when $b \geqslant 0$. With a bracket argument as in (24), we see that it is equivalent to proving

$$
\begin{equation*}
\liminf _{k \rightarrow+\infty}\left(\left(B \partial_{t} u^{k}, \partial_{t} u^{k}\right)_{L^{2}}+\left(B \nabla^{\mathrm{sR}} u^{k}, \nabla^{\mathrm{sR}} u^{k}\right)_{L^{2}}\right) \geqslant 0 \tag{26}
\end{equation*}
$$

But there exists $B^{\prime} \in \Psi_{\mathrm{phg}}^{0}\left((0, T) \times M_{H}\right)$ such that $B^{\prime}-B \in \Psi_{\mathrm{phg}}^{-1}\left((0, T) \times M_{H}\right)$ and $B^{\prime}$ is positive (this is the so-called Friedrichs quantization, see for example Tay74, Chapter VII]). It immediately implies that (26) and (25) hold.

Therefore, setting $p=\sigma_{p}(P)$ and denoting by $\mathcal{C}(p)$ the characteristic manifold $\mathcal{C}(p)=$ $\{p=0\}$, there exists a positive Radon measure $\nu$ on $S^{*}(\mathcal{C}(p))=\mathcal{C}(p) /(0,+\infty)$ such that

$$
\left(\mathrm{Op}(b) u^{k}, u^{k}\right)_{\mathcal{H}_{T}} \rightarrow \int_{S^{*}(\mathcal{C}(p))} b d \nu
$$

for any $b \in S_{\mathrm{phg}}^{0}\left((0, T) \times M_{H}\right)$.
Let $C \in \Psi_{\mathrm{phg}}^{-1}\left((0, T) \times M_{H}\right)$ of principal symbol $c$. We have $\vec{H}_{p} c=\{p, c\} \in S_{\mathrm{phg}}^{0}((0, T) \times$ $\left.M_{H}\right)$ and, for any $k \in \mathbb{N}$,

$$
\left((C P-P C) u^{k}, u^{k}\right)_{\mathcal{H}_{T}}=\left(C P u^{k}, u^{k}\right)_{\mathcal{H}_{T}}-\left(C u^{k}, P u^{k}\right)_{\mathcal{H}_{T}}=0
$$

since $P u^{k}=0$. Taking principal symbols, we get $\left\langle\nu, \vec{H}_{p} c\right\rangle=0$.
Therefore, denoting by $\left(\phi_{t}\right)_{\in \mathbb{R}}$ the sub-Riemannian normal flow introduced in Section 1.4,

$$
0=\left\langle\nu, \vec{H}_{p} c \circ \phi_{t}\right\rangle=\left\langle\nu, \frac{d}{d t} c \circ \phi_{t}\right\rangle=\frac{d}{d t}\left\langle\nu, c \circ \phi_{t}\right\rangle
$$

and hence

$$
\begin{equation*}
\langle\nu, c\rangle=\left\langle\nu, c \circ \phi_{t}\right\rangle . \tag{27}
\end{equation*}
$$

From the second convergence in (23), we can deduce that

$$
\begin{equation*}
\nu=0 \text { in } S^{*}(\mathcal{C}(p)) \cap T^{*}((0, T) \times \operatorname{Supp}(a)) \tag{28}
\end{equation*}
$$

The proof of this fact, which is standard (see for example BG02, Section 6.2]), is given in Appendix C

We note that any geodesic of $M_{H}$ with momentum $\xi \in V_{\varepsilon}$ enters $\omega$ in time at most $\kappa \varepsilon^{-1}$ for some $\kappa>0$ which does not depend on $\varepsilon$ (this is because $\omega=M_{H} \backslash B$, where $B$ is a small ball). Hence, together with (28), the propagation property (27) implies that $\nu \equiv 0$. It follows that $\left\|u^{k}\right\|_{\mathcal{H}_{T}} \rightarrow 0$. By conservation of energy, it is a contradiction with the normalization $\left\|\left(u_{0}^{k}, u_{1}^{k}\right)\right\|_{\mathcal{H} \times L^{2}}=1$. Hence, (22) holds.

Our goal is to remove the $\left\|\left(u_{0}, u_{1}\right)\right\|_{L^{2} \times \mathcal{H}^{\prime}}$ term in (22) in order to get (77). We follow the same scheme of proof as in BS19, Section 8].

Let us denote by $S(t)$ the semigroup of the wave equation, which means that $S(t)\left(u_{0}, u_{1}\right)$ is the solution at time $t$ of (2) associated to the initial data $\left(u_{0}, u_{1}\right) \in \mathcal{H} \times L^{2}$.

For $T>0$, define the space of invisible solutions

$$
\begin{equation*}
\mathcal{N}_{T}=\left\{\left(u_{0}, u_{1}\right) \in \mathcal{H}_{0} \times L^{2}:\left(\left(\mathrm{Op}(a) \partial_{t} u, \partial_{t} u\right)_{L^{2}(M, \mu)}\right)_{\mid[0, T]}=0\right\} \tag{29}
\end{equation*}
$$

where $u$ denotes the solution of (2) with initial data $\left(u_{0}, u_{1}\right)$.
Lemma 19. For any $T>\kappa \varepsilon^{-1}$, there holds $\mathcal{N}_{T}=\{0\}$.
Proof. Fix $T>\kappa \varepsilon^{-1}$. We will prove that there exists $\kappa \varepsilon^{-1}<T_{1} \leqslant T$ such that $\mathcal{N}_{T_{1}}=\{0\}$, which will immediately imply that $\mathcal{N}_{T}=\{0\}$ since $\mathcal{N}_{T^{\prime}} \subset \mathcal{N}_{T^{\prime \prime}}$ for $T^{\prime} \geqslant T^{\prime \prime}$.

Let us remark that for any $T^{\prime}>\kappa \varepsilon^{-1}$, Lemma 18 implies that there exists $C_{T^{\prime}}>0$ such that any $\left(u_{0}, u_{1}\right) \in \mathcal{N}_{T^{\prime}}$ satisfies

$$
C_{T^{\prime}}\left\|\left(u_{0}, u_{1}\right)\right\|_{\mathcal{H} \times L^{2}} \leqslant\left\|\left(u_{0}, u_{1}\right)\right\|_{L^{2} \times \mathcal{H}^{\prime}}
$$

Since $\mathcal{H} \times L^{2}$ is compactly embedded in $L^{2} \times \mathcal{H}^{\prime}$ by Lemma 17 we get that $\operatorname{dim} \mathcal{N}_{T^{\prime}}<+\infty$. Consider the mapping $A(\delta):=\delta^{-1}(S(\delta)-\mathrm{Id}): \mathcal{N}_{T^{\prime}} \rightarrow \mathcal{N}_{T^{\prime}-\delta}$. Since dim $\mathcal{N}_{T^{\prime}}$ is an integer, there exist $\kappa \varepsilon^{-1}<T_{1} \leqslant T$ and $\delta_{0}>0$ such that $\operatorname{dim} \mathcal{N}_{T_{1}-\delta}=\operatorname{dim} \mathcal{N}_{T_{1}}$ for any $\delta \in\left(0, \delta_{0}\right)$. Therefore, $A(\delta)$ is a linear map on $\mathcal{N}_{T_{1}}$. Letting $\delta \rightarrow 0$, we obtain that $\partial_{t}: \mathcal{N}_{T_{1}} \rightarrow \mathcal{N}_{T_{1}}$
which associates to $\left(u_{0}, u_{1}\right) \in \mathcal{N}_{T_{1}}$ the couple $\left(\partial_{t} u_{\mid t=0}, \partial_{t t}^{2} u_{\mid t=0}\right)$ with $u=S(t)\left(u_{0}, u_{1}\right)$ is a well-defined linear operator.

For the sake of a contradiction, assume that $\operatorname{dim} \mathcal{N}_{T_{1}} \geqslant 1$. Then there exists $\lambda \in \mathbb{C}$ an eigenvalue of $\partial_{t}: \mathcal{N}_{T_{1}} \rightarrow \mathcal{N}_{T_{1}}$ and a corresponding eigenfunction $\left(u_{0}, u_{1}\right) \in \mathcal{N}_{T_{1}}$. We consider $u=S(t)\left(u_{0}, u_{1}\right)$. We have $\Delta_{H} u=\partial_{t t}^{2} u=\lambda^{2} u$, which means that $u$ is an eigenfunction of $\Delta_{H}$. Since $u_{\mid \omega}=0$, a result of Bony [Bon69, Corollaire 4.1] ensures that $u=0$, which is a contradiction. Therefore, $\mathcal{N}_{T_{1}}=\{0\}$ and $\mathcal{N}_{T}=\{0\}$.

End of the proof of Theorem 3. We argue again by contradiction. Let us assume that there exists a sequence $\left(u^{k}\right)_{k \in \mathbb{N}}$ of solutions of the wave equation such that $\left(u_{0}^{k}, u_{1}^{k}\right) \in \mathcal{H}_{0} \times L^{2}$, $\left\|\left(u_{0}^{k}, u_{1}^{k}\right)\right\|_{\mathcal{H} \times L^{2}}=1$ for any $k \in \mathbb{N}$ and

$$
\begin{equation*}
\int_{0}^{T}\left|\left(\operatorname{Op}(a) \partial_{t} u^{k}, \partial_{t} u^{k}\right)_{L^{2}\left(M_{H}, \mu\right)}\right| d t \rightarrow 0 \tag{30}
\end{equation*}
$$

as $k \rightarrow+\infty$. The sequence $\left(u_{0}^{k}, u_{1}^{k}\right)$ is bounded in $\mathcal{H}_{0} \times L^{2}$, and therefore, extracting if necessary a subsequence, it converges weakly to some $\left(u_{0}, u_{1}\right) \in \mathcal{H}_{0} \times L^{2}$. Denoting by $u$ the associated solution of (2), it follows that $\partial_{t} u^{k} \rightharpoonup \partial_{t} u$ for the weak topology of $L^{2}\left((0, T) \times M_{H}\right)$.

Let us prove that

$$
\begin{equation*}
\int_{0}^{T}\left|\left(\operatorname{Op}(a) \partial_{t} u, \partial_{t} u\right)_{L^{2}\left(M_{H}, \mu\right)}\right| d t=0 \tag{31}
\end{equation*}
$$

We write $\mathrm{Op}(a)=A^{*} A+B$ with $A \in \Psi_{0}^{\mathrm{phg}}\left(T^{*} M\right)$ and $B \in \Psi_{-1}^{\mathrm{phg}}\left(T^{*} M\right)$ and we have

$$
0 \leqslant \int_{0}^{T}\left(\left(A^{*} A+B\right) \partial_{t} u^{k}, \partial_{t} u^{k}\right)_{L^{2}\left(M_{H}, \mu\right)} d t+\mathrm{o}(1) \underset{k \rightarrow+\infty}{\rightarrow} 0
$$

where the $\mathrm{o}(1)$ comes from Gårding inequality $(\mathrm{Op}(a)$ is asymptotically non-negative). It follows that $\int_{0}^{T}\left(B \partial_{t} u^{k}, \partial_{t} u^{k}\right)_{L^{2}\left(M_{H}, \mu\right)} d t \rightarrow 0$, and since $B \partial_{t} u^{k} \rightarrow B \partial_{t} u$ strongly in $L^{2}((0, T) \times$ $M_{H}$ ), we get that

$$
\int_{0}^{T}\left(B \partial_{t} u, \partial_{t} u\right)_{L^{2}\left(M_{H}, \mu\right)} d t=0
$$

We also have

$$
\left\|A \partial_{t} u\right\|_{L^{2}\left((0, T) \times M_{H}\right)} \leqslant \liminf _{k \rightarrow+\infty}\left\|A \partial_{t} u^{k}\right\|_{L^{2}\left((0, T) \times M_{H}\right)}=0 .
$$

These last two lines imply (31).
From (31), we deduce that $u \in \mathcal{N}_{T}$. From Lemma 19, we deduce that $u=0$. In particular, we have $\left(u_{0}, u_{1}\right)=(0,0)$, and $\left(u_{0}^{k}, u_{k}^{1}\right)$ converges to $(0,0)$ for the weak topology of $\mathcal{H}_{0} \times L^{2}$, and hence for the strong topology of $L^{2} \times \mathcal{H}_{0}^{\prime}$ by Lemma 17. Together with (30) and the fact that $\left\|\left(u_{0}^{k}, u_{1}^{k}\right)\right\|_{\mathcal{H} \times L^{2}}=1$, this contradicts (22). It concludes the proof of Theorem 3,

## A Pseudodifferential calculus

We denote by $\Omega$ an open set of a $d$-dimensional manifold (typically $d=n$ or $d=n+1$ with the notations of this paper) equipped with a smooth volume $\mu$. We denote by $q$ the variable in $\Omega$, typically $q=x$ or $q=(t, x)$ with our notations. Let $\pi: T^{*} \Omega \rightarrow \Omega$ be the canonical projection. We recall briefly some facts concerning pseudodifferential calculus, following Hör07, Chapter 18].

We denote by $S_{\text {hom }}^{m}\left(T^{*} \Omega\right)$ the set of homogeneous symbols of degree $m$ with compact support in $\Omega$. We also write $S_{\mathrm{phg}}^{m}\left(T^{*} \Omega\right)$ the set of polyhomogeneous symbols of degree $m$ with compact support in $\Omega$. Hence, $a \in S_{\mathrm{phg}}^{m}\left(T^{*} \Omega\right)$ if $a \in C^{\infty}\left(T^{*} \Omega\right), \pi(\operatorname{Supp}(a))$ is a compact of $\Omega$, and there exist $a_{j} \in S_{\mathrm{hom}}^{m-j}\left(T^{*} \Omega\right)$ such that for all $N \in \mathbb{N}, a-\sum_{j=0}^{N} a_{j} \in S_{\mathrm{phg}}^{m-N-1}\left(T^{*} \Omega\right)$. We denote by $\Psi_{\mathrm{phg}}^{m}\left(T^{*} \Omega\right)$ the space of polyhomogeneous pseudodifferential operators of order $m$ on $\Omega$, with a compactly supported kernel in $\Omega \times \Omega$. For $A \in \Psi_{\mathrm{phg}}^{m}(\Omega)$, we denote by
$\sigma_{p}(A) \in S_{\mathrm{phg}}^{m}\left(T^{*} \Omega\right)$ the principal symbol of $A$. The sub-principal symbol is characterized by the action of pseudodifferential operators on oscillating functions: if $A \in \Psi_{\mathrm{phg}}^{m}(\Omega)$ and $f(q)=b(q) e^{i k S(q)}$ with $b, S$ smooth and real-valued, then

$$
\int_{\Omega} A(f) \bar{f} d \mu=k^{m} \int_{\Omega}\left(\sigma_{p}(A)\left(q, S^{\prime}(q)\right)+\frac{1}{k} \sigma_{\operatorname{sub}}(A)\left(q, S^{\prime}(q)\right)\right)|f(q)|^{2} d \mu(q)+O\left(k^{m-2}\right)
$$

A quantization is a continuous linear mapping

$$
\mathrm{Op}: S_{\mathrm{phg}}^{m}\left(T^{*} \Omega\right) \rightarrow \Psi_{\mathrm{phg}}^{m}(\Omega)
$$

satisfying $\sigma_{p}(\operatorname{Op}(a))=a$. An example of quantization is obtained by using partitions of unity and, locally, the Weyl quantization, which is given in local coordinates by

$$
\mathrm{Op}^{W}(a) f(q)=\frac{1}{(2 \pi)^{d}} \int_{\mathbb{R}_{q^{\prime}}^{d} \times \mathbb{R}_{p}^{d}} e^{i\left\langle q-q^{\prime}, p\right\rangle} a\left(\frac{q+q^{\prime}}{2}, p\right) f\left(q^{\prime}\right) d q^{\prime} d p .
$$

We have the following properties:

1. If $A \in \Psi_{\mathrm{phg}}^{l}(\Omega)$ and $B \in \Psi_{\mathrm{phg}}^{m}(\Omega)$, then $[A, B] \in \Psi_{\mathrm{phg}}^{l+m-1}(\Omega)$ and $\sigma_{p}([A, B])=$ $\frac{1}{i}\left\{\sigma_{p}(a), \sigma_{p}(b)\right\}$ where the Poisson bracket is taken with respect to the canonical symplectic structure of $T^{*} \Omega$.
2. If $X$ is a vector field on $\Omega$ and $X^{*}$ is its formal adjoint in $L^{2}(\Omega, \mu)$, then $X^{*} X \in \Psi_{\mathrm{phg}}^{2}(\Omega)$ and $\sigma_{p}\left(X^{*} X\right)=h_{X}^{2}$.
3. If $A \in \Psi_{\mathrm{phg}}^{m}(\Omega)$, then A maps continuously the space $H^{s}(\Omega)$ to the space $H^{s-m}(\Omega)$.

## B Proof of Proposition 8

Let us first prove the result when $M \subset \mathbb{R}^{n}$, following the proof of Ral82.
In this proof, we suppress the time variable $t$. Thus we use $x=\left(x_{0}, x_{1}, \ldots, x_{n}\right)$ where $x_{0}=t$ in our earlier notations. Similarly, $\xi=\left(\xi_{0}, \xi_{1}, \ldots, \xi_{n}\right)$ where $\xi_{0}=\tau$ previously. Let $\Gamma$ be the projection a null-bicharacteristic curve, i.e., a solution of (5) given by $x(s) \in \mathbb{R}^{n+1}$. We insist on the fact that the bicharacteristics are parametrized by $s$, as in (5). We consider functions of the form

$$
v_{k}(x)=k^{\frac{n}{4}-1} a_{0}(x) e^{i k \psi(x)}
$$

We would like to choose $\psi(x)$ such that for all $s \in \mathbb{R}, \psi(x(s))$ is real-valued and $\operatorname{Im} \frac{\partial^{2} \psi}{\partial x_{i} \partial x_{j}}(x(s))$ is positive definite on vectors orthogonal to $\dot{x}(s)$. Roughly speaking, $\left|e^{i k \psi(x)}\right|$ will then look like a Gaussian distribution on planes perpendicular to $\Gamma$ in $\mathbb{R}^{n+1}$.

We first observe that $\partial_{t t}^{2} v_{k}-\Delta v_{k}$ may be decomposed as

$$
\partial_{t t}^{2} v_{k}-\Delta v_{k}=\left(k^{\frac{n}{4}+1} A_{1}+k^{\frac{n}{4}} A_{2}+k^{\frac{n}{4}-1} A_{3}\right) e^{i k \psi}
$$

with

$$
\begin{aligned}
& A_{1}(x)=p_{2}\left(x, \frac{\partial \psi(x)}{\partial x}\right) a_{0}(x) \\
& A_{2}(x)=L a_{0}(x) \\
& A_{3}(x)=\partial_{t t}^{2} a_{0}(x)-\Delta a_{0}(x)
\end{aligned}
$$

Here we have set

$$
\begin{equation*}
L a_{0}=\frac{1}{i}\left(\frac{\partial p_{2}}{\partial \xi_{j}}\left(x, \frac{\partial \psi}{\partial x}\right) \frac{\partial a_{0}}{\partial x_{j}}\right)+\left(\frac{1}{2 i} \frac{\partial^{2} p_{2}}{\partial \xi_{j} \partial \xi_{k}}\left(x, \frac{\partial \psi}{\partial x}\right) \frac{\partial^{2} \psi}{\partial x_{j} \partial x_{k}}+p_{1}\left(x, \frac{\partial \psi}{\partial x}\right)\right) a_{0} \tag{32}
\end{equation*}
$$

and $p_{1}$ is the sub-principal symbol of $\partial_{t t}^{2}-\Delta$ (see Appendix A).
We will use several times the following classical lemma. It is particularly useful to estimate oscillating integrals and its proof is given in Ral82, Lemma 2.8].

Lemma 20. Let $c(x)$ be a function on $\mathbb{R}^{n+1}$ which vanishes at order $S-1$ on a curve $\Gamma$ for some $S \geqslant 1$. Suppose that Supp $c \cap\left\{\left|x_{0}\right| \leqslant T\right\}$ is compact and that Im $\psi(x) \geqslant a d^{2}(x)$ on this set for some constant $a>0$, where $d(x)$ denotes the distance from the point $x \in \mathbb{R}^{d+1}$ to the curve $\Gamma$. Then there exists a constant $C$ such that

$$
\int_{\left|x_{0}\right| \leqslant T}\left|c(x) e^{i k \psi(x)}\right|^{2} d x \leqslant C k^{-S-n / 2}
$$

In what follows, we construct $a_{0}$ and $\psi$ so that $A_{1}(x)$ vanishes at order 2 along $\Gamma$ and $A_{2}(x)$ vanishes at order 0 along the same curve. We will then be able to use Lemma 20 with $S=3$ and $S=1$ respectively.

Analysis of $A_{1}(x)$. Our goal is to show that, if we choose $\psi$ adequately, we can make the quantity

$$
f(x)=p_{2}\left(x, \frac{\partial \psi(x)}{\partial x}\right)
$$

vanish at order 2 on $\Gamma$. For the vanishing at order 0 , we prescribe $\frac{\partial \psi}{\partial x}(x(s))=\xi(s)$, and then $f(x(s))=0$ since $(x(s), \xi(s))$ is a null-bicharacteristic. For the vanishing at order 1 , using (5), we remark that

$$
\begin{align*}
\frac{\partial f}{\partial x_{j}}(x(s)) & =\frac{\partial p_{2}}{\partial x_{j}}(x(s))+\sum_{k=0}^{n} \frac{\partial p_{2}}{\partial \xi_{k}}(x(s)) \frac{\partial \psi}{\partial x_{j} \partial x_{k}}(x(s)) \\
& =-\dot{\xi}_{j}(s)+\sum_{k=0}^{n} \dot{x}_{k}(s) \frac{\partial \psi}{\partial x_{j} \partial x_{k}}(x(s))  \tag{33}\\
& =-\frac{d}{d s}\left(\frac{\partial \psi}{\partial x_{j}}(x(s))\right)+\sum_{k=0}^{n} \dot{x}_{k}(s) \frac{\partial \psi}{\partial x_{j} \partial x_{k}}(x(s)) \\
& =0
\end{align*}
$$

Therefore, $f$ vanishes automatically at order 1 along $\Gamma$ (without making any particular choice for $\psi$ ): it just follows from the bicharacteristic equations (5). But for $f(x)$ to vanish at order 2 along $\Gamma$, it is required to choose a particular $\psi$. Using the Einstein summation notation, we want

$$
\begin{aligned}
0 & =\frac{\partial^{2} f}{\partial x_{j} \partial x_{i}} \\
& =\frac{\partial^{2} p_{2}}{\partial x_{j} \partial x_{i}}+\frac{\partial^{2} p_{2}}{\partial \xi_{k} \partial x_{i}} \frac{\partial^{2} \psi}{\partial x_{j} \partial x_{k}}+\frac{\partial^{2} p_{2}}{\partial x_{j} \partial \xi_{k}} \frac{\partial^{2} \psi}{\partial x_{i} \partial x_{k}}+\frac{\partial^{2} p_{2}}{\partial \xi_{l} \partial \xi_{k}} \frac{\partial^{2} \psi}{\partial x_{i} \partial x_{k}} \frac{\partial^{2} \psi}{\partial x_{j} \partial x_{l}}+\frac{\partial p_{2}}{\partial \xi_{k}} \frac{\partial^{3} \psi}{\partial x_{j} \partial x_{k} \partial x_{i}}
\end{aligned}
$$

to hold along $\Gamma$ for $i, j=0,1, \ldots, n$. Introducing the matrices

$$
\begin{array}{ll}
(M(s))_{i j}=\frac{\partial^{2} \psi}{\partial x_{i} \partial x_{j}}(x(s)), & (A(s))_{i j}=\frac{\partial^{2} p_{2}}{\partial x_{i} \partial x_{j}}(x(s), \xi(s)), \\
(B(s))_{i j}=\frac{\partial^{2} p_{2}}{\partial \xi_{i} \partial x_{j}}(x(s), \xi(s)), & (C(s))_{i j}=\frac{\partial^{2} p_{2}}{\partial \xi_{i} \partial \xi_{j}}(x(s), \xi(s))
\end{array}
$$

this amounts to solving the matricial Riccati equation

$$
\begin{equation*}
\frac{d M}{d s}+M C M+B^{T} M+M B+A=0 \tag{34}
\end{equation*}
$$

on the time-interval $(0, T)$. While solving (34), we also require $M(s)$ to be symmetric, $\operatorname{Im}(M(s))$ to be positive definite on the orthogonal complement of $\dot{x}(s)$, and $M(s) \dot{x}(s)=\dot{\xi}(s)$ to hold for all $s \in \mathbb{R}$ due to (33).

It is shown in Ral82 that for any $(n+1) \times(n+1)$ matrix $M_{0}$ with $\operatorname{Im}\left(M_{0}\right)>0$ on the orthogonal complement of $\dot{x}(0)$ and $M_{0} \dot{x}(0)=\dot{\xi}(0)$, there exists a global solution $M(s)$ on
$[0, T]$ of (34) which satisfies all these conditions and such that $M(0)=M_{0}$. The proof just requires that $A, C$ are symmetric, but does not need anything special about $p_{2}$ (in particular, it applies to our sub-Riemannian case where $p_{2}$ is degenerate). For the sake of completeness, we recall the proof below.

We first extend $y^{0}=\dot{x}(0)$ to a basis $y^{0}, \ldots, y^{n}$ of $\mathbb{R}^{n+1}$, and we define $\eta^{i}=M(0) y^{i}$. Finally, $M(s)$ is defined by the equations

$$
\begin{equation*}
M(s) y^{i}(s)=\eta^{i}(s), \quad i=0, \ldots, n \tag{35}
\end{equation*}
$$

where $\left(y^{i}(s), \eta^{i}(s)\right)$ is the solution with initial data $\left(y^{i}, \eta^{i}\right)$ to the linear system

$$
\begin{cases}\dot{y}=B y+C \eta, & y^{i}(0)=y^{i}  \tag{36}\\ \dot{\eta}=-A y-B^{T} \eta & \eta^{i}(0)=\eta^{i}\end{cases}
$$

All the coefficients in (36) are real and $A$ and $C$ are symmetric, therefore the flow defined by (36) preserves both the real symplectic form acting on pairs $(y, \eta) \in\left(\mathbb{R}^{n}\right)^{2}$ and $\left(y^{\prime}, \eta^{\prime}\right) \in$ $\left(\mathbb{R}^{n}\right)^{2}$ given by

$$
\sigma\left((y, \eta),\left(y^{\prime}, \eta^{\prime}\right)\right)=y \cdot \eta^{\prime}-\eta \cdot y^{\prime}
$$

and the complexified form $\sigma_{\mathbb{C}}\left((y, \eta),\left(y^{\prime}, \eta^{\prime}\right)\right)=\sigma\left((y, \eta),\left(\overline{y^{\prime}}, \overline{\eta^{\prime}}\right)\right)$ for $(y, \eta) \in\left(\mathbb{C}^{n}\right)^{2}$ and $(\tilde{y}, \tilde{\eta}) \in$ $\left(\mathbb{C}^{n}\right)^{2}$. When we say that $\sigma_{\mathbb{C}}$ is invariant under (36), it means that we allow complex initial data in (36). Note that, decomposing $y=y_{1}+i y_{2}$ (with $y_{1}, y_{2} \in \mathbb{R}^{n}$ ) and similarly for $\eta$, we have in particular

$$
\sigma_{\mathbb{C}}((y, \eta),(y, \eta))=-2 i\left(y_{1} \cdot \eta_{2}-y_{2} \cdot \eta_{1}\right)
$$

Thus, if $\sum_{i=0}^{n} b_{i} y^{i}\left(s_{0}\right)=0$ with $b_{i} \in \mathbb{R}$, and if we set $v=\sum_{i=0}^{n} b_{i}\left(y^{i}\left(s_{0}\right), \eta^{i}\left(s_{0}\right)\right)$, then $0=\sigma_{\mathbb{C}}(v, v)$, and therefore

$$
-2 i y \cdot \operatorname{Im}(M(0)) \bar{y}=0
$$

where $y=\sum_{i=0}^{n} b_{i} y^{i}(0)$. From this, we deduce that $b_{i}=0$ for $i=0, \ldots, n$, and $M(s)$ is well-defined by (35). One may check directly thanks to (36) that $M$ satisfies (34). Moreover, the invariance of $\sigma$ implies that $M(s)$ is symmetric. We may also check that $\left(y^{0}(s), \eta^{0}(s)\right)=$ $(\dot{x}(s), \dot{\xi}(s))$, which implies that $M(s) \dot{x}(s)=\dot{\xi}(s)$. Together with the invariance of $\sigma_{\mathbb{C}}$, it also implies that $\operatorname{Im}(M(s))$ is positive definite on the orthogonal complement of $\dot{x}(s)$. Therefore, we found a choice for the second order derivatives of $\psi$ along $\Gamma$ which meets all our conditions. For $x=\left(t, x^{\prime}\right) \in \mathbb{R} \times \mathbb{R}^{n}$ and $s$ such that $t=t(s)$, we set

$$
\begin{equation*}
\psi(x)=\xi^{\prime}(s) \cdot\left(x^{\prime}-x^{\prime}(s)\right)+\frac{1}{2}\left(x^{\prime}-x^{\prime}(s)\right) \cdot M(s)\left(x^{\prime}-x^{\prime}(s)\right) \tag{37}
\end{equation*}
$$

To sum up, as in the Riemannian (or "strictly hyperbolic") case handled by Ralston in [Ral82], the key observation is that the invariance of $\sigma$ and $\sigma_{\mathbb{C}}$ prevents the solutions of (34) with positive imaginary part on the orthogonal complement of $\dot{x}(0)$ to blowup.

Analysis of $A_{2}(x)$. We note that $A_{2}$ vanishes along $\Gamma$ if and only if $L a_{0}(x(s))=0$. According to (32), this turns out to be a linear transport equation on $a_{0}$, which has a solution. Note that, given $a_{0}$ at $(t=0, x=x(0))$, this determines $a_{0}$ only along $\Gamma$, and we may choose $a_{0}$ in a smooth (and arbitrary) way outside $\Gamma$. We choose it to vanish outside a small neighborhood of $\Gamma$.

Proof of (9). We apply Lemma 20 to $S=3, c=A_{1}$ and to $S=1, c=A_{2}$, and we get

$$
\left\|\partial_{t t}^{2} v_{k}-\Delta v_{k}\right\|_{L^{1}\left(0, T ; L^{2}(M)\right)} \leqslant C\left(k^{-\frac{1}{2}}+k^{-\frac{1}{2}}+k^{-1}\right)
$$

which implies (9).

Proof of (10). We first observe that since $\operatorname{Im}(M(s))$ is positive definite on the orthogonal complement of $\dot{x}(s)$ and continuous as a function of $s$, there exist $\alpha, C>0$ such that for any $t(s) \in[0, T]$ and any $x^{\prime} \in M$,

$$
\left|\partial_{t} v_{k}\left(t(s), x^{\prime}\right)\right|^{2}+\left|\partial_{t} v_{k}\left(t(s), x^{\prime}\right)\right|^{2} \geqslant C k^{\frac{n}{2}} e^{-\alpha k d\left(x^{\prime}, x^{\prime}(s)\right)^{2}}
$$

where $d(\cdot, \cdot)$ denotes the Euclidean distance in $\mathbb{R}^{n}$. Using the observation that for any function $f$,

$$
\begin{equation*}
\int_{M} f\left(x^{\prime}\right) e^{-\alpha k d\left(x^{\prime}, x^{\prime}(s)\right)^{2}} d \mu\left(x^{\prime}\right) \sim \frac{\pi^{n / 2}}{k^{n / 2} \sqrt{\alpha}} f\left(x^{\prime}(s)\right) \frac{d \mu}{d \ell_{n}}\left(x^{\prime}(s)\right) \tag{38}
\end{equation*}
$$

as $k \rightarrow+\infty$, we obtain (10). Here $\ell_{n}$ denotes the Lebesgue measure on $\mathbb{R}^{n}$.
Proof of (11). We observe that since $\operatorname{Im}(M(s)$ ) is positive definite (uniformy in $s$ ) on the orthogonal complement of $\dot{x}(s)$, there exist $C, \alpha^{\prime}>0$ such that for any $t \in[0, T]$, for any $x^{\prime} \in M,\left|\partial_{t} v_{k}\left(t(s), x^{\prime}\right)\right|$ and $\left|\nabla^{\mathrm{sR}} v_{k}\left(t(s), x^{\prime}\right)\right|$ are both bounded above by $C k^{\frac{n}{4}} e^{-\alpha^{\prime} k d\left(x^{\prime}, x^{\prime}(s)\right)^{2}}$. Therefore

$$
\begin{align*}
& \int_{M \backslash V_{t(s)}}\left(\left|\partial_{t} v_{k}\left(t(s), x^{\prime}\right)\right|^{2}+\left|\nabla^{\mathrm{sR}} v_{k}\left(t(s), x^{\prime}\right)\right|^{2}\right) d \mu\left(x^{\prime}\right) \\
& \quad \leqslant C k^{n / 2} \int_{M \backslash V_{t(s)}} e^{-2 \alpha^{\prime} k d\left(x^{\prime}, x^{\prime}(s)\right)^{2}} d \mu\left(x^{\prime}\right) \\
& \leqslant C k^{n / 2} \int_{M \backslash V_{t(s)}} e^{-2 \alpha^{\prime} k d\left(x^{\prime}, x^{\prime}(s)\right)^{2}} d \ell_{n}\left(x^{\prime}\right)+\mathrm{o}(1) \tag{39}
\end{align*}
$$

where, in the last line, we used the fact that $\left|d \mu / d \ell_{n}\right| \leqslant C$ in a fixed compact subset of $M$ (since $\mu$ is a smooth volume), and the o(1) comes from the eventual blowup of $\mu$ at the boundary of $M$.

Now, $M \subset \mathbb{R}^{n}$, and there exists $r>0$ such that $B_{d}(x(s), r) \subset V_{t(s)}$ for any $s$ such that $t(s) \in\left(0, T_{0}\right)$, where $d(\cdot, \cdot)$ still denotes the Euclidean distance in $\mathbb{R}^{n}$. Therefore, we bound above the integral in (39) by

$$
\begin{equation*}
C k^{n / 2} \int_{\mathbb{R}^{n} \backslash B_{d}(x(s), r)} e^{-2 \alpha^{\prime} k d\left(x^{\prime}, x^{\prime}(s)\right)^{2}} d \ell_{n}\left(x^{\prime}\right) \tag{40}
\end{equation*}
$$

Making the change of variables $y=k^{-1 / 2}(y-x(s))$, we bound above (40) by

$$
C \int_{\mathbb{R}^{n} \backslash B_{d}\left(0, r k^{1 / 2}\right)} e^{-2 \alpha^{\prime}\|y\|^{2}} d \ell_{n}(y)
$$

with $\|\cdot\|$ the Euclidean norm. This last expression is bounded above by

$$
C e^{-\alpha^{\prime} r^{2} k} \int_{\mathbb{R}^{n}} e^{-\alpha^{\prime}\|y\|^{2}} d \ell_{n}(y)
$$

which implies (11).
Extension of the result to any manifold $M$. In the case of a general manifold $M$, not necessarily included in $\mathbb{R}^{n}$, we use charts together with the above construction. We cover $M$ by a set of charts $\left(U_{\alpha}, \varphi_{\alpha}\right)$, where $\left(U_{\alpha}\right)$ is a family of open sets of $M$ covering $M$ and $\varphi_{\alpha}: U_{\alpha} \rightarrow \mathbb{R}^{n}$ is an homeomorphism $U_{\alpha}$ onto an open subset of $\mathbb{R}^{n}$. Take a solution $(x(t), \xi(t))_{t \in[0, T]}$ of (6). It visits a finite number of charts in the order $U_{\alpha_{1}}, U_{\alpha_{2}}, \ldots$, and we choose the charts and $a_{0}$ so that $v_{k}(t, \cdot)$ is supported in a unique chart at each time $t$. The above construction shows how to construct $a_{0}$ and $\psi$ as long as $x(t)$ remains in the same chart. For any $l \geqslant 1$, we choose $t_{l}$ so that $x\left(t_{l}\right) \in U_{\alpha_{l}} \cap U_{\alpha_{l+1}}$ and $a_{0}\left(t_{l}, \cdot\right)$ is supported in $U_{\alpha_{l}} \cap U_{\alpha_{l+1}}$. Since there is a (local) solution $v_{k}$ for any choice of initial $a_{0}\left(t_{l}, x\left(t_{l}\right)\right)$ and $\operatorname{Im}\left(\frac{\partial^{2} \psi}{\partial x_{i} \partial x_{j}}\right)\left(t_{l}, x\left(t_{l}\right)\right)$ in Proposition 8, we see that $v_{k}$ may be continued from the chart $U_{\alpha_{l}}$ to the chart $U_{\alpha_{l+1}}$. This continuation is smooth since the two solutions coincide as long as $a_{0}(t, \cdot)$ is supported in $U_{\alpha_{l}} \cap U_{\alpha_{l+1}}$. Patching all solutions on the time intervals $\left[t_{l}, t_{l+1}\right]$ together, it yields a global in time solution $v_{k}$, as desired.

## C Proof of

Because of the second convergence in (23), it amounts to proving that

$$
\int_{0}^{T}\left|\left(\mathrm{Op}(a) \nabla^{\mathrm{sR}} u^{k}, \nabla^{\mathrm{sR}} u^{k}\right)_{L^{2}\left(M_{H}, \mu\right)}\right| d t \rightarrow 0
$$

Replacing $\mathrm{Op}(a)$ by its Friedrichs quantization $\mathrm{Op}^{F}(a)$, which is positive (see Tay74, Chapter VII]), we see that it is sufficient to prove

$$
\begin{equation*}
\left(\mathrm{Op}^{F}(a) \nabla^{\mathrm{sR}} u^{k}, \nabla^{\mathrm{sR}} u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)} \rightarrow 0 \tag{41}
\end{equation*}
$$

Let $\delta>0$ and $\tilde{a} \in S_{\mathrm{phg}}^{0}\left((-\delta, T+\delta) \times M_{H}\right), 0 \leqslant \tilde{a} \leqslant \sup (a)$ and such that $\tilde{a}=a$ for $0 \leqslant t \leqslant T$. Using integrations by parts, we have

$$
\begin{aligned}
\left(\mathrm{Op}^{F}(\tilde{a}) \nabla^{\mathrm{sR}} u^{k}, \nabla^{\left.\mathrm{sR}^{k} u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)}}\right. & =-\left(\operatorname{div}_{\mu}\left(\mathrm{Op}^{F}(\tilde{a}) \nabla^{\mathrm{sR}} u^{k}\right), u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)} \\
& =-\left(\operatorname{Op}^{F}(\tilde{a}) \Delta u^{k}, u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)}+\mathrm{o}(1) \\
& =-\left(\operatorname{Op}^{F}(\tilde{a}) \partial_{t}^{2} u^{k}, u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)}+\mathrm{o}(1) \\
& =-\left(\partial_{t} \operatorname{Op}^{F}(\tilde{a}) \partial_{t} u^{k}, u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)}+\mathrm{o}(1) \\
& =\left(\operatorname{Op}^{F}(\tilde{a}) \partial_{t} u^{k}, \partial_{t} u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)}+\mathrm{o}(1)
\end{aligned}
$$

where the $\mathrm{o}(1)$ comes from direct commutator estimates together with the fact that $\left\|u^{k}\right\|_{\mathcal{H}_{T}}=$ 1. Finally we note that since $\mathrm{Op}^{F}$ is a positive quantization, we have

$$
\begin{aligned}
\left(\mathrm{Op}^{F}(a) \nabla^{\mathrm{sR}} u^{k}, \nabla^{\mathrm{sR}} u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)} & \leqslant\left(\mathrm{Op}^{F}(\tilde{a}) \nabla^{\mathrm{sR}} u^{k}, \nabla^{\mathrm{sR}} u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)} \\
& =\left(\operatorname{Op}^{F}(\tilde{a}) \partial_{t} u^{k}, \partial_{t} u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)}+\mathrm{o}(1) \\
& \leqslant C \delta+\left(\mathrm{Op}^{F}(a) \partial_{t} u^{k}, \partial_{t} u^{k}\right)_{L^{2}\left((0, T) \times M_{H}\right)}+\mathrm{o}(1)
\end{aligned}
$$

where $C$ does not depend on $\delta$. Making $\delta \rightarrow 0$, it concludes the proof of (41), and consequently (28) holds.

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