## Chapter 3. Dispersive Behavior

### 3.1. Orientation.

In this chapter we return to Fourier analysis techniques as in Chapter 1. The Fourier transform of the solution is written exactly and then analysed.
The results show how the geometry of the characteristic variety of $L=L_{1}\left(\partial_{y}\right)$ is reflected in qualitative properties of the solutions of $L u=0$. The main idea is that when the characteristic variety is curved, the corresponding solutions tend to spread out in space. This dispersive effect is reflected in solutions becoming smaller in $L^{\infty}\left(\mathbb{R}^{d}\right)$ in contast to $L^{2}\left(\mathbb{R}^{d}\right)$ conservation.

Three simple examples illustrate the theme. The scalar advection operator

$$
\begin{equation*}
L:=\partial_{t}+\mathbf{v} \cdot \partial_{x} \tag{3.1.1}
\end{equation*}
$$

in dimension $d$ and the system

$$
\frac{\partial v}{\partial t}+\left(\begin{array}{cc}
1 & 0  \tag{3.1.2}\\
0 & -1
\end{array}\right) \frac{\partial v}{\partial x}=0
$$

in dimension $d=1$ have only purely translating modes. The characteristic variety of (3.1.1) is the hyperplane $\tau+\mathbf{v} . \xi=0$ and for (3.1.2) it is the pair of lines $\tau \pm \xi=0$. In both cases the variety is not curved at all.
The system analogue of $\square_{1+2}$,

$$
L:=\partial_{t}+\left(\begin{array}{cc}
1 & 0  \tag{3.1.3}\\
0 & -1
\end{array}\right) \partial_{1}+\left(\begin{array}{ll}
0 & 1 \\
1 & 0
\end{array}\right) \partial_{2}
$$

behaves differently. Each component satisfies $\square_{1+2} u=0$. For smooth compactly supported data, they decay (in sup norm) as $t^{-1 / 2}$. The characteristic variety is $\tau^{2}-|\xi|^{2}=0$. Since all charateristic varieties are conic their Gauss curvatures vanish. The present variety intersects $\tau=1$ in a strictly convex set. So the variety is as curved as a conic set can be.

Exercise. Prove the decay rate for compactly supported solutions of $\square_{1+2} u=0$ by expressing solutions as convolutions with fundamental solution(s).

For all three examples the $L^{2}\left(\mathbb{R}^{d}\right)$ norm is perserved during the time evolution.
For the solutions of the transport equation associated to (3.1.1), the size of the support of solutions does not change in time. For (3.1.3), solutions spread out over a set whose two dimensional area grows with time. The spread together with $L^{2}$ conservation, explains the decay.
In optics, the word dispersion is used to mean that the speed of light depends on its wavelength. In that sense, none of the above models is dispersive. The dispersion relations of the first and third models are

$$
\tau=-\mathbf{v} \cdot \xi, \quad \text { and } \quad \tau= \pm|\xi|
$$

Both are positive homogeneous of degree one in $\xi$. Therefore the group velocities $-\nabla_{\xi} \tau$ satisfy

$$
\left|\nabla_{\xi} \tau\right|=|\mathbf{v}|, \quad \text { and } \quad\left|\nabla_{\xi} \tau\right|=1
$$

respectively. In neither case does the speed depend on the wavelength. However for (3.1.2), the velocity depends strongly on $\xi$, though not on $|\xi|$. The fact that the group velocities point in different directions has the effect of spreading the solution, and for large time the solutions decay.
The variation of the group velocity with $\xi$ is given by the matrix of second derivatives $\nabla_{\xi}^{2} \tau$. For our homogeneous operators, $\nabla_{\xi} \tau$ is homogeneous of degree zero, so $\xi$ belongs to the kernel of matrix. The rank can be at most $d-1$. The D'Alembertian $\square_{1+d}$ achieves this maximal rank so is as dispersive as a homogeneous operator can be.
At the extreme opposite is $\nabla_{\xi}^{2} \tau \equiv 0$, in which case the dispersion relation is linear in $\xi$. The associated graph is a hyperplane which belongs to the characteristic variety. The characteristic variety for (3.1.1) and (3.1.2) consist of hyperplanes while for (3.1.3) the variety is curved. Where the variety is flat, $\tau=-\mathbf{v} . \xi$, the group velocity is identically equal to $\mathbf{v}$ so does not depend on $\xi$. This is the completely nondispersive situation. Solutions translate without spread.
If the variety contains no hyperplanes, the variation of the group velocity spreads wavepackets. We will show that as $t \rightarrow \infty$, solutions decay in $L^{\infty}$. These results, presented in $\S 3.2-\S 3.3$, are taken from [JMR, Indiana Math. J. 1998].
An even stronger notion of uniform dispersion is when the rank of $\nabla_{\xi}^{2} \tau$ is everywhere equal to $d-1$. In this case, the sheets of the characteristic variety are uniformly convex cones and smooth compactly supported solutions decay at the rate $t^{-(d-1) / 2}$ as $t \rightarrow \infty$. This is investigated in $\S 3.4$. In $\S 3.4 .1 L^{1} \rightarrow L^{\infty}$ decay estimates are proved. These are applied in $\S 6.7$ to prove global solvability of for nonlinear problems with small initial data and high dimension. In §3.4.3 they are used to derive Strichartz estimates. In §6.8, these estimates are applied to study the nonlinear Klein-Gordon equation in the natural energy space.

## §3.2. Spectral decomposition of solutions.

Since $(\tau, 0)$ is noncharacteristic for $L$, any hyperplane $\{a \tau+b . \xi=0\}$ contained in the characteristic variety must have $a \neq 0$. Therefore, it is necessarily a graph $\{\tau=-\mathbf{v} . \xi\}$.
Over each $\xi \in \mathbb{R}^{d}$ there are at most $N$ points in the characteristic variety. Therefore, the number of distinct hyperplanes in the variety can be no larger than $N$. Denote by $0 \leq M \leq N$ the number of such hyperplanes, $H_{1}, \ldots, H_{M}$,

$$
\begin{equation*}
H_{j}=\left\{(\tau, \xi): \tau=-\mathbf{v}_{j} \cdot \xi\right\}, \quad j=1, \ldots, M \leq N \tag{3.2.1}
\end{equation*}
$$

Examples. 1. When $d=1$ the characteristic variety is a union of lines so consists only of hyperplanes. There are no curved sheets.
2. The operator from (3.1.3) has characteristic variety is given by $\tau^{2}=|\xi|^{2}$ so the variety is the classical light cone, and there are no hyperplanes.
3. The characteristic varieties of Mawell's Equations and the the linearization at $u=0$ of the compressible Euler equations are the union of a convex light cone and a single horizontal hyperplane $\tau=0$.

Definition. A wave number $\omega \in \mathbb{R}^{d} \backslash\{0\}$ is good when there is a neighborhood $\Omega$ of $\omega$ and a finite number of real valued real analytic functions $\lambda_{1}(\xi)<\lambda_{2}(\xi)<\cdots<\lambda_{m}(\xi)$ so that the spectrum of $\sum_{j=1}^{d} A_{j} \xi_{j}$ is $\left\{\lambda_{1}(\xi), \ldots, \lambda_{m}(\xi)\right\}$ for $\xi \in \Omega$. The complementary set consists of bad wave numbers. The set of bad wave numbers is denoted $\mathcal{B}(L)$.

Over a good $\xi$, the characteristic variety of $L$ cotains exactly $m$ nonintersecting sheets $\tau=-\lambda_{j}(\xi)$. At bad points, eigenvalues cross and multiplicities change. The examples above have no bad points.

Examples. Consider the characteristic equation $\left(\tau^{2}-|\xi|^{2}\right)\left(\tau-c \xi_{1}\right)=0$ with $c \in \mathbb{R}$. If $|c|<1$ then the variety is a cone and a hyperplane intersecting only at the origin and all points are good. If $|c|>1$ the plane and cone intersect in a cone whose projection on $\xi$ space is the set of bad points

$$
\mathcal{B}=\left\{\xi:\left(c^{2}-1\right) \xi_{1}^{2}=\xi_{2}^{2}+\ldots+\xi_{d}^{2}\right\} .
$$

When $|c|=1, \mathcal{B}(L)$ degenerates to a line of tangency.

Proposition 3.2.1. i. $\mathcal{B}(L)$ is a closed conic set of measure zero in $\mathbb{R}^{d} \backslash\{0\}$.
ii. The complementary set, $\mathbb{R}^{d} \backslash(\mathcal{B} \cup\{0\})$, is the disjoint union of a finite family of conic connected open sets $\Omega_{g} \subset \mathbb{R}^{d} \backslash\{0\}, g \in \mathcal{G}$.
iii. The mulitplicity of $\tau=-\mathbf{v}_{j} . \xi$ as a root of $\operatorname{det} L(\tau, \xi)=0$ is independent of $\xi \in$ $\mathbb{R}^{d} \backslash(\mathcal{B} \cup\{0\})$.
iv. If $\lambda(\xi) \in C^{\omega}\left(\Omega_{g}\right)$ is an eigenvalue of $\sum A_{j} \xi_{j}$ depending real analytically on $\xi$, then either there is $j \in\{1, \ldots, M\}$ such that $\lambda(\xi)=-\mathbf{v}_{j} \cdot \xi$ or $\nabla^{2} \lambda \neq 0$ almost everywhere on $\Omega_{g}$.

Proof. i. Use the basic stratification theorem of real algebraic geometry (see [BR], [CR]). The characteristic variety is a conic real algebraic variety in $\mathbb{R}^{1+d} \backslash\{0\}$.
Over each $\xi$ it contains at least 1 and at most $N$ points. Therefore its projection on $\mathbb{R}_{\xi}^{d}$ is the whole space so the variety has dimension at least $d$. On the other hand it has measure zero by Fubini's Theorem so the dimension is at most $d$, since $d+1$ dimensional algebraic sets contain open sets.
The singular points are therefore a stratum of dimension at most $d-1$. The bad frequencies are exactly the projection of this singular locus and so is a real algebraic subvariety of $\mathbb{R}_{\xi}^{d}$ of dimension at most $d-1$ and $\mathbf{i}$ follows.
ii. That there are at most a finite number of components in the complementary set is a classical theorem of Whitney (see [BR], [CR]).
iii. Denote by $m$ the mulitplicity on $\Omega_{g}$ and $m^{\prime}$ the mulitplicity on $\Omega_{g^{\prime}}$. by definition of mulitplicity,

$$
\begin{equation*}
\xi \in \Omega_{g} \quad \text { and } \quad k<\left.m \quad \Longrightarrow \quad \frac{\partial^{k} \operatorname{det} L(\tau, \xi)}{\partial \tau^{k}}\right|_{\tau=-\mathbf{v}_{j} . \xi}=0 \tag{3.2.2}
\end{equation*}
$$

Then $\partial_{\tau}^{k} L\left(-\mathbf{v}_{j} \cdot \xi, \xi\right)$ is a polynomial in $\xi$ which vanishes on the nonempty open set $\Omega_{g}$, so must vanish identically. Thus it vanishes on $\Omega_{g^{\prime}}$ and it follows that $m^{\prime} \geq m$. by symmetry one has $m \geq m^{\prime}$.
iv. If $\lambda$ is a linear function $\lambda=-\mathbf{v} . \xi$ on $\Omega_{g}$, then $\operatorname{det} L(-\mathbf{v} \cdot \xi, \xi)=0$ for $\xi \in \Omega_{g}$ so by analytic continuation, must vanish for all $\xi$. It follows that the hyperplane $\tau=-\mathbf{v} . \xi$ lies in the characteristic variety and therefore that $\lambda=-\mathbf{v}_{j} . \xi$ for some $j$.
If $\lambda$ is not a linear function, then the matrix $\nabla_{\xi}^{2} \lambda$ is a real analytic function on $\Omega_{g}$ which is not identically zero and therefore vanishes at most on a set of measure zero in $\Omega_{g}$.

Definitions. Enumerate the roots of $\operatorname{det} L(\tau, \xi)=0$ as follows. Let $\mathcal{A}_{f}:=\{1, \ldots, M\}$ denote the indices of the flat parts, and for $\alpha \in \mathcal{A}_{f}, \tau_{\alpha}(\xi):=-\mathbf{v}_{\alpha} \cdot \xi$. For $g \in \mathcal{G}$ and $\xi \in \Omega_{g}$, number the roots other than the $\left\{\tau_{\alpha}: \alpha \in \mathcal{A}_{f}\right\}$ in order $\tau_{g, 1}(\xi)<\tau_{g, 2}(\xi)<\cdots<\tau_{g, M(g)}$. Multiple roots are not repeated in this list. Let $\mathcal{A}_{c}$ denote the indices of the curved sheets

$$
\begin{equation*}
\mathcal{A}_{c}:=\{(g, j): g \in \mathcal{G} \text { and } 1 \leq j \leq M(g)\} \tag{3.2.3}
\end{equation*}
$$

Let $\mathcal{A}:=\mathcal{A}_{f} \cup \mathcal{A}_{c}$. For $\alpha \in \mathcal{A}_{f}$ and $\xi \in \mathbb{R}^{d}$ define $E_{\alpha}(\xi):=\pi\left(-\mathbf{v}_{j} . \xi, \xi\right)$. For $\alpha \in \mathcal{A}_{c}$ define

$$
E_{\alpha}(\xi):=\left\{\begin{array}{cc}
\pi\left(\tau_{\alpha}(\xi), \xi\right) & \text { for } \quad \xi \in \Omega_{g}  \tag{3.2.4}\\
0 & \text { for } \quad \xi \notin \Omega_{g}
\end{array}\right.
$$

The next proposition decomposes an arbitrary solution of $L u=0$ as a finite sum of simpler waves.

Proposition 3.2.2. 1. For each $\alpha \in \mathcal{A}, E_{\alpha}(\xi) \in C^{\omega}\left(\mathbb{R}^{d} \backslash(\mathcal{B} \cup\{0\})\right)$ is an orthogonal projection valued function positive homogeneous of degree zero.
2. For each $\xi \in \mathbb{R}^{d} \backslash(\mathcal{B} \cup\{0\}), \mathbb{C}^{N}$ is equal to the orthogonal direct sum

$$
\begin{equation*}
\mathbb{C}^{N}=\oplus_{\alpha \in \mathcal{A}} \text { Image } E_{\alpha}(\xi) \tag{3.2.5}
\end{equation*}
$$

3. The operators $E_{\alpha}\left(D_{x}\right):=\mathcal{F}^{*} E(\xi) \mathcal{F}$ are orthogonal projectors on $H^{s}\left(\mathbb{R}^{d}\right)$, and for each $s \in \mathbb{R}, H^{s}\left(\mathbb{R}^{d}\right)$ is equal to the orthogonal direct sum,

$$
\begin{equation*}
H^{s}\left(\mathbb{R}^{d}\right)=\oplus_{\alpha \in \mathcal{A}} \text { Image } E_{\alpha}\left(D_{x}\right) \tag{3.2.6}
\end{equation*}
$$

4. If $f \in \mathcal{S}^{\prime}\left(\mathbb{R}^{d}\right)$ has Fourier transform equal to a locally integrable function, then the solution of the initial value problem

$$
\begin{equation*}
L\left(\partial_{y}\right) u=0,\left.\quad u\right|_{t=0}=f \tag{3.2.7}
\end{equation*}
$$

is given by the formula

$$
\begin{equation*}
\hat{u}(t, \xi)=\sum_{\alpha \in \mathcal{A}} \hat{u}_{\alpha}(t, \xi):=\sum_{\alpha \in \mathcal{A}} e^{i t \tau_{\alpha}(\xi)} E_{\alpha}(\xi) \hat{f}(\xi) \tag{3.2.8}
\end{equation*}
$$

Remarks. 1. The last decomposition is also written

$$
u:=\sum_{\alpha \in \mathcal{A}} u_{\alpha}:=\sum_{\alpha \in \mathcal{A}} e^{i t \tau_{\alpha}\left(D_{x}\right)} E_{\alpha}\left(D_{x}\right) f .
$$

2. Since $\tau_{\alpha}$ is real valued on the support of $E_{\alpha}(\xi)$ the operator $e^{i t \tau_{\alpha}\left(D_{x}\right)} E_{\alpha}\left(D_{x}\right)$ is a contraction on $H^{s}\left(\mathbb{R}^{d}\right)$ for all $s$.
3. If $\alpha \in \mathcal{A}_{f}$ then $-i \tau_{\alpha}\left(D_{x}\right)=\mathbf{v}_{\alpha} . \partial_{x}$. For $\alpha=(g, j) \in \mathcal{A}_{c},\left|\tau_{\alpha}(\xi)\right| \leq C|\xi|$, so the operator $\tau_{\alpha}\left(D_{x}\right) f$ is continuous from $H^{s}$ to $H^{s-1}$. The mode $u_{\alpha}=e^{i t \tau_{\alpha}\left(D_{x}\right)} E_{\alpha}\left(D_{x}\right) f$ satisfies $\partial_{t} u_{\alpha}=i \tau_{\alpha}\left(D_{x}\right) u_{\alpha}$. For $\alpha \in \mathcal{A}_{f}$ this is $\left(\partial_{t}+\mathbf{v}_{\alpha} . \partial_{x}\right) u_{\alpha}=0$, so

$$
u_{\alpha}=\left(E_{\alpha}(D) f\right)\left(x-\mathbf{v}_{\alpha} t\right)
$$

4. Over $\mathcal{B}(L)$ only the $E_{\alpha}$ corresponding to the hyperplanes are defined. One does not have a decomposition of $\mathbb{C}^{N}$. It is important that $\mathcal{B}$ is a negligible set for $\hat{f}$. The $\hat{f} \in L_{\text {loc }}^{1}$ assumption in 4 is essential.

## §3.3. Large time asymptotics.

Definition. Define $\mathbb{A}$ as the set of tempered distributions whose Fourier transforms belong to $L^{1}\left(\mathbb{R}^{d}\right)$. Then $\mathbb{A}$ is a Banach space with norm

$$
\begin{equation*}
\|f\|_{\mathbb{A}}:=(2 \pi)^{-d / 2} \int_{\mathbb{R}^{d}}|\hat{f}(\xi)| d \xi . \tag{3.3.2}
\end{equation*}
$$

The Fourier Inversion Formula implies that $\mathbb{A} \subset L^{\infty}\left(\mathbb{R}^{d}\right)$ and

$$
\begin{equation*}
\|f\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \leq\|f\|_{\mathbb{A}} \tag{3.3.3}
\end{equation*}
$$

The elements of $\mathbb{A}$ are continuous and tend to zero as $x \rightarrow \infty$. Moreover, the Fourier transform of $f^{2}$ is a multiple of $\hat{f} * \hat{f}$ and therefore in $L^{1}$, so $\mathbb{A}$ is an algebra. It is called the Wiener algebra. It was a centerpiece of the Tauberian Theorems of N. Wiener.

Theorem 3.3.1 ( $L^{\infty}$ asymptotics for symmetric systems). Suppose that $f \in \mathbb{A}$ and $u$ is the solution of the initial value problem $L\left(\partial_{x}\right) u=0,\left.u\right|_{t=0}=f$. Then with the notation introduced in the preceding section,

$$
\begin{equation*}
\lim _{t \rightarrow \infty}\left\|u(t)-\sum_{\alpha \in \mathcal{A}_{f}}\left(E_{\alpha}\left(D_{x}\right) f\right)\left(x-\mathbf{v}_{\alpha} t\right)\right\|_{L^{\infty}\left(\mathbb{R}^{d}\right)}=0 \tag{3.3.1}
\end{equation*}
$$

Remarks. 1. This result shows that a general solution of the Cauchy problem is the sum of $M$ rigidly translating waves, one for each hyperplane in the characteristic variety, plus a term which tends to zero in sup norm. The last part decays because of the dispersion of waves.
2. The Theorem does not extend to $f$ whose Fourier Transform is a bounded measure. For example, $u:=\left(e^{i\left(x_{1}-t\right)}, 0\right)$ is a solution of (3.1.3) with $\hat{f}$ equal to a point mass. The characteristic variety contains no hyperplanes so (3.3.1) asserts that solutions with $\hat{f} \in L^{1}$ tend to zero in $L^{\infty}\left(\mathbb{R}^{d}\right)$ while $u(t)$ has sup norm equal to 1 for all $t$.

Proof of Theorem. Step 1. Approximation-decomposition. Symmetric hyperbolicity implies that for each $t, \xi, \exp \left(i t \sum A_{j} \xi_{j}\right)$ is unitary on $\mathbb{C}^{N}$. Therefore $S(t):=$ $\exp \left(-t \sum_{j} A_{j} \partial_{j}\right)$ is isometric on $\mathbb{A}$. Since the family of linear maps

$$
f \longmapsto S(t) f-\sum_{\alpha \in \mathcal{A}_{f}}\left(E_{\alpha}\left(D_{x}\right) f\right)\left(x-\mathbf{v}_{\alpha} t\right)
$$

is uniformly bounded from $\mathbb{A}$ to $L^{\infty}\left(\mathbb{R}^{d}\right)$, it suffices to prove (3.3.1) for a set of $f$ dense in A.

For $\alpha \in \mathcal{A}_{c}$, Propostion 3.2.1.iv shows that the matrix of second derivatives, $\nabla_{\xi}^{2} \tau_{\alpha}$ can vanish at most on a closed set of measure zero. The set of $f$ we choose is those with

$$
\hat{f} \in C_{0}^{\infty}\left(\mathbb{R}^{d} \backslash\left\{\mathcal{B} \cup\{0\} \cup \bigcup_{\alpha \in \mathcal{A}_{c}}\left\{\xi \in \Omega_{g}: \nabla_{\xi}^{2} \tau_{\alpha}(\xi)=0\right\}\right\}\right)
$$

Since the removed set is a closed null set, these $f$ are dense.
To prove (3.3.1) for such $f$ decompose

$$
\begin{equation*}
f=\sum_{\alpha \in \mathcal{A}} f_{\alpha}:=\sum_{\alpha \in \mathcal{A}} E_{\alpha}\left(D_{x}\right) f, \quad u(t)=S(t) f=\sum u_{\alpha}(t):=\sum S(t) f_{\alpha} \tag{3.3.4}
\end{equation*}
$$

For $\alpha \in \mathcal{A}_{f}, u_{\alpha}(t)=\left(E_{\alpha}\left(D_{x}\right) f\right)\left(x-\mathbf{v}_{\alpha} t\right)$ which recovers the summands in (3.15). To prove (3.15) it suffices to show that for $\alpha \in \mathcal{A}_{c}$

$$
\begin{equation*}
\lim _{t \rightarrow \infty}\left\|u_{\alpha}(t)\right\|_{L^{\infty}\left(\mathbb{R}^{d}\right)}=0 \tag{3.3.5}
\end{equation*}
$$

Step 2. Stationary and nonstationary phase. Part 4 of Proposition 3.2. shows that for $\alpha \in \mathcal{A}_{c}$,

$$
\begin{equation*}
u_{\alpha}(t, x)=\int_{\Omega_{g}} e^{i\left(\tau_{\alpha}(\xi) t+x \cdot \xi\right)} \hat{f}_{\alpha}(\xi) d \xi, \quad \hat{f}_{\alpha} \in C_{0}^{\infty}\left(\Omega_{g}\right) \tag{3.3.6}
\end{equation*}
$$

For each $\xi$ in the support of $\hat{f}_{\alpha}$, there is a vector $\mathbf{r} \in \mathbb{R}^{d}$ so that $\left\langle\nabla_{\xi}^{2} \tau(\xi) \mathbf{r}, \mathbf{r}\right\rangle \neq 0$ on a neighborhood of $\xi$. Using a partition of unity we can write $\hat{f}_{\alpha}$ as a finite sum of functions
$h_{\mu} \in C_{0}^{\infty}\left(\Omega_{g}\right)$ so that for each $\mu$ there is a $\mathbf{r}_{\mu} \in \mathbb{C}^{N}$ so that on an open ball containing the support of $h_{\mu},\left\langle\nabla_{\xi}^{2} \tau(\xi) \mathbf{r}_{\mu}, \mathbf{r}_{\mu}\right\rangle \neq 0$. It suffices to show that for each $\mu$

$$
\begin{equation*}
\lim _{t \rightarrow \infty} \sup _{x \in \mathbb{R}^{d}} \int e^{i\left(\tau_{\alpha}(\xi) t+x \cdot \xi\right)} h_{\mu}(\xi) d \xi=0 \tag{3.3.7}
\end{equation*}
$$

For ease of reading we suppress the subscripts. Write $x=t z$. For each $t>0$, the sup in $x$ is equal to the sup in $z$ so it suffices to show that

$$
\lim _{t \rightarrow \infty} \sup _{z \in \mathbb{R}^{d}}\left|\int e^{i t(\tau(\xi)+z . \xi)} h(\xi) d \xi\right|=0
$$

Choose

$$
\sigma>\sup _{\xi \in \operatorname{supp} h}\left|\nabla_{\xi} \tau(\xi)\right|
$$

There is a $\delta>0$ so that for all $|z| \geq \sigma$,

$$
\left|\nabla_{\xi}(\tau(\xi)+z . \xi)\right| \geq \delta
$$

The method of nonstationary phase implies that

$$
\forall N>0, \exists C_{N}, \forall|z| \geq \sigma, t>1, \quad\left|\int e^{i t(\tau(\xi)+z \cdot \xi)} h(\xi) d \xi\right| \leq C_{N} t^{-N}
$$

It remains to show that

$$
\begin{equation*}
\lim _{t \rightarrow \infty} \sup _{|z| \leq \sigma}\left|\int e^{i t(\tau(\xi)+z \cdot \xi)} h(\xi) d \xi\right|=0 \tag{3.3.8}
\end{equation*}
$$

Make a linear change of variables in $\xi$ so that $\mathbf{r}=(1,0, \ldots, 0)$ and therefore

$$
\frac{\partial^{2} \tau}{\partial^{2} \xi_{1}} \neq 0, \quad \text { on } \quad \operatorname{supp} h
$$

Choose $R>0$ so that for $\xi \in \operatorname{supp} h,|\xi| \leq R$. Set

$$
\Gamma:=\left\{\left|z_{1}\right| \leq \sigma\right\} \times\left\{\left|\xi_{2}, \ldots, \xi_{d}\right| \leq R\right\} \subset \mathbb{R}^{1} \times \mathbb{R}^{d-1}
$$

Define

$$
\begin{aligned}
K(t) & :=\sup _{|z| \leq \sigma,\left|\xi_{2}, \ldots, \xi_{d}\right| \leq R}\left|\int e^{i t\left(\tau(\xi)+z_{1} \cdot \xi_{1}\right)} h(\xi) d \xi_{1}\right| \\
& =\sup _{\Gamma}\left|\int e^{i t\left(\tau(\xi)+z_{1} \cdot \xi_{1}\right)} h(\xi) d \xi_{1}\right| .
\end{aligned}
$$

Then

$$
\begin{aligned}
& \sup _{|z| \leq \sigma}\left|\int e^{i t(\tau(\xi)+z . \xi)} h(\xi) d \xi\right| \\
& \leq \int_{\left|\xi_{2}, \ldots, \xi_{d}\right| \leq R} e^{i\left(z_{2} \xi_{2}+\ldots+z_{d} \cdot \xi_{d}\right)}\left(\int e^{i t\left(\tau+z_{1} \xi_{1}\right)} h(\xi) d \xi_{1}\right) d \xi_{2} \ldots d \xi_{d} \\
& \leq\left|\left\{\left|\xi_{2}, \ldots, \xi_{d}\right| \leq R\right\}\right| K(t)
\end{aligned}
$$

It therefore suffices to show that

$$
\begin{equation*}
\lim _{t \rightarrow \infty} K(t)=0 \tag{3.3.9}
\end{equation*}
$$

The points of $\Gamma$ are split according to whether the phase $\tau(\xi)+z_{1} \xi_{1}$ has a stationary point with respect to $\xi_{1}$ or not. If $\underline{\gamma} \in \Gamma$ is such that

$$
\left|\frac{\partial \tau}{\partial \xi_{1}}+z_{1}\right|>\delta>0 \quad \text { for all } \quad\left|z_{1}\right| \leq \sigma,|\xi| \leq R
$$

the same is true on a neighborhood of $\underline{\gamma}$. The principal of nonstationary phase shows that

$$
\int e^{i t\left(\tau_{\alpha}(\xi)+z \cdot \xi\right)} \hat{h}_{\mu}(\xi) d \xi_{1}=O\left(t^{-N}\right)
$$

uniformly on such a neighborhood.
On the other hand if for $\underline{\gamma}$ there is a stationary point, then the strict convexity of $\tau$ in $\xi_{1}$ shows that it is unique and nondegenerate. Therefore for nearby $\gamma$ there is a nearby unique and nondegenerate stationary point. The inequality of stationary phase (see Appendix) implies that

$$
\int e^{i t\left(\tau_{\alpha}(\xi)+z . \xi\right)} \hat{h}_{\mu}(\xi) d \xi_{1}=O\left(t^{-1 / 2}\right)
$$

uniformly on a neighborhood of $\underline{\gamma}$.
Covering the compact set $\Gamma$ by a finite family of neighborhoods proves (3.39) and therefore the Theorem.

Definition. The operator $L$ purely dispersive when its characteristic variety contains no hyperplanes. It is call nondispersive when its characteristic variety is equal to a union of hyperplanes.

The nondispersive operators have a discrete set of group velocities. The characteristic variety of purely dispersive operators have only curved sheets. The latter name is justified by the next Corollary.

Corollary 3.3.2. If $L=L_{1}\left(\partial_{x}\right)$ is a constant coefficient homogeneous symmetric hyperbolic operator, then the following are equivalent.

1. The characteristic variety of $L$ contains no hyperplanes (i.e. $L$ is purely dispersive).
2. Every solution of $L u=0$ with $\left.u\right|_{t=0} \in C_{0}^{\infty}\left(\mathbb{R}^{d}\right)$ satisfies,

$$
\begin{equation*}
\lim _{t \rightarrow \infty}\|u(t)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \rightarrow 0 \tag{3.3.10}
\end{equation*}
$$

3. Every solution of $L u=0$ with $\left.u\right|_{t=0} \in \mathbb{A}$ satisfies (3.3.10).
4. If $\tau(\xi)$ is a $C^{\infty}$ solution of $\operatorname{det} L(\tau, \xi)=0$ defined on a open set of $\xi \in \mathbb{R}^{d}$ then for every $\mathbf{v} \in \mathbb{R}^{d},\left\{\xi \in \mathbb{R}^{d}: \nabla_{\xi} \tau=-\mathbf{v}\right\}$ has measure zero.

Proof. Theorem 3.3 shows that $\mathbf{1} \Leftrightarrow \mathbf{3}$. To complete the proof we show that $\mathbf{3} \Leftrightarrow \mathbf{2}$ and $1 \Leftrightarrow 4$.

The assertions $\mathbf{2}$ and $\mathbf{3}$ are equivalent because the family of mappings $u(0) \mapsto u(t)$ is uniformly bounded from $\mathbb{A} \rightarrow L^{\infty}$, and $C_{0}^{\infty}$ is dense in $\mathbb{A}$.
That $\sim 1 \Longrightarrow \sim 4$ is immediate.
If $\mathbf{4}$ is violated there is a smooth solution $\tau$ so that $\nabla_{\xi} \tau=-\mathbf{v}$ on a set of positive measure. It follows from the Fundamental Stratification Theorem (see [BR], [CR]) that $\nabla_{\xi} \tau=-\mathbf{v}$ on a conic open real algebraic set of dimension $d$ in $\mathbb{R}^{d} \backslash 0$. Then $\tau=-\mathbf{v} . \xi$ on this set and we conclude that the polynomial $\operatorname{det} L(-\mathbf{v} \cdot \xi, \xi)$ vanishes on this set and therefore everywhere. Thus the hyperplane $\{\tau=-\mathbf{v} . \xi\}$ is contained in the characteristic variety and 1 is violated.

Thus $\mathbf{1}$ and $\mathbf{4}$ are equivalent.
Remark. Part four of this Corollary shows that for any velocity $\mathbf{v}$ the group velocity $-\nabla_{\xi} \tau$ associated to a curved sheet of the characteristic variety takes the value $\mathbf{v}$ for at most a set of frequencies $\xi$ of measure zero.

The nondispersive evolutions are described in the next results.
Corollary 3.3.3. If $L=L_{1}\left(\partial_{y}\right)$ is a constant coefficient homogeneous symmetric hyperbolic operator with $A_{0}=I$, then the following are equivalent.

1. The characteristic variety of $L$ is a finite union of hyperplanes.
2. (Motzkin and Tausky) The matrices $A_{j}$ commute.
3. If $u$ satisfies $L u=0$ with $u(0) \in \mathbb{A}$ and $\|u(t)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \rightarrow 0$ as $t \rightarrow \infty$, then $u$ is identically equal to zero.

Proof. $2 \Rightarrow 3$. A unitary change of variable $u=V v$ replaces the equation $L u=0$ with the equivalent equation $\tilde{L} v=0$ with $\tilde{A}_{j}:=V^{*} A_{j} V$. When the $A_{j}$ commute, $V$ can be chosen so that the $\tilde{A}_{j}$ are all real diagonal matrices. Property $\mathbf{3}$ is clear for the tilde equation as each component of the solution rigidly translates as time goes on. The only way its supremum can tend to zero at $t \rightarrow \infty$ is for it to vanish.
$\mathbf{3} \Rightarrow \mathbf{1}$. This is an immediate consequence of Theorem 3.3.
$\mathbf{1} \Rightarrow \mathbf{2}$. This is a result of Motzkin and Tausky.

Theorem 3.3.4. (Motzkin and Tausky) Suppose that $A$ and $B$ are hermitian $N \times N$ matrices. The eigenvalues of $\xi A+\eta B$ are linear functions of $\xi, \eta$ if and only if $A$ and $B$ commute.

Proof. We must show that linear eigenvalue implies commutation. The proof is by induction on $N$. The case $N=1$ is trivial. We suppose that $N>1$ and the result is known for dimensions $\leq N-1$.
Consider the characteristic variety $\operatorname{det}(\tau+\xi A+\eta B)=0$. Choose a good point $(\underline{\xi}, \underline{\eta})$ so that above this point the variety has $k \leq N$ real analytic sheets. If $\eta=0$, leave the spatial coordinates as they are. If $\eta \neq 0$, change orthogonal coordinates in $\mathbb{R}^{2}$ so that $(\underline{\xi}, \underline{\eta})$ is a mutiple of $d y_{1}$. In this way we can without loss of generality assume that above $\eta=0$ the variety consists of $k$ real analytic sheets.
For $s$ small the eigenvalues of $A+s B$ are real analytic function $\lambda_{j}(s)$ with $\lambda_{j}(0)<\lambda_{j+1}(0)$ for $1 \leq j<k-1$. Denote by $\mu_{j}$ the multiplicity of $\lambda_{j}(0)$ and therefore of $\lambda_{j}(s)$ for $s$ small. By hypothesis the $\lambda_{j}(s)$ are affine functions of $s$ so $\lambda^{\prime \prime}=0$. We use this only at $s=0$.
By a unitary change of variable in $\mathbb{C}^{N}$ one can arrange that $A$ is block diagonal with diagonal entries $\lambda_{j}(0) I_{\mu_{j} \times \mu_{j}}$.
Corresponding to this block structure and the eigenvalue $\lambda_{1}$, one has one has

$$
\begin{align*}
\pi & =\operatorname{diag}\left(I_{\mu_{1} \times \mu_{1}}, 0_{\mu_{2} \times \mu_{2}}, \ldots, 0_{\mu_{k} \times \mu_{k}}\right) \\
Q & =\operatorname{diag}\left(0_{\mu_{1} \times \mu_{1}}, \frac{1}{\lambda_{2}-\lambda_{1}} I_{\mu_{2} \times \mu_{2}}, \ldots, \frac{1}{\lambda_{k}-\lambda_{1}} I_{\mu_{k} \times \mu_{k}}\right) . \tag{3.3.11}
\end{align*}
$$

The matrix $B$ has block structure

$$
B=\left(\begin{array}{cccc}
B_{1,1} & B_{1,2} & \ldots \ldots & B_{1, k} \\
B_{2,1} & B_{2,2} & \ldots \ldots & B_{2, k} \\
& & & \\
B_{k, 1} & B_{k, 2} & \ldots & B_{k, k}
\end{array}\right)
$$

with $B_{i j}$ a $\mu_{i} \times \mu_{j}$ matrix and $B_{i j}=B_{j i}^{*}$.
The fundamental formula of second order perturbation theory (see Appendix) yields $\lambda^{\prime \prime} \pi=$ $2 \pi B Q B \pi$. By hypothesis this is equal to zero.
Straightforward calculation shows that

$$
\pi B=\left(\begin{array}{cccc}
B_{1,1} & B_{1,2} & \ldots . . & B_{1, k} \\
0 & 0 & \ldots . & 0 \\
0 & 0 & & 0 \\
0 & 0 & \ldots . & 0
\end{array}\right), \quad Q B \pi=\left(\begin{array}{cccc}
0 & 0 & \ldots . & 0 \\
\frac{1}{\lambda_{2}-\lambda_{1}} B_{2,1} & 0 & \ldots . & 0 \\
\frac{1}{\lambda_{k}-\lambda_{1}} B_{k, 1} & 0 & \ldots . & 0
\end{array}\right)
$$

Therefore, the $\mu_{1} \times \mu_{1}$ upper left hand block block of $\pi Q B Q \pi$ is equal to

$$
\sum_{j=2}^{k} \frac{1}{\lambda_{j}-\lambda_{1}} B_{1, j} B_{1, j}^{*}
$$

Conclude that this sum of positive square matrices vanishes. Thus, for $j \geq 2, B_{1, j}=0$ and $B_{j, 1}=0$.
Thus $B$ and $A$ are reduced by the splitting

$$
\mathbb{C}^{N}=\mathbb{C}^{\mu_{1}} \times \mathbb{C}^{N-\mu_{1}}
$$

The commutation then follows by the inductive hypothesis applied to the diagonal blocks. This proves the case $N$ and completes the induction.

This completes the proof of Corollary 3.3.3.
Example. The implication $\mathbf{1} \Rightarrow \mathbf{2}$ is not true without the symmetry hypothesis. For example, the hypberbolic system

$$
\partial_{t}+\left(\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right) \partial_{1}+\left(\begin{array}{ll}
1 & 1 \\
0 & 1
\end{array}\right) \partial_{2}
$$

has flat characteristic variety with equation

$$
\left(\tau+\xi_{1}+\xi_{2}\right)\left(\tau-\xi_{1}+\xi_{2}\right)=0
$$

and the coefficient matrices do not commute. The conclusion is correct assuming that the hyperbolic system generates a semigroup in $L^{2}\left(\mathbb{R}^{d}\right)$ (see [GR, Hyperbolic multipliers are translations]).

Theorem (P. Brenner). If $L=L\left(\partial_{y}\right)$ is a constant coefficient homogeneous symmetric hyperbolic operator then the conditions of Corollary 3.3.3 are equivalent to each of the following.
i. For all $t \in \mathbb{R}$ and $p \in[1, \infty]$ the Fourier multiplication operator

$$
S(t):=\mathcal{F}^{-1} e^{-i t \sum A_{j} \xi_{j}} \mathcal{F}
$$

is a bounded from $L^{p}\left(\mathbb{R}^{d}\right)$ to itself.
ii. For some $\underline{t} \in \mathbb{R} \backslash 0$ and $2 \neq \underline{p} \in[1, \infty]$ the operator $S(\underline{t})$ is bounded from $L^{\underline{p}}\left(\mathbb{R}^{d}\right)$ to itself.

Remark. The Fourier multiplication operators are unitary on $L^{2}$. The properties ii means that the restriction to $\mathcal{S}(\mathbb{R})$ extend to bounded operators on $L^{p}$, equivalently

$$
\sup _{f \in \mathcal{S}\left(\mathbb{R}^{d}\right) \backslash 0} \frac{\|S(t) f\|_{L^{p}\left(\mathbb{R}^{d}\right)}}{\|f\|_{L^{p}\left(\mathbb{R}^{d}\right)}}<\infty .
$$

Proof. The conditions of Corollary 3.3.3 implies that after an orthogonal change of basis, the $A_{j}$ are all real diagonal matrices. It is then elementary to verify that $\mathbf{i}$ is satisfied.

Clearly $\mathbf{i}$ implies ii. It remains to show that ii implies the conditions of the Corollary. If the conditions of the the Corollary are violated, then $\mathbf{i i}$ is violated. The first remark is that ii is stronger than it appears. Since $S(\underline{t})$ is unitary on $L^{2}$, if ii is satisfied then $S(t)$ is bounded on $L^{p}$ for all $p$ between 2 and $\underline{p}$. Thus we may assume that $\underline{p}$ in not equal to 1 or $\infty$.
For $\sigma \in \mathbb{R} \backslash 0, L u=0$ if and only if $u^{\sigma}(t, x):=u(\sigma t, \sigma x)$ satisfies $L u^{\sigma}=0$. It follows that if ii is satisfied then

$$
\begin{equation*}
\|S(t)\|_{\operatorname{Hom}(L \underline{\underline{p}})}=\|S(\underline{t})\|_{\operatorname{Hom}(L \underline{\underline{p}})}<\infty, \quad \forall t \neq 0 \tag{3.3.12}
\end{equation*}
$$

If $\underline{q}$ is the conjugate index to $\underline{p}$, that is $\frac{1}{\underline{p}}+\frac{1}{\underline{q}}=1$, then

$$
\begin{aligned}
\|S(t)\|_{\operatorname{Hom}\left(L^{\underline{q}}\left(\mathbb{R}^{d}\right)\right)} & =\sup _{f, g \in \mathcal{S} \backslash 0} \frac{(S(t) f, g)}{} \frac{\|f\|_{L^{\underline{q}}\left(\mathbb{R}^{d}\right)}\|g\|_{L^{\underline{p}}\left(\mathbb{R}^{d}\right)}}{} \\
& =\sup _{f, g \in \mathcal{S} \backslash 0} \frac{(f, S(-t) g)}{\|f\|_{L^{\underline{q}}\left(\mathbb{R}^{d}\right)}\|g\|_{L^{\underline{p}}\left(\mathbb{R}^{d}\right)}}=\|S(-t)\|_{\operatorname{Hom}\left(L^{\underline{p}}\left(\mathbb{R}^{d}\right)\right.} .
\end{aligned}
$$

Thus when ii is satisfied for $\underline{p}$ it is satisfied for $\underline{q}$ so we may suppose that $\underline{p}>2$.
When the conditions of Corollary 3.3.3 are violated, there is a conic set of good points $\Omega_{g}$ and a sheet $\tau=\tau(\xi)$ over $\Omega_{g}$ with $\nabla_{\xi \xi}^{2} \tau \neq 0$ for almost all $\xi \in \Omega_{g}$. Denote by $\pi(\xi)$ with associated spectral projection. Choose an $f \in \mathcal{S}\left(\mathbb{R}^{d}\right)$ with $\hat{f}$ compactly supported in $\Omega_{g}$. Replacing $\hat{f}$ by $\pi(\xi) \hat{f}$ we may assume that $\pi(D) f=f$. Theorem 3.3.1 implies that

$$
\lim _{t \rightarrow \infty}\|S(t) f\|_{L^{\infty}\left(\mathbb{R}^{d}\right)}=0
$$

Then

$$
\|S(t) f\|_{L^{\underline{p}}\left(\mathbb{R}^{d}\right)}^{\underline{p}} \leq\|S(t) f\|_{L^{\infty}\left(\mathbb{R}^{d}\right)}^{\frac{p-2}{}}\|S(t) f\|_{L^{2}\left(\mathbb{R}^{d}\right)}^{2}=\|S(t) f\|_{L^{\infty}\left(\mathbb{R}^{d}\right)}^{\underline{p-2}}\|f\|_{L^{2}\left(\mathbb{R}^{d}\right)}^{2} \rightarrow 0
$$

as $t \rightarrow \infty$.
Therefore,

$$
\|S(-t)\|_{\operatorname{Hom}\left(L^{\underline{p}}\right)} \geq \frac{\|S(-t)(S(t) f)\|_{L^{\underline{p}}}}{\|S(t) f\|_{L^{\underline{p}}}}=\frac{\|f\|_{L^{\underline{p}}}}{\|S(t) f\|_{L^{\underline{p}}}} \rightarrow \infty
$$

Thus (3.3.12) is violated and the proof is complete.
Example. It may seem that (3.3.12) together with $\lim _{t \rightarrow 0} S(t) f=f$ might imply that $S(t)$ has norm equal to 1 . That this is not true can be seen from the one dimensional example

$$
\partial_{t}+\left(\begin{array}{ll}
0 & 0 \\
0 & 1
\end{array}\right) \partial_{x}
$$

and $L^{p}$ norm chosen so that for $p=2$ one has unitarity,

$$
\left.\left\|\left(u_{1}, u_{2}\right)\right\|_{p}:=\int\left\|\left(u_{1}, u_{2}\right)\right\|^{p} d x\right)^{1 / p}, \quad\left\|\left(u_{1}, u_{2}\right)\right\|:=\left(\left|u_{1}\right|^{2}+\left|u_{2}\right|^{2}\right)^{1 / 2}
$$

Choosing $u_{1}(0)=u_{2}(0)=f \in C_{0}^{\infty}(\{|x| \leq \rho\})$ one has

$$
\|u(0)\|_{p}^{p}=(\sqrt{2})^{p}\|f\|_{p}^{p}
$$

and for $|t|>\rho$,

$$
\|u(t)\|_{p}^{p}=2\|f\|_{p}^{p}
$$

It follows that for all $t \neq 0$ and $p<2,\|S(t)\|_{\operatorname{Hom}\left(L^{p}\right)}^{p} \geq 2^{1-p / 2}>1$. Reversing time, treats $p>2$.

## §3.4. Maximally dispersive systems.

## $\S$ 3.4.1. The $L^{1} \rightarrow L^{\infty}$ decay estimate.

If $\tau=\tau(\xi)$ parametrizes a real analytic patch of the characteristic variety of a hyperbolic operator then $\tau$ is homogeneous of degree 1 in $\xi$. The group velocity $\mathbf{v}(\xi)=-\nabla_{\xi} \tau(\xi)$ is homogeneous of degree 0 . Therefore $\xi . \nabla_{\xi} \mathbf{v}=0$ so $\xi$ belongs to the kernel of the symmetric matrix $\nabla_{\xi} \mathbf{v}(\xi)=\nabla_{\xi}^{2} \tau(\xi)$. Thus the rank of $\nabla_{\xi}^{2} \tau$ is at most $d-1$. When the rank is equal to $d-1$ the group velocity depends as strongly on $\xi$ as possible. The dispersion is as strong as possible.

Definition. The homogeneous constant coefficient symmetric hyperbolic operator is maxmally dispersive when

$$
\operatorname{Char} L=\cup_{j=1}^{m}\left\{(\tau, \xi): \tau=\tau_{j}(\xi)\right\}
$$

where for $\xi \in \mathbb{R}^{d} \backslash 0$

$$
\tau_{1}(\xi)<\tau_{2}(\xi) \ldots<\tau_{m}(\xi)
$$

the $\tau_{j}$ are real analytic, positive homogeneous of degree one in $\xi$, and

$$
\begin{equation*}
\forall j, \quad \forall \xi \in \mathbb{R}^{d} \backslash 0, \quad \operatorname{rank} \nabla_{\xi}^{2} \tau(\xi)=d-1 \tag{3.4.1}
\end{equation*}
$$

Examples. i. The simplest example is

$$
\left(\tau^{2}-|\xi|^{2}\right)\left(\tau^{2}-c^{2}|\xi|^{2}\right)=0, \quad 0<c \neq 1
$$

The variety in this case consists of two sheets $\tau=|\xi|$ and $\tau=c|\xi|$ which have $d-1$ strictly positive principal curvatures. The other sheets bound $\tau \leq-|\xi|$ and $\tau \leq-c|\xi|$ and have $d-1$ strictly negative curvatures.
ii. The next figure gives an example with two sheets bounding strictly convex regions for which the functions $\tau_{j}$ change sign. In particular the generator $G=-\sum A_{j} \partial_{j}$ is not elliptic since the points where the cone crosses $\tau=0$ are characteristic for $G$.


The next result is closely related to Hadamard's Ovaloid Theorem which is recalled in Appendix III.

Propostion 3.4.1. If $\tau(\xi)$ is smooth in $\xi \neq 0$, homogeneous of degree one and the hessian has rank equal to $d-1$ at all points, then the nonzero eigenvalues of $\nabla_{\xi}^{2} \tau$ have the same sign. When they are positive (resp. negative) $\tau$ is convex (resp. concave).

Proof. When $d=2, \nabla_{\xi}^{2} \tau$ has only one nonzero eigenvalue and the result is immediate. For $d \geq 3$, consider the mapping

$$
\Gamma(\xi):=\mathbf{v}(\xi)=-\nabla_{\xi} \tau(\xi) .
$$

The differential of the mapping $\Gamma$ is equal to $-\nabla_{\xi}^{2} \tau$ so $\xi$ is in its kernel and it is invertible when restricted to the orthogonal to $\xi$.
Since $\Gamma$ is homogeneous of degree 0 , it is natural to consider $\Gamma$ as a map from $S^{d-1}=$ $\{|\xi|=1\}$. As such it is an immersion onto a compact $d-1$ dimensional manifold, $\mathcal{M}$. The image is oriented by the image of the orientation of $S^{d-1}$.
Since $\xi$ is orthogonal to the image of $-\nabla_{\xi}^{2} \tau(\xi)$ it follows that the $\xi$ is the unit normal to $\mathcal{M}$ at $\Gamma(\xi)$. Thus, at least locally, $\Gamma$ is the inverse of the Gauss map of $\mathcal{M}$. Since the differential is invertible it follows that the Gauss curvature of $\mathcal{M}$ is nowhere vanishing.
Since $\xi \in \operatorname{ker}\left(\nabla_{\xi}^{2} \tau(\xi)\right.$, the unit normal to $\mathcal{M}$ at $\mathbf{v}(\xi)$ is equal to $\xi$. Since the map from $\xi \in S^{d-1}$ to $\mathbf{v}(\xi)$ has invertible jacobian, the Gauss curvature of $\mathcal{M}$ is nowhere vanishing. Since $d \geq 3$, it follows from Hadamard's Ovaloid Theorem, that $\mathcal{M}$ is the boundary of a strictly convex set and $\Gamma: S^{d-1} \rightarrow \mathcal{M}$ is a diffeomorphism.
Thus each value $-\nabla_{\xi} \tau(\xi) \in \mathcal{M}$ is attained at a unique $\xi \in S^{d-1}$.
The normals to $\tau=\tau(\xi)$ are the nonzero multiples of $(1, v(\xi))$. Thus, the hyperplane $\{\tau+v(\underline{\xi}) \cdot \xi=0\}$ is tangent at $\tau=\tau(\underline{\xi})$ and at no other point $\tau=\tau\left(\underline{\xi}^{\prime}\right)$ with $\underline{\xi}^{\prime} \in S^{d-1}$.
It follows that the cone $\tau=\tau(\xi)$ is strictly convex in the sense that its intersection with its tangent plane conists exactly of the line $(\mathbb{R} \backslash 0)(\tau(\xi), \xi)$.

This implies that the $d-1$ nonzero eigenvalues must have one sign.
Examples. The characteristic variety of a maximally dispersive system consists of $m$ disjoint sheets, each the boundary of a strictly convex cone.

Lemma 3.4.2 (Pointwise decay). If $d \geq 2, \tau$ is as above and $k \in C_{0}^{\infty}\left(\mathbb{R}^{d} \backslash 0\right)$ then there is a constant $C$ so that

$$
u(t, x):=\int e^{i t \tau(\xi)} e^{i x . \xi} k(\xi) d \xi
$$

satisfies

$$
\begin{equation*}
\|u(t)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \leq C(1+|t|)^{-(d-1) / 2} . \tag{3.4.2}
\end{equation*}
$$

Remark. This is the decay rate for solutions of $\square_{1+d} u=0$ which corresponds to the choice $\tau(\xi)= \pm|\xi|$.

Proof. The easy estimate

$$
\|u(t, x)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \leq \int|k(\xi)| d \xi
$$

shows that only the decay for $|t| \geq 1$ needs to be proved.
Let

$$
y:=\frac{x}{t}, \quad x=t y
$$

Then

$$
\sup _{x}|u(t, x)|=\sup _{y}|u(t, t y)|=\sup _{y}\left|\int e^{i t(\tau(\xi)+y \cdot \xi)} k(\xi) d \xi\right| .
$$

The phase $\tau(\xi)+y . \xi$ is stationary when

$$
-\nabla_{\xi} \tau(\xi)=y
$$

The left hand side is the group velocity.
As in Lemma 3.4.1, denote by $\mathcal{M}$ the set of attained group velocities which is an embedded strictly convex compact $d-1$ manifold.
For any open neighborhood $\mathcal{O}$ of $\mathcal{M}$, the method of nonstationary phase shows that for any $N$,

$$
\sup _{y \in \mathbb{R}^{d} \backslash \mathcal{O}}\left|\int e^{i t(\tau(\xi)+y . \xi)} k(\xi) d \xi\right| \leq C_{N}|t|^{-N}
$$

as $t \rightarrow \infty$.
Choose $0<r_{1}<r_{2}$ so that

$$
\operatorname{supp} k \subset\left\{r_{1} \leq|\xi| \leq r_{2}\right\}
$$

Write

$$
\int e^{i t(\tau(\xi)+y \cdot \xi)} k(\xi) d \xi=\int_{r_{1}}^{r_{2}}\left(\int_{|\xi|=1} e^{i t(\tau(\xi)+y \cdot \xi)} k(r \xi) d \sigma(\xi)\right) r^{d-1} d r
$$

It suffices to show that for any $\underline{y} \in \mathcal{M}$ and $\underline{r} \in\left[r_{1}, r_{2}\right]$ one has

$$
\int_{|\xi|=1} e^{i t(\tau(\xi)+y \cdot \xi)} k(r \xi) d \sigma(\xi) \leq C|t|^{-(d-1) / 2}
$$

uniformly for $r, y$ in a neighborhood of $\underline{r}, \underline{y}$.
For $\underline{r}, \underline{y}$ fixed, there is a unique $\underline{\xi}$ with $|\underline{\xi}|=\underline{r}$ for which the phase is stationary and the stationary point is nondegenerate because of the rank equal to $d-1$ hypothesis. It follows that for $r, y$ in a neighborhood, there is a unique uniformly nondegenerate startionary point. The desired estimate follows from the inequality of stationary phase (see Appendix II).

Proposition 3.4.3. Suppose that $0<R_{1}<R_{2}<\infty$ and $\omega:=\left\{\xi \in \mathbb{R}: R_{1}<|\xi|<R_{2}\right\}$. There is a constant $C$ so that for all $f \in L^{1}\left(\mathbb{R}_{x}^{d}\right)$ with $\operatorname{supp} \hat{f} \subset \bar{\omega}$,

$$
u(t, x):=(2 \pi)^{-d / 2} \int e^{i\left(t \tau_{j}(\xi)+x . \xi\right)} \hat{f}(\xi) d \xi:=e^{i t \tau_{j}\left(D_{x}\right)} f
$$

satisfies

$$
\begin{equation*}
\|u(t)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \leq C(1+|t|)^{-(d-1) / 2}\|f\|_{L^{1}\left(\mathbb{R}^{d}\right)} . \tag{3.4.3}
\end{equation*}
$$

The proof is based on a simple idea. The solution $u$ is equal to the convolution of the fundamental solution with $f$. The Fourier transform of the fundamental solution at $t=0$ is equal to a constant. To have an analogous but more regular representation, it is sufficient that one convolve with a solution whose initial data has Fourier Transform equal to this constant on the spectrum of $f$.

Proof. Choose a $k \in C_{0}^{\infty}\left(\mathbb{R}^{d} \backslash 0\right)$ with $k$ equal to $(2 \pi)^{-d / 2}$ on a neighborhood of $\bar{\omega}$. Define $G$ so that $\hat{G}:=k$. Then since $(2 \pi)^{d / 2} k \hat{f}=\hat{f}$ one has $G * f=f$. Since $e^{i t \tau\left(D_{x}\right)}$ is a Fourier multiplier, one has

$$
u(t):=e^{i t \tau\left(D_{x}\right)} f=e^{i t \tau\left(D_{x}\right)}(f * G)=f *\left(e^{i t \tau\left(D_{x}\right)} G\right)
$$

Then

$$
\|u(t)\|_{L^{\infty}} \leq\|f\|_{L^{1}}\left\|e^{i \tau\left(D_{x}\right)} G\right\|_{L^{\infty}}
$$

The preceding Lemma shows that

$$
\left\|e^{i \tau\left(D_{x}\right)} G\right\|_{L^{\infty}} \leq C(1+|t|)^{-(d-1) / 2}
$$

The next subsections consist of two different paths for exploiting the estimates just proved. The first is more elementary and will be used in Chapter 6 to derive, in the spirit of JohnKlainerman, that in high dimensions there is global solvability for maximally dispersive nonlinear problems with small data. The second is devoted to Strichartz estimates which are important in trying to treat existence problems with low regularity data. That in turn is important in trying to pass from local solvability to global solvability for nonlinear problems for which the natural a priori estimates control few derivatives.

## $\S 3.4 .2$. Fixed time dispersive Sobolev estimates.*

We next find decay estimates for $\|u(t)\|_{L^{1}}$ for sources with Fourier transform supported in $\lambda \bar{\omega}$ for $0<\lambda$. The starting point is

$$
\begin{equation*}
\|u(t)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \leq C|t|^{-(d-1) / 2}\|f\|_{L^{1}\left(\mathbb{R}^{d}\right)}, \quad \operatorname{supp} \hat{f} \subset \omega \tag{3.4.4}
\end{equation*}
$$

Proposition 3.4.4. There is a constant $C$ so that for all $\lambda>0$ and $f \in L^{1}$ with $\operatorname{supp} \hat{f} \subset \lambda \omega$, the solution of

$$
L u=0,\left.\quad u\right|_{t=0}=f
$$

satisfies

$$
\begin{equation*}
\|u(t)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \leq C|t|^{-(d-1) / 2}\left\||D|^{(d+1) / 2} f\right\|_{L^{1}\left(\mathbb{R}^{d}\right)} \tag{3.4.5}
\end{equation*}
$$

First verify the dimensions of the homogeneous estimate (3.4.5). With $t, x$ having the dimensions of a length $\ell$, the factor $|t|^{(d-1) / 2}$ has dimension $\ell^{(d-1) / 2}$. On the other hand, in

$$
\left\||D|^{\gamma} f\right\|_{L^{1}\left(\mathbb{R}^{d}\right)}=\int \|\left. D\right|^{\gamma} f \mid d x
$$

the integrand has dimension $\ell^{-\gamma}$ and $d x$ has dimension $\ell^{d}$. In total the right hand side of (3.4.5) has dimension $\ell^{d-\gamma-(d-1) / 2}$. It is dimensionless as is the left hand side exactly when

$$
\gamma:=\frac{d+1}{2} .
$$

Proof. Choose $\psi \in C_{0}^{\infty}\left(\mathbb{R}_{\xi}^{d}\right)$ so that $\psi_{ \pm}=|\xi|^{ \pm \gamma}$ on $\bar{\omega}$. Then

$$
|D|^{\gamma} f=C \hat{\psi}_{+} * f, \quad \text { and } \quad f=C \hat{\psi}_{-} *\left(|D|^{\gamma} f\right)
$$

Young's inequality implies that $\left\||D|^{\gamma} f\right\|_{L^{1}}$ is a norm equivalent to that on the right in (3.4.4) so

$$
\|u(t)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \leq C|t|^{-(d-1) / 2}\left\||D|^{\gamma} f\right\|_{L^{1}\left(\mathbb{R}^{d}\right)}, \quad \operatorname{supp} \hat{f} \subset \omega
$$

* The material in this subsection is not needed for the Strichartz estimates in the next subsection

If $u_{\lambda}(t, x):=u(\lambda t, \lambda x)$ then, $L u_{\lambda}=0$ if and only if $L u=0$ and $\hat{u}_{\lambda}(\lambda t, \xi)=\lambda^{-d} \hat{u}(t, \xi / \lambda)$. The spectrum of $u$ is contained in $\omega$ if and only if the spectrum of $u_{\lambda}$ is contained in $\lambda \omega$.

Exercise. Show that if $f_{\lambda}(x):=f(\lambda x)$,

$$
|D|^{\gamma} f_{\lambda}(x)=\lambda^{-\gamma}\left(|D|^{\gamma} f\right)(\lambda x) .
$$

The change of variable $z=\lambda x$ yields

$$
\left\||D|^{\gamma} f_{\lambda}\right\|_{L^{1}\left(\mathbb{R}^{d}\right)}=\int \lambda^{-\gamma}\left(|D|^{\gamma} f\right)(\lambda x) d x=\lambda^{-\gamma-d}\left\||D|^{\gamma} f\right\|_{L^{1}\left(\mathbb{R}^{d}\right)}
$$

Then (3.4.3) yields

$$
\begin{aligned}
\left\|u_{\lambda}(t)\right\|_{L^{\infty}} & =\|u(\lambda t)\|_{L^{\infty}} \leq C|\lambda t|^{-(d-1) / 2}\left\||D|^{\gamma} f\right\|_{L^{1}\left(\mathbb{R}^{d}\right)} \\
& =C|\lambda t|^{-(d-1) / 2}\left\||D|^{\gamma} f\right\|_{L^{1}\left(\mathbb{R}^{d}\right)} \\
& =C \lambda^{-(d-1) / 2+\gamma+d}|t|^{-(d-1) / 2}\left\||D|^{\gamma} f_{\lambda}\right\|_{L^{1}\left(\mathbb{R}^{d}\right)}
\end{aligned}
$$

The choice $\gamma=(d+1) / 2$ is made so that the $\lambda$ factors cancel.
Since $\hat{u}$ and $\hat{f}$ are locally integrable functions, the point $\xi=0$ is negligible so we have the Littlewood-Paley decompositions

$$
u=\sum_{j=-\infty}^{\infty} \chi\left(2^{-j} D_{x}\right) u:=\sum_{j=-\infty}^{\infty} u_{j}, \quad f=\sum_{j=-\infty}^{\infty} \chi\left(2^{-j} D\right) f:=\sum_{j=-\infty}^{\infty} f_{j},
$$

where the dyadic decomposition of unity is constructed in the Appendix on the stationary phase inequality. This expresses a solution of $L u=0$ as a sum of spectrally localized solutions. The estimates of the next exercise show that $|D|^{\sigma}$ acts like multiplication by $2^{\sigma j}$ on $f_{j}$.

Exercise. Show that there is an integer $k$ and a constant $C$ depending on $\sigma$ and $\chi$ so that for $p \in[1, \infty]$

$$
\begin{align*}
& \left\||D|^{\sigma} f_{j}\right\|_{L^{p}} \leq C 2^{\sigma j} \sum_{|n-j| \leq k}\left\|f_{n}\right\|_{L^{p}},  \tag{3.4.6}\\
& \left\|f_{j}\right\|_{L^{p}} \leq C 2^{-\sigma j} \sum_{|n-j| \leq k}\left\||D|^{\sigma} f_{n}\right\|_{L^{p}} . \tag{3.4.7}
\end{align*}
$$

Theorem 3.4.5. i. If $L u=0$ and $\left.u\right|_{t=0}=f$ then,

$$
\begin{equation*}
\|u\|_{L^{\infty}} \leq C|t|^{-(d-1)} \sum_{j=-\infty}^{\infty}\left\||D|^{\gamma} f_{j}\right\|_{L^{1}}, \quad \gamma=\frac{d+1}{2} \tag{3.4.8}
\end{equation*}
$$

ii. If $0<\delta<\gamma$ there is a constant $C(\gamma, \delta)$ so that

$$
\begin{equation*}
\sum_{j=-\infty}^{\infty}\left\||D|^{\gamma} f_{j}\right\|_{L^{1}} \leq C\left(\left\||D|^{\gamma-\delta} f\right\|_{L^{1}\left(\mathbb{R}^{d}\right)}+\left\||D|^{\gamma+\delta} f\right\|_{L^{1}\left(\mathbb{R}^{d}\right)}\right) \tag{3.4.9}
\end{equation*}
$$

Remarks. 1. The sum on the right of (3.4.8) is the definition of the norm in the homogeneous Besov space $\dot{B}_{1,1}^{\gamma}$. Estimate (3.4.9) yields a bound which is not as sharp but avoids these spaces.
2. A slightly weaker estimate than (3.4.8-3.4.9) was proved by [Lucente-Ziliotti].
S. Lucente and G. Ziliotti, A decay estimate for a class of hyperbolic pseudo-differential equations, Math. App. 10(1999), 173-190, Atti, Acc. Naz. Linccei Cl. Sci. Fis. Mat. Natur. Rend. Lincei (9).
3. It is impossible to have a decay estimate of the form

$$
\|u(t)\|_{L^{\infty}} \leq g(t)\|f\|_{H^{s}}, \quad \lim _{t \rightarrow \infty} g(t)=0
$$

with a conserved norm on the right hand side. If there were such an estimate one can apply it to $v(t)=v(t-T)$ at $t=T \rightarrow \infty$ to find

$$
\|u(0)\|_{L^{\infty}} \leq g(T)\|f\|_{H^{s}} \rightarrow 0
$$

The appearance of norms which are not propagated by the equation is necessary.
4. An $L^{1}$ condition encodes more rapid decay as $|x| \rightarrow \infty$ than an $L^{2}$ condition. This is natural since the energy in a ring $R<|x|<R+1$ can focus at time $t \sim R$ into a ball of radius $O(1)$. If the amplitude in the initial ring in $\sim a$ the $L^{2}$ norm is $\sim a^{2} R^{d-1}$. If the focused amplitude is $\sim A$ one obtains $A^{2} \sim a^{2} R^{d-1}$. If this focussing is to take place at $t \sim R$ and also $A^{2} \leq t^{-(d-1)}$ that yields $a \leq R^{-(d-1)}$ which is on the $L^{1}$ borderline. Thus one cannot have $t^{-(d-1) / 2}$ decay estimates as in the Theorem with $L^{p}$ norms on the right with $p>1$.

Proof of Theorem. i. Estimate (3.4.5) implies

$$
\left\|u_{j}(t)\right\|_{L^{\infty}} \leq C|t|^{-(d-1)}\left\||D|^{\gamma} f_{j}\right\|_{L^{1}} .
$$

Summing yields

$$
\|u\|_{L^{\infty}} \leq \sum\left\|u_{j}\right\|_{L^{\infty}} \leq C|t|^{-(d-1)} \sum\left\||D|^{\gamma} f_{j}\right\|_{L^{1}} .
$$

ii. For $j \geq 0$, estimate (3.4.6) implies

$$
\left\||D|^{\gamma} f_{j}\right\|_{L^{1}} \leq C 2^{\gamma j} \sum_{|n-j| \leq k}\left\|f_{n}\right\|_{L^{1}} .
$$

Estimate (3.4.7) implies

$$
\left\|f_{n}\right\|_{L^{1}} \leq C 2^{-\sigma n} \sum_{|m-n| \leq k}\left\||D|^{\sigma} f_{m}\right\|_{L^{1}}
$$

Finally,

$$
\left\||D|^{\sigma} f_{m}\right\|_{L^{1}} \leq C\left\||D|^{\sigma} f\right\|_{L^{1}}
$$

Combining yields

$$
\sum_{j \geq 0}\left\||D|^{\gamma} f_{j}\right\|_{L^{1}} \leq C\left\||D|^{\sigma} f\right\|_{L^{1}} \sum_{j \geq 0} \sum_{|n-j| \leq k} 2^{\gamma j-\sigma n}
$$

With $\sigma=\gamma+\delta$, the sum is finite, so

$$
\sum_{j \geq 0}\left\||D|^{\gamma} f_{j}\right\|_{L^{1}} \leq C\left\||D|^{\gamma+\delta} f\right\|_{L^{1}}
$$

Exercise. Prove the complementary estimate

$$
\sum_{j<0}\left\||D|^{\gamma} f_{j}\right\|_{L^{1}} \leq C\left\||D|^{\gamma-\delta} f\right\|_{L^{1}}
$$

This completes the proof.
Corollary 3.4.6. For any $d / 2>\delta>0$ there is a constant $C$ so that if $L u=0$, then

$$
\begin{align*}
\|u(t)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \leq & C\langle t\rangle^{-(d-1) / 2}\left(\|f\|_{H^{d / 2+\delta}\left(\mathbb{R}^{d}\right)}\right. \\
& \left.+\left\||D|^{(d+1) / 2+\delta} f\right\|_{L^{1}\left(\mathbb{R}^{d}\right)}+\left\||D|^{(d+1) / 2-\delta} f\right\|_{L^{1}\left(\mathbb{R}^{d}\right)}\right) \tag{3.4.10}
\end{align*}
$$

Remark. The smaller is $\delta>0$ the stronger is the conclusion.
Proof. Sobolev's inequality yields

$$
\|u(t)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)} \leq C\|u(t)\|_{H^{\delta+d / 2}\left(\mathbb{R}^{d}\right)}=C\|f\|_{H^{\delta+d / 2}\left(\mathbb{R}^{d}\right)} .
$$

This yields (3.4.10) for $|t| \leq 1$.
For $|t| \geq 1$ use the two estimates of the Theorem.

## §3.4.3. Strichartz estimates.

The estimates involve norms

$$
\|u\|_{L_{t}^{q} L_{x}^{r}}:=\left(\int_{0}^{\infty}\|u(t)\|_{L^{r}\left(\mathbb{R}_{x}^{d}\right)}^{q} d t\right)^{1 / q}
$$

which integrate over space and time. If such a norm is finite, then the integrand must be small for large times. This requires $r>2$. The estimates express time decay because of dispersion.
The group velocities lie on the strictly convex manifold $\mathcal{M}$. For a typical Fourier Transform, an open set of these velocites is sampled. The method of nonstationary phase shows that for large time the solution is concentrated on the rays with these speeds, starting from the support of the initial data. Thus, a solution is expected to be concentrated on and spread over a region of measure which grows like $t^{d-1}$. An example is concentration in an annulus $\rho_{1}<|x|-t<\rho_{2}$. Or even finer, concentration on that part of the annulus subtending a fixed solid angle.
The conservation of $L^{2}\left(\mathbb{R}^{d}\right)$ and also Lemma 3.4.2 show that the expected amplitude is $O\left(t^{-(d-1) / 2}\right)$. Then

$$
\|u(t)\|_{L^{r}}^{r} \sim t^{-r(d-1) / 2} t^{d-1}
$$

so

$$
\|u\|_{L_{t}^{q} L_{x}^{r}}^{q} \sim \int_{1}^{\infty}\left(t^{-r(d-1) / 2} t^{d-1}\right)^{q / r} d t .
$$

The limiting indices are those for which the power of $t$ is equal to -1 , that is with

$$
\begin{aligned}
& \sigma:=d-1 \\
& \left(\frac{-r \sigma}{2}+\sigma\right) \frac{q}{r}=-1, \quad \text { equivalently, } \quad \frac{-\sigma}{2}+\frac{\sigma}{r}=\frac{-1}{q}
\end{aligned}
$$

The admissible indices are those for which the power is less than or equal to -1 ,

$$
\frac{-\sigma}{2}+\frac{\sigma}{r} \leq \frac{-1}{q}
$$

Definitions. The pair $2<q, r<\infty$ is $\sigma$-admissible if

$$
\frac{1}{q}+\frac{\sigma}{r} \leq \frac{\sigma}{2}
$$

It is sharp $\sigma$-admissible when equality holds.
The estimates involve the homogeneous Sobolev norms

$$
\left\||D|^{\gamma} f\right\|_{L^{2}}:=\left(\left.\left.\int| | \xi\right|^{\gamma} \hat{f}(\xi)\right|^{2} d \xi\right)^{1 / 2}
$$

Theorem 3.4.7 (Strichartz inequalities). Suppose that $L(\partial)$ is maximally dispersive, $\sigma=d-1, q, r$ is $\sigma$-admissible, and $\gamma$ is the solution of

$$
\frac{1}{q}+\frac{d}{r}=\frac{d}{2}-\gamma
$$

There is a constant $C$ so that for $f \in L^{2}$ with $\left\||D|^{\gamma} f\right\|_{L^{2}}<\infty$, the solution of $L u=0$, $\left.u\right|_{t=0}=f$ satisfies

$$
\begin{equation*}
\|u\|_{L_{t}^{q} L_{x}^{r}} \leq C\left\||D|^{\gamma} f\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} . \tag{3.4.11}
\end{equation*}
$$

There are two complicated relations in this assertion. The first is the definition of admissibility. It is the crucial one which encodes the rate of decay of solutions. The second is the definition of $\gamma$. Once admissible $q, r$ are chosen, $\gamma$ is forced so that the two sides of (3.4.11) scale the same for $(t, x) \mapsto(a t, a x)$. From this perspective the dispersion is key as it constrains the $q, r$.
There is a diametrically opposite perspective which starts from the scaling relation which is independent of the dispersion. For example if you are obliged to work with a specific $\gamma$ then the scaling restricts $1 / q, 1 / r$ to lie on a line. Then the admissability chooses an interval on that line. Changing the dispersion, for example considering a problem with the same scaling but weaker dispersion leaves the line fixed but constrains the $1 / q, 1 / r$ to lie on a smaller subinterval.

We follow the proof of [Keel-Tao]. Another standard reference is [Ginibre-Velo]. The limit point case is treated in the first reference. The key step is an estimate for spectrally localized data.

Lemma 3.4.8. Suppose that $\sigma:=d-1, q, r$ is $\sigma$-admissible, and $\omega$ is as in the Corollary. There is a constant $C$ so that for all $f \in L^{2}\left(\mathbb{R}^{d}\right)$ with supp $\hat{f} \subset \bar{\omega}$,

$$
u(t):=e^{i t \tau_{j}\left(D_{x}\right)} f:=U(t) f, \quad U(t)^{*}=U(-t)
$$

satisfies

$$
\begin{equation*}
\|u\|_{L_{t}^{q} L_{x}^{r}} \leq C\|f\|_{L^{2}} \tag{3.4.12}
\end{equation*}
$$

Futhermore, for all $F \in L_{t}^{q^{\prime}} L_{x}^{r^{\prime}}$ with $\operatorname{supp} \hat{F}(t, \cdot) \subset \bar{\omega}$,

$$
\begin{equation*}
\left\|\int_{0}^{\infty} U(s)^{*} F(s) d s\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} \leq C\|F\|_{L_{t}^{q^{\prime}} L_{x}^{r^{\prime}}} \tag{3.4.13}
\end{equation*}
$$

Discussion. The estimate is true in the sharp admissible case even though for the heuristics given before the definition, the integral diverged. It is not possible to achieve the concentration suggested in the heuristics with data which has spectrum with support in an annulus. For example, if one considers the wave operator $\square$ on $\mathbb{R}^{1+3}$ with data supported
in $|x| \leq 1$ the solutions are supported in $|x|-t \leq 1$ and decay along with their derivatives exactly as in the heuristic. Thus one gets divergent integrals. However, compact support and compactly supported Fourier transform are not compatible, and the compact spectrum is enough to overcome the divergence.

Proof. Denote by (, ) the $L^{2}\left(\mathbb{R}^{d}\right)$ scalar product. Since.

$$
\int_{0}^{\infty}(U(t) f, F(t)) d t=\int_{0}^{\infty}\left(f, U(t)^{*} F(t)\right) d t=\left(f, \int_{0}^{\infty} U(t)^{*} F(t) d t\right)
$$

estimates (3.4.12) and (3.4.13) are equivalent thanks to the duality representations of the norms,

$$
\begin{gathered}
\left\|\int_{0}^{\infty} U(t)^{*} F(t) d t\right\|_{L^{2}\left(\mathbb{R}^{d}\right)}=\sup \left\{\left(f, \int_{0}^{\infty} U(t)^{*} F(t) d t\right): \hat{f} \in C_{0}^{\infty}(\omega),\|f\|_{L^{2}}=1\right\} \\
\|U(t) f\|_{L^{q} L^{r}}=\sup \left\{\int_{0}^{\infty}(U(t) f, F(t)) d t: \hat{F} \in C_{0}^{\infty}(] 0, \infty[\times \omega), \quad\|F\|_{L^{q^{\prime}} L^{r^{\prime}}}=1\right\}
\end{gathered}
$$

Estimate (3.4.13) holds if and only if

$$
\left(\int_{0}^{\infty}\left(U(t)^{*} F(t)\right) d t, \int_{0}^{\infty}\left(U(s)^{*} G(s)\right) d s\right)
$$

is a continuous bilinear form on $L^{q^{\prime}} L^{r^{\prime}}$, that is

$$
\begin{equation*}
\left|\int_{0}^{\infty} \int_{0}^{\infty}\left(U(s)^{*} F(s), U(t)^{*} G(t)\right) d s d t\right| \leq C\|F\|_{L_{t}^{q^{\prime} L_{x}^{r^{\prime}}}}\|G\|_{L_{t}^{q^{\prime} L_{x}^{r^{\prime}}}} \tag{3.4.14}
\end{equation*}
$$

Unitarity implies that

$$
\forall s, t, \quad B:=U(t) U^{*}(s), \quad \text { satisfies } \quad\|B f\|_{L^{2}} \leq\|f\|_{L^{2}}
$$

The dispersive estimate (3.4.3) is

$$
\forall s, t, \quad\|B f\|_{L^{\infty}} \leq C\langle t-s\rangle^{-\sigma}\|f\|_{L^{1}}
$$

With $\left.r^{\prime} \in\right] 1,2[$ the dual index to $r$, choose $\theta \in] 0,1[$ so that

$$
\begin{equation*}
\frac{1}{r^{\prime}}=\theta \frac{1}{1}+(1-\theta) \frac{1}{2}, \quad \text { then, } \quad \theta=\frac{2-r^{\prime}}{r^{\prime}}=\frac{r-2}{r} \tag{3.4.15}
\end{equation*}
$$

The Riesz-Thorin Theorem implies that

$$
\|B f\|_{L^{r}} \leq C^{\theta}\langle t-s\rangle^{-\sigma \theta}\|f\|_{L^{r^{\prime}}}
$$

With Hölder's inequality, this yields the interpolated bilinear estimate,

$$
\left|\left(U(s)^{*} F(s), U(t)^{*} G(t)\right)\right| \leq C^{\theta}\langle t-s\rangle^{-\sigma \theta}\|F(s)\|_{L^{r^{\prime}}}\|G(t)\|_{L^{r^{\prime}}}
$$

Admissibility implies that

$$
\frac{1}{q} \leq \sigma\left(\frac{1}{2}-\frac{1}{r}\right)=\sigma\left(\frac{r-2}{2 r}\right)=\frac{\sigma \theta}{2}
$$

When strict inequality holds in the definition of admissibility, $\langle t-s\rangle^{-\sigma \theta} \in L^{q / 2}\left(\mathbb{R}_{t}\right)$. The hypothesis $q>2$ is used here. For the limiting case, it is nearly so. The HardyLittlewood inequality shows that convolution with $|t|^{-2 / q}$ has the $L^{p}$ mapping properties that convolution with an element of $L^{q / 2}(\mathbb{R})$ would have.
The Hausdorff-Young inequality shows that

$$
\begin{equation*}
L^{p_{1}} * L^{p_{2}} \subset L^{p_{3}}, \quad \text { provided } \quad \frac{1}{p_{1}}+\frac{1}{p_{2}}=1+\frac{1}{p_{3}} \tag{3.4.16}
\end{equation*}
$$

The Hardy Littlewood inequality asserts that when $1<p_{1}, p_{2}, p_{3}<\infty$

$$
\begin{equation*}
\frac{1}{\langle t\rangle^{1 / p_{1}}} * L^{p_{2}}(\mathbb{R}) \subset L^{p_{3}}(\mathbb{R}), \quad \text { provided } \quad \frac{1}{p_{1}}+\frac{1}{p_{2}}=1+\frac{1}{p_{3}} \tag{3.4.17}
\end{equation*}
$$

Set

$$
\begin{equation*}
p_{1}=\frac{q}{2}, \quad p_{2}=q^{\prime}, \quad \text { and }, \quad p_{3}=q \tag{3.4.18}
\end{equation*}
$$

The index conditions in (3.4.16)-(3.4.17) become

$$
\frac{2}{q}+\frac{1}{q^{\prime}}=1+\frac{1}{q}
$$

which holds by definition of $q^{\prime}$. Then (3.4.16) in the admissible case and (3.4.17) in the sharp admissible case imply that

$$
\begin{equation*}
\left\|\int_{-\infty}^{\infty}\langle t-s\rangle^{-\sigma \theta}\right\| F(s)\left\|_{L^{r^{\prime}}} d s\right\|_{L^{q}\left(\mathbb{R}_{t}\right)} \leq C\|F\|_{L_{t}^{q^{\prime}} L_{x}^{r^{\prime}}} \tag{3.4.19}
\end{equation*}
$$

Hölder's inequality yields

$$
\int_{0}^{\infty}\left(\int_{0}^{\infty}\langle t-s\rangle^{-\sigma \theta}\|F(s)\|_{L^{r^{\prime}}} d s\right)\|G(t)\|_{L^{r^{\prime}}} d t \leq C\|F\|_{L_{t}^{q^{\prime}} L_{x}^{r^{\prime}}}\|G\|_{L_{t}^{q^{\prime}} L_{x}^{r^{\prime}}}
$$

This proves the desired estimate (3.4.14).
A scaling yields estimates for sources with Fourier transform suppoerted in $\lambda \bar{\omega}$ for $0<\lambda$.

Lemma 3.4.9. With $q, r, \omega, \sigma$ as in the previous lemma and $\gamma$ as in the Theorem, there is a $C$ so that for all $0<\lambda$ and $f \in L^{2}$ with $\operatorname{supp} \hat{f} \subset \lambda \bar{\omega}$,

$$
u(t):=e^{i t \tau_{j}\left(D_{x}\right)} f:=U(t) f
$$

satisfies

$$
\begin{equation*}
\|u\|_{L_{t}^{q} L_{x}^{r}} \leq C\left\||D|^{\gamma} f\right\|_{L^{2}} \tag{3.4.20}
\end{equation*}
$$

Proof of Lemma. If $u_{\lambda}(t, x):=u(\lambda t, \lambda x)$ then, $L u_{\lambda}=0$ and the spectrum of $u_{\lambda}$ is contained in $\bar{\omega}$.
The two sides of (3.4.12) scale differently. Compute

$$
\left\|u_{\lambda}(t)\right\|_{L^{r}}=\left(\int\left|u_{\lambda}(t, x)\right|^{r} d x\right)^{1 / r}=\left(\int|u(\lambda t, \lambda x)|^{r} d x\right)^{1 / r}
$$

The substitution $y=\lambda x, d x=\lambda^{-d} d x$ yields

$$
=\lambda^{-d / r}\left(\int|u(\lambda t, y)|^{r} d y\right)^{1 / r}=\lambda^{-d / r}\|u(\lambda t)\|_{L^{r}}
$$

A similar change of variable for the time integral shows that

$$
\left\|u_{\lambda}\right\|_{L_{t}^{q} L_{x}^{r}}=\lambda^{-1 / q-d / r}\|u\|_{L_{t}^{q} L_{x}^{r}} .
$$

For any $\gamma,\left\||D|^{\gamma} f\right\|_{L^{2}}$ is a norm equivalent to the norm on the right hand side for sources with spectrum in $\bar{\omega}$. Compute

$$
\begin{aligned}
\left\||D|^{\gamma} f_{\lambda}\right\|_{L^{2}} & =\left(\int|\xi|^{2 \gamma}\left|\hat{f}_{\lambda}(\xi)\right|^{2} d \xi\right)^{1 / 2}=\left(\int|\xi|^{2 \gamma}\left|\lambda^{-d} \hat{f}(\xi / \lambda)\right|^{2} d \xi\right)^{1 / 2} \\
& =\lambda^{\gamma-d / 2}\left(\int|\xi|^{2 \gamma}|\hat{f}(\xi)|^{2} d \xi\right)^{1 / 2}=\lambda^{\gamma-d / 2}\left\||D|^{\gamma} f\right\|_{L^{2}}
\end{aligned}
$$

Given $q, r$, the $\gamma$ of the Theorem is the unique value so that the two norms scale the same. Therefore the estimate of the present Lemma follows from the preceding Lemma.

Proof of Theorem. With $\chi$ from the dyadic partition of unity for $\mathbb{R}_{\xi}^{d} \backslash 0$ constructed in the stationary phase inequality, introduce the Littlewood-Paley decomposition of tempered distributions

$$
g=\sum_{J \in \mathbb{Z}} g_{j}, \quad g_{j}:=\chi\left(D / 2^{j}\right) g:=(2 \pi)^{-d / 2} \int e^{i x \xi} \chi\left(\xi / 2^{j}\right) \hat{g}(\xi) d \xi
$$

Then for $1<r<\infty$ the classical square function estimate (see [Stein, Singular Integrals]) asserts that there is a $C>1$ so that

$$
C^{-1}\|g\|_{L^{r}} \leq\left\|\left(\sum_{j \in \mathbb{Z}}\left|g_{j}\right|^{2}\right)^{1 / 2}\right\|_{L^{r}} \leq C\|g\|_{L^{r}}
$$

Lemma 3.4.10. If $2 \leq q, r<\infty$, there is a constant $C$ so that

$$
\begin{equation*}
\|F\|_{L_{t}^{q} L_{x}^{r}}^{2} \leq C \sum_{j \in \mathbb{Z}}\left\|F_{j}\right\|_{L_{t}^{q} L_{x}^{r}}^{2}, \tag{3.4.21}
\end{equation*}
$$

where $F(t)=\sum_{j} F_{j}(t)$ is the Littlewood-Paley decomposition in $x$.
Proof of Lemma. The square function estimate yields

$$
\|F(t)\|_{L_{x}^{r}}^{2} \leq C \int\left(\sum_{j}\left|F_{j}(t)\right|^{2}\right)^{r / 2} d x=C\left\|\sum_{j}\left|F_{j}(t)\right|^{2}\right\|_{L^{r / 2}}
$$

Minkowski's inequality in $L^{r / 2}$ shows that this is

$$
\leq C \sum_{j}\left\|F_{j}(t)^{2}\right\|_{L^{r / 2}}=C \sum_{j}\left\|F_{j}(t)\right\|_{L^{r}}^{2}
$$

Using this yields

$$
\|F\|_{L_{t}^{q} L_{x}^{r}}^{2} \leq C\left(\int_{0}^{\infty}\left(\sum_{j}\left\|F_{j}(t)\right\|_{L^{r}}^{2}\right)^{q / 2} d t\right)^{2 / q}=C\left\|\sum_{j}\right\| F_{j}(t)\left\|_{L^{r}\left(\mathbb{R}_{x}^{d}\right)}^{2}\right\|_{L^{q / 2}\left(\mathbb{R}_{t}\right)}
$$

Minkowski's inequality in $L^{q / 2}\left(\mathbb{R}_{t}\right)$ shows this is

$$
\leq C \sum_{j}\| \| F_{j}(t)\left\|_{L^{r}\left(\mathbb{R}_{x}^{d}\right)}^{2}\right\|_{L^{q / 2}\left(\mathbb{R}_{t}\right)}=C \sum_{j}\left\|F_{j}(t)\right\|_{L_{t}^{q} L_{x}^{r}}^{2} .
$$

Return now to the proof of the Theorem. Associate to the sheet $\tau=\tau_{k}(\xi)$ the projector $\pi_{k}(\xi):=\pi\left(\tau_{k}(\xi), \xi\right)$ from $\S 3.2$. The $\pi_{k}$ are real analytic on $\xi \neq 0$ and homogeneous of degree 0 in $\xi$. In addition $\sum_{k} \pi_{k}=I$. The solution $u$ satisfies

$$
u=\sum_{k} e^{i t \tau_{k}(D)} \pi_{k}(D) f:=\sum_{k} u_{k}
$$

Apply (3.4.21) to $u_{k}$ to find using (3.4.20)

$$
\left\|u_{k}\right\|_{L_{t}^{q} L_{x}^{r}}^{2} \leq C \sum_{j}\left\|u_{k, j}\right\|_{L_{t}^{q} L_{x}^{r}}^{2} \leq C^{\prime} \sum_{j}\left\||D|^{\gamma} \pi_{k}(D) f_{j}\right\|_{L^{2}}^{2} \leq C^{\prime}\left\||D|^{\gamma} f\right\|_{L^{2}}^{2} .
$$

The finite sum on $k$ completes the proof of the Theorem.
Corollary 3.4.11. Denote by $S(t)$ the $L^{2}$ unitary mapping $u(0) \mapsto u(t)$ for solutions of $L u=0$. With the indices of the Theorem one has

$$
\begin{equation*}
\left\|\int_{0}^{\infty} S(s)^{*} F(s) d s\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} \leq C\left\|\left|D_{x}\right|^{\gamma} F\right\|_{L_{t}^{q^{\prime}} L_{x}^{r^{\prime}}} \tag{3.4.22}
\end{equation*}
$$

Proof. Estimate (3.4.22) is equivalent to the Strichartz estimate (3.4.11) by a duality like that used to establish the equivalence of (3.4.12) and (3.4.13).

Exercise. Prove the following complement to (3.4.21) which comes from the other side of the square function inequality. If $1<p \leq 2$ and $1 \leq r \leq 2$ then there is a $C$ so that

$$
\begin{equation*}
\sum_{j=-\infty}^{\infty}\left\|F_{j}\right\|_{L_{t}^{r} L_{x}^{p}}^{2} \leq C\|F\|_{L_{t}^{r} L_{x}^{p}}^{2} \tag{3.4.23}
\end{equation*}
$$

## §6.8. The subcritical nonlinear Klein-Gordon equation in the energy space.

## §6.8.1. Introductory remarks.

The mass zero nonlinear Klein-Gordon equation is

$$
\begin{equation*}
\square_{1+d} u+F(u)=0 . \tag{6.8.1}
\end{equation*}
$$

where

$$
\begin{equation*}
F \in C^{1}(\mathbb{R}), \quad F(0)=0, \quad F^{\prime}(0)=0 \tag{6.8.2}
\end{equation*}
$$

The classic examples from quantum field theory are the equations with $F(u)=u^{p}$ with $p \geq 3$ an odd integer. For ease of reading we consider only real solutions.
The equation (6.8.1) is Lorentz invariant and if

$$
\begin{equation*}
G^{\prime}(s)=F(s), \quad G(0)=0 \tag{6.8.3}
\end{equation*}
$$

The local energy density is defined as

$$
\begin{equation*}
e(u):=\frac{u_{t}^{2}+\left|\nabla_{x} u\right|^{2}}{2}+G(u) \tag{6.8.4}
\end{equation*}
$$

Solutions $u \in H_{\text {loc }}^{2}\left(\mathbb{R}^{1+d}\right)$ satisfy the differential energy law,

$$
\begin{equation*}
\partial_{t} e-\operatorname{div}\left(u_{t} \nabla_{x} u\right)=u_{t}(\square u+F(u))=0 \tag{6.8.5}
\end{equation*}
$$

The corresponding integral conservation law for solutions suitably small at infinity is,

$$
\begin{equation*}
\partial_{t} \int_{\mathbb{R}^{d}} \frac{u_{t}^{2}+\left|\nabla_{x} u\right|^{2}}{2}+G(u) d x=0 \tag{6.8.6}
\end{equation*}
$$

is one of the fundamental estimates in this section. Solutions are stationary for the Lagrangian,

$$
\int_{0}^{T} \int_{\mathbb{R}^{d}} \frac{u_{t}^{2}-\left|\nabla_{x} u\right|^{2}}{2}-G(u) d t d x .
$$

When $F$ is smooth, the methods of $\S 6.3-6.4$ yield local smooth existence.
Theorem 6.8.1. If $F \in C^{\infty}, s>d / 2, f \in H^{s}\left(\mathbb{R}^{d}\right)$, and $g \in H^{s-1}\left(\mathbb{R}^{d}\right)$, then there is a unique maximal solution

$$
u \in C\left(\left[0, T_{*}\left[; H^{s}\left(\mathbb{R}^{d}\right)\right) \cap C^{1}\left(\left[0, T_{*}\left[; H^{s-1}\left(\mathbb{R}^{d}\right)\right)\right.\right.\right.\right.
$$

satisfying

$$
u(0, x)=f, \quad u_{t}(0, x)=g
$$

If $T_{*}<\infty$ then

$$
\limsup _{t \rightarrow T_{*}}\|u(t)\|_{L^{\infty}\left(\mathbb{R}^{d}\right)}=\infty
$$

In favorable cases, the energy law (6.8.6) gives good control of the norm of $u, u_{t} \in H^{1} \times L^{2}$. Controling the norm of the difference of two solutions is, in contrast, a very difficult problem for which many fundamental questions remain unresolved.
An easy first case is nonlinearities $F$ which are uniformly lipschitzean. In this case, there is global existence in the energy space.

Theorem 6.8.2. If $F$ satisfies $F^{\prime} \in L^{\infty}(\mathbb{R})$, then for all Cauchy data $f, g \in H^{1} \times L^{2}$ there is a unique solution

$$
u \in C\left(\mathbb{R} ; H^{1}\left(\mathbb{R}^{d}\right)\right) \cap C^{1}\left(\mathbb{R} ; L^{2}\left(\mathbb{R}^{d}\right)\right)
$$

For any finite $T$, the map from data to solution is uniformly lipschitzean from $H^{1} \times L^{2}$ to $C\left(\left[-T, T ; H^{1}\right) \cap C^{1}\left([-T, T] ; L^{2}\right)\right.$. If $f, g \in H^{2} \times H^{1}$ then

$$
u \in L^{\infty}\left(\mathbb{R} ; H^{2}\left(\mathbb{R}^{d}\right)\right), \quad u_{t} \in L^{\infty}\left(\mathbb{R} ; H^{1}\left(\mathbb{R}^{d}\right)\right)
$$

If $f, g \in H^{s} \times H^{s-1}$ with $1 \leq s<2$, then

$$
u \in C\left(\mathbb{R} ; H^{s}\left(\mathbb{R}^{d}\right)\right), \quad u_{t} \in C\left(\mathbb{R} ; H^{s-1}\left(\mathbb{R}^{d}\right)\right)
$$

Sketch of Proof. The key estimate is the following. If $u$ and $v$ are solutions then

$$
\square(u-v)=F(v)-F(u), \quad|F(u)-F(v)| \leq C|u-v|
$$

Multiplying by $u_{t}-v_{t}$ yields
$\frac{d}{d t} \int\left(u_{t}-v_{t}\right)^{2}+\left|\nabla_{x}(u-v)\right|^{2} d x=2 \int\left(u_{t}-v_{t}\right)(F(v)-F(u)) d x \leq C\left\|u_{t}-v_{t}\right\|_{L^{2}}^{2}\|u-v\|_{L^{2}}^{2}$.
It follows that for any $T$ there is an a priori estimate
$\sup _{|t| \leq T}\left(\|u(t)-v(t)\|_{H^{1}}+\left\|u_{t}-v_{t}\right\|_{L^{2}}\right) \leq C(T)\left(\|u(0)-v(0)\|_{H^{1}}+\left\|u_{t}(0)-v_{t}(0)\right\|_{L^{2}}\right)$.
This estimate exactly corresponds to the asserted Lipschitz continuity of the map from data to solutions.
Applying the estimate to $v=u(x+h)$ and taking the supremum over small vectors $h$, yields an a priori estimate

$$
\sup _{|t| \leq T}\left(\|u(t)\|_{H^{2}}+\left\|u_{t}\right\|_{L^{2}}\right) \leq C(T)\left(\|u(0)\|_{H^{2}}+\left\|u_{t}(0)\right\|_{H^{1}}\right)
$$

which is the estimate correponding to the $H^{2}$ regularity.
Higher regularity for dimensions $d \geq 10$ is an outstanding open problem. For example, for $d \geq 10$, smooth compactly supported initial data, and $F \in C_{0}^{\infty}$ or $F=\sin u$, it is not known if the above global unique solutions are smooth. For $d \leq 9$ the result can be found in [Brenner-vonWahl 1982]. Smoothness would follow if one could prove that $u \in L_{\text {loc }}^{\infty}$. What is needed is to show that the solutions do not get large in the pointwise sense. Compared to the analogous regularity problem for Navier-Stokes this problem has the advantage that solutions are known to be unique and depend continuously on the data.

## §6.8.2. The ordinary differential equation and nonlipshitzean $F$.

Concerning global existence for functions $F(u)$ which may grow more rapidly than linearly as $u \rightarrow \infty$, the first considerations concern solutions which are independent of $x$ and therefore satisfy the ordinary differential equation,

$$
\begin{equation*}
u_{t t}+F(u)=0 \tag{6.8.7}
\end{equation*}
$$

Global solvability of the ordinary differential equation is analysed using the energy conservation law

$$
\left(\frac{u_{t}^{2}}{2}+G(u)\right)^{\prime}=u_{t}\left(u_{t t}+F(u)\right)=0
$$

Think of the equation as modeling a nonlinear spring. The spring force is attractive, that is pulls the spring toward the origin when

$$
F(u)>0 \quad \text { when } \quad u>0 \quad \text { and, } \quad F(u)<0 \quad \text { when } \quad u<0 .
$$

In this case one has $G(u)>0$ for all $u \neq 0$. Conservation of energy then gives a pointwise bound on $u_{t}$ uniform in time

$$
u_{t}^{2}(t) \leq u_{t}^{2}(0)+2 G(u(0)), \quad\left|u_{t}(t)\right| \leq\left(u_{t}^{2}(0)+2 G(u(0))\right)^{1 / 2}
$$

This gives a pointwise bound

$$
|u(t)| \leq|u(0)|+|t|\left(u_{t}^{2}(0)+2 G(u(0))\right)^{1 / 2}
$$

In particular the ordinary differential equation has global solutions.
In the extreme opposite case consider the replusive spring force $F(u)=-u^{2}$ and $G(u)=$ $-u^{3} / 3$. The energy law asserts that $u_{t}^{2} / 2-u^{3} / 3:=E$ is independent of time. Consider solutions with

$$
u(0)>0, \quad u_{t}(0)>0 \quad \text { so } \quad E>-\frac{u^{3}(0)}{3} .
$$

For all $t>0$,

$$
\left|u_{t}\right|=\left|\frac{u^{3}}{3}+E\right|^{1 / 2}
$$

At $t=0$ one has

$$
u_{t}(0)=\left(\frac{u^{3}(0)}{3}+E\right)^{1 / 2}>0
$$

Therefore $u$ increases and $u^{3} / 3+E$ stays positive and one has for $t \geq 0$

$$
u_{t}(t)=\left(\frac{u^{3}(t)}{3}+E\right)^{1 / 2}>0
$$

Both $u$ and $u_{t}$ are strictly increasing.
Since

$$
\frac{d u}{\left(\frac{u^{3}}{3}+E\right)^{1 / 2}}=d t
$$

$u(t)$ approaches $\infty$ at time

$$
T:=\int_{u(0)}^{\infty} \frac{d u}{\left(\frac{u^{3}}{3}+E\right)^{1 / 2}} .
$$

Exercise. Show that if there is an $M>0$ so that $G(s)<0$ for $s \geq M$ and

$$
\int_{M}^{\infty} \frac{1}{\sqrt{|G(s)|}} d s<\infty
$$

then there are solutions of the ordinary differential equation which blow up in finite time.

Proposition 6.8.3 [J.B. Keller 1957]. If

$$
a, \delta>0, \quad d \leq 3, \quad E:=\delta^{2} / 2-a^{3} / 3, \quad T:=\int_{a}^{\infty}\left|\frac{u^{3}}{3}+E\right|^{-1 / 2} d u
$$

and $\phi, \psi \in C^{\infty}\left(\mathbb{R}^{d}\right)$ satisfy

$$
\phi \geq a \quad \text { and } \quad \psi \geq \delta \quad \text { for } \quad|x| \leq T
$$

the the smooth solution of

$$
\square_{1+d} u-u^{2}, \quad u(0)=\phi, \quad u_{t}(0)=\psi
$$

blows up on or before time $T$.
Proof. Denote by $\underline{u}$ the solution of the ordinary differential equation with initial data $\underline{u}(0)=a, \underline{u}_{t}(0)=\delta$.
If $u \in C^{\infty}\left([0, \underline{t}] \times \mathbb{R}^{d}\right)$, then finite speed of propagation and positivity of the fundamental solution of $\square_{1+d}$ imply that

$$
u \geq \underline{u} \quad \text { on } \quad\{|x| \leq T-\underline{t}\} .
$$

Since $\underline{u}$ diverges as $t \rightarrow T$ it follows that $\underline{t} \leq T$
In the case of attractive forces where $G \geq 0$ one can hope that there is global smooth solvability for smooth initial data. This question has received much attention and is very far from being understood. For example even in the uniformly lipschitzean case where solutions $H^{2}$ in $x$ exist globally, higher regularity is unknown in high dimensions.

In the remainder of this section we will study solvability in the energy space defined by $u, u_{t} \in H^{1} \times L^{2}$. This regularity is suggested by the basic energy law. For uniformly lipschitzean nonllinearities the global solvability is given by Theorem 6.8.2. The interest is in attractive nonlinearities with superlinear growth at infinity.

A crucial role is played by the rate of growth of $F$ at infinity. There is a critical growth rate so that for nonlinearities which are subcritical and critical there is a good theory based on Strichartz estimates. The analysis is valid in all dimensions.
To concentrate on essentials, we present the family of attractive (repulsive) nonlinearities $F=u|u|^{p-1}\left(F=-u|u|^{p-1}\right)$ with potential energies given by $\pm \int|u|^{p+1} /(p+1) d x$. Start with four natural notions of subcriticality. They are in increasing order of strength. One could expect to call $p$ subcritical when

1. $H^{1}\left(\mathbb{R}^{d}\right) \subset L^{p}\left(\mathbb{R}^{d}\right)$ so the nonlinear term makes sense for elements of $H^{1}$.
2. $H^{1}\left(\mathbb{R}^{d}\right) \subset L^{p+1}\left(\mathbb{R}^{d}\right)$ so the potential energy makes sense for elements of $H^{1}$.
3. $H^{1}\left(\mathbb{R}^{d}\right)$ is compact in $L_{\text {loc }}^{p+1}\left(\mathbb{R}^{d}\right)$ so the potential energy is in a sense small compared to the kinetic energy.
4. $H^{1}\left(\mathbb{R}^{d}\right) \subset L^{2 p}\left(\mathbb{R}^{d}\right)$ so the nonlinear term belongs to $L^{2}\left(\mathbb{R}^{d}\right)$ for elements of $H^{1}$.

The Sobolev embedding is

$$
\begin{equation*}
H^{1}\left(\mathbb{R}^{d}\right) \subset L^{q}\left(\mathbb{R}^{d}\right), \quad \text { for, } \quad q=\frac{2 d}{d-2} \tag{6.8.8}
\end{equation*}
$$

The above conditions then read (with the values for $d=3$ given in parentheses),

1. $p \leq 2 d /(d-2), \quad(p \leq 6)$,
2. $p+1 \leq 2 d /(d-2), \quad($ equiv. $\quad p \leq(d+2) /(d-2)), \quad(p \leq 5)$,
3. $p<(d+2) /(d-2), \quad(p<5)$,
4. $p \leq d /(d-2), \quad(p \leq 3)$.

The correct answer is 3. Much that will follow can be extended to the critical case $p=$ $(d+2) /(d-2)$. The case $\mathbf{1}$ in contrast is supercritical and comparatively little is known. It is known that in the supercritical case, solutions are very sensitive to initial data. The dependence is not lipschitzean, and it is lipschitzean in the subcritical and critical cases. The books of Sogge, and Shatah-Struwe and the orignal 1985 article of Ginibre and Velo are good references. The sensitive dependence is a recent result of Lebeau.

Notation. Denote by $L_{t}^{q} L_{x}^{r}([0, T])$ the space $L_{t}^{q} L_{x}^{r}\left([0, T] \times \mathbb{R}^{d}\right)$, Denote with an open interval

$$
L_{t}^{q} L_{x}^{r}\left(\left[0, T[):=\cup_{0<\underline{T}<T} L_{t}^{q} L_{x}^{r}([0, \underline{T}])\right.\right.
$$

Theorem 6.8.4. i. If $p$ is subcritical for $H^{1}$, that is $p<(d+2) /(d-2)$, then for any $f \in H^{1}\left(\mathbb{R}^{d}\right)$ and $g \in L^{2}\left(\mathbb{R}^{d}\right)$ there is $T_{*}>0$ and a unique solution

$$
\begin{equation*}
u \in C\left(\left[0, T_{*}\left[H^{1}\left(\mathbb{R}^{d}\right)\right) \cap C^{1}\left(\left[0, T_{*}\left[; L^{2}\left(\mathbb{R}^{d}\right)\right) \cap L_{t}^{p} L_{x}^{2 p}\left(\left[0, T_{*}[)\right.\right.\right.\right.\right.\right. \tag{6.8.9}
\end{equation*}
$$

of the repulsive problem

$$
\begin{equation*}
\square u-u|u|^{p-1}=0, \quad u(0)=f, \quad u_{t}(0)=g \tag{6.8.10}
\end{equation*}
$$

If $T_{*}<\infty$ then

$$
\begin{equation*}
\liminf _{t / T_{*}}\left\|\nabla_{t, x} u\right\|_{L^{2}\left(\mathbb{R}^{d}\right)}=\infty \tag{6.8.11}
\end{equation*}
$$

The energy conservation law (6.8.6) is satisfied.
ii. For the attractive problem

$$
\begin{equation*}
\square u+u|u|^{p-1}=0, \quad u(0)=f, \quad u_{t}(0)=g \tag{6.8.12}
\end{equation*}
$$

one has the same result with $T_{*}=\infty$ and with $u \in L_{t}^{p} L_{x}^{2 p}(\mathbb{R})$. For any $T>0$, the map from Cauchy data to solution is uniformly lipschitzean

$$
\left.H^{1} \times L^{2} \quad \rightarrow \quad C\left([-T, T] ; H^{1}\right) \cap C\left([-T, T] ; L^{2}\right) \cap L_{t}^{p} L_{x}^{2 p}([0, T])\right)
$$

In the proof of this result and all that follows a central role is played by the linear wave equation and its solution for which we recall the basic energy estimate

$$
\left\|\nabla_{t, x} u(t)\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} \leq\left\|\nabla_{t, x} u(0)\right\|_{L^{2}\left(\mathbb{R}^{d}\right)}+\int_{0}^{t}\|\square u(t)\|_{L^{2}\left(\mathbb{R}^{d}\right)} d t .
$$

This is completed by the $L^{2}$ estimate

$$
\|u(t)\|_{L^{2}\left(\mathbb{R}^{d}\right)} \leq \int_{0}^{t}\left\|u_{t}(t)\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} d t
$$

In particular, for $h \in L_{\text {loc }}^{1}\left(\mathbb{R} ; L^{2}\left(\mathbb{R}^{d}\right)\right)$ there is a unique solution

$$
u \in C\left(\mathbb{R} ; H^{1}\left(\mathbb{R}^{d}\right)\right) \cap C^{1}\left(\mathbb{R} ; L^{2}\left(\mathbb{R}^{d}\right)\right)
$$

to

$$
\square u=h, \quad u(0)=0, \quad u_{t}(0)=0 .
$$

This solution is denoted

$$
\square^{-1} h .
$$

In order to take advantage of this we seek solutions so that

$$
F_{p}(u):= \pm u|u|^{p-1} \in L_{t}^{1} L_{x}^{2} .
$$

Compute

$$
\left\|F_{p}(u)\right\|_{L_{t}^{1} L_{x}^{2}}=\int_{0}^{T}\left(\int\left|u^{p}\right|^{2} d x\right)^{1 / 2} d t
$$

where

$$
\left(\int|u|^{2 p} d x\right)^{1 / 2}=\left[\left(\int|u|^{2 p}\right)^{1 / 2 p}\right]^{p}=\|u\|_{L^{2 p}\left(\mathbb{R}^{d}\right)}^{p}
$$

so

$$
\begin{equation*}
\left\|F_{p}(u)\right\|_{L_{t}^{1} L_{x}^{2}}=\int_{0}^{T}\|u\|_{L_{t}^{2 p} \mathbb{R}_{x}^{d}}^{p} d t=\|u\|_{L_{t}^{p} L_{x}^{2 p}}^{p} \tag{6.8.13}
\end{equation*}
$$

The above calculation proves the first part of the next lemma.
Lemma 6.8.5. The mapping $u \mapsto F_{p}(u)$ takes $L_{t}^{p} L_{x}^{2 p}\left([0, T]\right.$ to $L_{t}^{1} L_{x}^{2}([0, T])$. It is uniformly Lipshitzean on bounded subsets.

Proof. Write

$$
F_{p}(v)-F_{p}(w)=G(v, w)(v-w), \quad|G(v, w)| \leq C\left(|v|^{p-1}+|w|^{p-1}\right)
$$

Write

$$
\|G(v, w)(v-w)\|_{L_{x}^{2}}^{2}=\int|G|^{2}|v-w|^{2} d x
$$

Use Hölder's inequality for $L_{x}^{p /(p-1)} \times L_{x}^{p}$ to estimate by

$$
\leq\left(\int|G(v, w)|^{2 p /(p-1)} d x\right)^{\frac{p-1}{p}}\left(\int|v-w|^{2 p} d x\right)^{\frac{1}{p}}
$$

Then

$$
\left\|F_{p}(v)-F_{p}(w)\right\|_{L^{2}} \leq C\|v, w\|_{L_{x}^{2 p}}^{p-1}\|v-w\|_{L_{x}^{2 p}}
$$

Finally estimate the integral in time using Hölder's inequality for $L_{t}^{p /(p-1)} \times L_{t}^{p}$.
It is natural to seek solutions $u \in L_{t}^{p} L_{x}^{2 p}([0, T])$. With that as a goal we ask when it is true that

$$
\square^{-1}\left(L_{t}^{1} L_{x}^{2}\right) \subset L_{t}^{p} L_{x}^{2 p}
$$

This is exactly in the family of questions addressed by the Strichartz inequalities. The next Lemma gives the inequalities adapted to the present situation.

Lemma 6.8.6. If

$$
\begin{equation*}
q>2, \quad \text { and } \quad \frac{1}{q}+\frac{d}{r}=\frac{d}{2}-1 \tag{6.8.14}
\end{equation*}
$$

then there is a constant $C>0$ so that for all $T>0, h, f, g \in L_{t}^{1}\left(L_{x}^{2}\right) \times H^{1} \times L^{2}$ the solution of

$$
\square u=h, \quad u(0)=f, \quad u_{t}(0)=g,
$$

satisfies

$$
\begin{equation*}
\|u\|_{L_{t}^{q} L_{x}^{r}([0, T])} \leq C\left(\|h\|_{L_{t}^{1} L_{x}^{2}([0, T])}+\left\|\nabla_{x} f\right\|_{L^{2}\left(\mathbb{R}^{d}\right)}+\|g\|_{L^{2}\left(\mathbb{R}^{d}\right)}\right) \tag{6.8.15}
\end{equation*}
$$

Proof. 1. Rewrite the wave equation as a symmetric hyperbolic pseudodifferential system motivated by D'Alembert's solution of the $1-d$ wave equation. Factor,

$$
\partial_{t}^{2}-\Delta=\left(\partial_{t}+i|D|\right)\left(\partial_{t}-i|D|\right)=\left(\partial_{t}+i|D|\right)\left(\partial_{t}-i|D|\right)
$$

Introduce

$$
v_{ \pm}:=\left(\partial_{t} \mp i|D|\right) u, \quad V:=\left(v_{+}, v_{-}\right)
$$

so

$$
V_{t}+\left(\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right) i|D| V=\binom{h}{h} .
$$

Lemma 3.4.8 implies that for $\sigma=d-1, q>2$, $(q, r) \sigma$ - admissible, and $h, f, g$ with spectrum in $\left\{R_{1} \leq|\xi| \leq R_{2}\right\}$ one has

$$
\|u\|_{L_{t}^{q} L_{x}^{r}} \leq C\left\|\nabla_{t, x} u\right\|_{L_{t}^{q} L_{x}^{r}} \leq C\|V\|_{L_{t}^{q} L_{x}^{r}} \leq C\left(\|h\|_{L_{t}^{1} L_{x}^{2}}+\||D| f\|_{L^{2}}+\|g\|_{L^{2}}\right)
$$

2. Denote by $\ell$ the dimensions of $t$ and $x$. With dimensionless $u$, the terms on right of this inequality have dimension $\ell^{d / 2-1}$.
The dimension of the term on the left is equal to

$$
\left(\ell^{d q / r} \ell\right)^{1 / q}=\ell^{\frac{d}{r}+\frac{1}{q}}
$$

The two sides have the same dimensions if and only if

$$
\begin{equation*}
\frac{d}{r}+\frac{1}{q}=\frac{d}{2}-1 \tag{6.8.16}
\end{equation*}
$$

Under this hypothesis it follows that the same inequality holds, with the same constant $C$ for data with support in $\lambda R_{1} \leq|\xi| \leq \lambda R_{2}$.
Comparing (6.8.16) with $\sigma$-admissibility which is equivalent to

$$
\frac{d}{r}+\frac{1}{q} \leq \frac{d}{2}-\frac{1}{2}-\frac{1}{r}
$$

shows that (6.8.16) implies admissibility since $r \geq 2$.
3. Lemma 6.8.6 follows using Littlewood-Paley theory as at the end of §3.4.3.

We now answer the question of when $\square^{-1}$ maps $L_{t}^{1} L_{x}^{2}$ to $L_{t}^{p} L_{x}^{2 p}$. This is the crucial calculation. In Lemma 6.8.6, take $r=2 p$ to find

$$
\frac{1}{q}+\frac{d}{2 p}=\frac{d-2}{2}
$$

so,

$$
\frac{1}{q}=\frac{d-2}{2}-\frac{d}{2 p}=\frac{p(d-2)-d}{2 p}, \quad q=p\left(\frac{2}{p(d-2)-d}\right)
$$

We want $q \geq p$, that is

$$
\frac{2}{p(d-2)-d} \geq 1, \quad \Leftrightarrow \quad p(d-2)-d \leq 2 \quad \Leftrightarrow \quad p \leq \frac{d+2}{d-2}
$$

The critical case is that of equality, and the subcritical case that we treat is the one with strict inequality. For $d=3$ the critical power is $p=5$ and for $d=4$ it is $p=3$. In the subcritical case the operator has small norm for $T \ll 1$.

The strategy of the proof is to write the solution $u$ as a perturbation of the solution of the linear problem, at least for small times. Define $u_{0}$ to be the solution of

$$
\begin{equation*}
\square u_{0}=0, \quad u_{0}(0)=f, \quad \frac{\partial u_{0}}{\partial t}(0)=g \tag{6.8.17}
\end{equation*}
$$

Write

$$
\begin{equation*}
u=u_{0}+v \tag{6.8.18}
\end{equation*}
$$

with the hope that $v$ will be small at least for $t$ small.
Lemma 6.8.7. If $u=u_{0}+v$ with $v \in L_{t}^{p} L_{x}^{2 p}([0, T])$ satisfying

$$
\begin{equation*}
v= \pm \square^{-1} F_{p}\left(u_{0}+v\right) . \tag{6.8.20}
\end{equation*}
$$

then

$$
\begin{equation*}
u \in C\left([0, T] ; H^{1}\left(\mathbb{R}^{d}\right)\right) \cap C^{1}\left([0, T] ; L^{2}\left(\mathbb{R}^{d}\right)\right) \cap L_{t}^{p} L_{x}^{2 p}([0, T]) \tag{6.8.21}
\end{equation*}
$$

satisfies

$$
\begin{equation*}
\square u \pm F_{p}(u)=0, \quad u(0)=f, \quad u_{t}(0)=g \tag{6.8.22}
\end{equation*}
$$

Conversely, if $u$ satisfies (6.8.21)-(6.8.22) then $v:=u-u_{0} \in L_{t}^{p} L_{x}^{2 p}([0, T])$ and satisfies (6.8.21)

Proof. The Strichartz inequality implies that $u_{0} \in L_{t}^{p} L_{x}^{2 p}$ and by hypothesis the same is true of $v$. Therefore $u_{0}+v$ belongs to $L_{t}^{p} L_{x}^{2 p}$ so $F_{p}\left(u_{0}+v\right) \in L_{t}^{1} L_{x}^{2}$.
Therefore $v= \pm \square^{-1} F_{p}$ is $C\left(H^{1}\right) \cap C^{1}\left(L^{2}\right)$. The differential equation and initial condition for $v$ are immediate.
The converse is similar, not used below, and left to the reader.
Proof of Theorem 6.8.4. For $K>0$ arbitrary but fixed, we prove unique local solvability with continuous dependence for $0 \leq t \leq T$ with $T$ uniform for all data $f, g$ with

$$
\|f\|_{H^{1}}+\|g\|_{L^{2}} \leq K
$$

Choose $R=R(K)$ so that for such data,

$$
\left\|u_{0}\right\|_{L_{t}^{p} L_{x}^{2 p}([0,1])} \leq \frac{R}{2}
$$

Define

$$
B=B(T):=\left\{v \in L_{t}^{p} L_{x}^{2 p}([0, T]):\|v\|_{L_{t}^{p} L_{x}^{2 p}([0, T])} \leq R\right\}
$$

We show that for $T=T(K)$ sufficiently small, the map $v \mapsto \square^{-1} F_{p}(u)$ is a contraction from $B$ to itself.
This is a consequence of three facts.

1. Lemma 6.8.5 shows that $F_{p}$ is uniformly lipschitzean from $B$ to $L_{t}^{1} L_{x}^{2}([0, T])$ uniformly for $0<T \leq 1$.
2. Lemma 6.8.6 together with subcriticality shows that there is an $r>p$ so that $\square^{-1}$ is uniformly lipshitzean from $L_{t}^{1} L_{x}^{2}$ to $L_{t}^{r} L_{x}^{2 p}$ uniformly for $0<T<1$.
3. The injection $L_{t}^{r} L_{x}^{2 p} \mapsto L_{t}^{p} L_{x}^{2 p}$ has norm which tends to zero as $T \rightarrow 0$.

This is enough to carry out the existence parts of Theorem 6.8.4.
If there are two solutions $u, v$ with the same initial data, compute

$$
\square(u-v)=G(u, v)(u-v)
$$

Lemma 6.8.6 together with subcriticality shows that with $r$ slightly larger than $p$,

$$
\|u-v\|_{L_{t}^{r} L_{x}^{2 p}} \leq C\|G(u, v)(u-v)\|_{L_{t}^{1} L_{x}^{2}} \leq C\|u-v\|_{L_{t}^{p} L_{x}^{2 p}} .
$$

Use this estimate for $0 \leq t \leq T \ll 1$ noting that Hölder's inequality shows that for $T \rightarrow 0$,

$$
\|u-v\|_{L_{t}^{p} L_{x}^{2 p}} \leq C T^{\rho}\|u-v\|_{L_{t}^{r} L_{x}^{2 p}} \leq C T^{\rho}\|u-v\|_{L_{t}^{p} L_{x}^{2 p}}, \quad \rho>0
$$

to show that the two solutions agree for small times. Thus the set of times where the solutions agree is open and closed proving uniqueness.
To prove the energy law note that $F_{p}(u) \in L_{t}^{1} L_{x}^{2}$ so the linear energy law shows that

$$
\begin{equation*}
\left.\int \frac{\left|u_{t}\right|^{2}+\left|\nabla_{x} u\right|^{2}}{2} d x\right|_{t=0} ^{t}=\mp \int_{0}^{t} \int u_{t} F_{p}(u) d x d t \tag{6.8.23}
\end{equation*}
$$

Now

$$
u_{t} \in L_{t}^{\infty} L_{x}^{2}, \quad \text { and } \quad F_{p}(u) \in L_{t}^{1} L_{x}^{2}
$$

Hölder's inequality shows that

$$
\int\left|u_{t} F_{p}(u)\right| d x \leq\left\|u_{t}(t)\right\|_{L_{x}^{2}} \| F_{p}\left(u(t) \|_{L_{x}^{2}}\right.
$$

The latter is the product of a bounded and an integrable function so

$$
\forall T, \quad u_{t} F_{p}(u) \in L^{1}\left([0, T] \times \mathbb{R}^{d}\right)
$$

Let

$$
w:=\frac{|u|^{p+1}}{p+1} .
$$

Since $p$ is subcritical, one has for some $0<\epsilon$,

$$
\|w(t)\|_{L_{x}^{1}} \leq C\|u(t)\|_{H^{1-\epsilon}\left(\mathbb{R}^{d}\right)} \in L^{\infty}([0, T])
$$

In particular $w \in L^{1}\left([0, T] \times \mathbb{R}^{d}\right)$ and the family $\{w(t)\}_{t \in[0, T]}$ is precompact in $L_{\text {loc }}^{1}$. Formally differentiating yields

$$
\begin{equation*}
w_{t}=u_{t} F_{p}(u) \in L^{1}\left([0, T] \times \mathbb{R}^{d}\right) \tag{6.8.24}
\end{equation*}
$$

Using the above estimates, it is not hard to justify (6.8.24).
It then follows that $w \in C\left([0, T] ; L^{1}\left(\mathbb{R}^{d}\right)\right)$ and

$$
\left.\int w(t, x) d x\right|_{t=0} ^{t=T}=\int_{0}^{T} \int u_{t} F_{p}(u) d x d t
$$

Together with (6.8.23) this proves the energy identity.
Once the energy law is known, one concludes global solvability in the attractive case since the blow up criterion (6.8.11) is ruled out by energy conservation.

## Chapter 3. Appendix I. Perturbation theory for semisimple eigenvalues.

The computation of the form of the operator $\pi L \pi$ requires formulas from the perturbation theory of eigenvalues. These results for multiple eigenvalues which are semisimple is not that well known. The key idea is that one should NOT make a choice of basis of eigenfunctions, but work systematically with the spectral projections.

Definition. An eigenvalue $\lambda$ of a matrix $A$ is semisimple when the kernel and range of $A-\lambda I$ are complementary subspaces. In this case denote by $\pi$ the spectral projection onto the kernel of $A-\lambda I$ along its range and by $Q$ the partial inverse defined by

$$
\begin{equation*}
Q \pi=0, \quad Q(A-\lambda I)=I-\pi . \tag{1}
\end{equation*}
$$

Theorem. Suppose that $] a, b[\ni s \rightarrow A(s)$ is a smooth family of complex matrices with an isolated smooth semisimple eigenvalue $\lambda(s)$. Then $\lambda(s)$ and $\pi(s)$ are smooth functions of $s$ whose first derivatives satisfy

$$
\begin{gather*}
\lambda^{\prime}(s) \pi(s)=\pi(s) A^{\prime}(s) \pi(s),  \tag{2}\\
\lambda^{\prime \prime} \pi=\pi A^{\prime \prime} \pi-2 \pi A^{\prime} Q A^{\prime} \pi  \tag{3}\\
\pi^{\prime}=-\pi A^{\prime} Q-Q A^{\prime} \pi \tag{4}
\end{gather*}
$$

Proof. For $\underline{s}$ fixed, choose $r>0$ so that $A(\underline{s})$ has only the eigenvalue $\lambda(\underline{s})$ in the disc $\mid z-$ $\lambda(\underline{s}) \mid \leq 2 r$. The smoothness of $\pi(s)$ near $\underline{s}$ follows from the contour integral representation

$$
\pi(s):=\frac{1}{2 \pi i} \oint_{|z-\lambda(\underline{s})|=r}(z-A(s))^{-1} d z
$$

The identity

$$
Q(s)=(I-\pi(s))(\pi(s)+A(s))^{-1} \in C^{\infty}
$$

The identity $A(s) \pi(s)=\lambda(s) \pi(s)$ implies that

$$
\lambda(s)=\frac{\operatorname{Trace}(A(s) \pi(s))}{\operatorname{Trace} \pi(s)} \in C^{\infty}
$$

The formulas (2-4) are proved by differentiating the identity $(A-\lambda) \pi=\pi(A-\lambda)=0$ with respect to $s$. The equation for each $d^{j} / d s^{j}$ is analysed by considering its projections $\pi$ and $I-\pi$. Equivalently, each equation is multiplied first by $\pi$, then by $Q$.
Denoting $d / d s$ with a ${ }^{\prime}$. Differentiating $(A-\lambda) \pi$ yields

$$
\begin{equation*}
(A-\lambda)^{\prime} \pi+(A-\lambda) \pi^{\prime}=0 \tag{5}
\end{equation*}
$$

Mulitplying on the left by $\pi$ eliminates the second term to yield

$$
\begin{equation*}
\pi(A-\lambda)^{\prime} \pi=0 \tag{6}
\end{equation*}
$$

which is equivalent to (2).
Multiply equation (5) on the left by $Q$ to find

$$
(I-\pi) \pi^{\prime}=-Q(A-\lambda)^{\prime} \pi
$$

Since $Q \pi=0$ this simplifies to

$$
\begin{equation*}
(I-\pi) \pi^{\prime}=-Q A^{\prime} \pi \tag{7}
\end{equation*}
$$

Equation (5) is exhausted and we take a second derivative,

$$
(A-\lambda)^{\prime \prime} \pi+2(A-\lambda)^{\prime} \pi^{\prime}+(A-\lambda) \pi^{\prime \prime}=0
$$

Mutiply on the left by $\pi$ to eliminate the last term,

$$
\pi(A-\lambda)^{\prime \prime} \pi+2 \pi(A-\lambda)^{\prime} \pi^{\prime}=0
$$

Subtract $2\left(\pi(A-\lambda)^{\prime} \pi\right) \pi^{\prime}=0 \pi^{\prime}=0$ to find

$$
\pi(A-\lambda)^{\prime \prime} \pi+2 \pi(A-\lambda)^{\prime}(I-\pi) \pi^{\prime}=0
$$

Then (7) yields

$$
\begin{equation*}
\pi(A-\lambda)^{\prime \prime} \pi+2 \pi(A-\lambda)^{\prime}\left(-Q A^{\prime} \pi\right)=0 \tag{8}
\end{equation*}
$$

Since $\pi Q=0$ one has

$$
\begin{equation*}
2 \pi \lambda^{\prime}\left(-Q A^{\prime} \pi\right)=0 \tag{9}
\end{equation*}
$$

Adding (8) and (9) yields (3).
To prove (4) knowing (7), what is needed is $\pi \pi^{\prime}$. Differentiate $\pi^{2}=\pi$ to find

$$
\begin{equation*}
\pi \pi^{\prime}+\pi^{\prime} \pi=\pi^{\prime}, \quad \text { whence } \quad \pi \pi^{\prime}=\pi^{\prime}(I-\pi) \tag{10}
\end{equation*}
$$

Differentiate $\pi(A-\lambda)=0$ to find

$$
\pi^{\prime}(A-\lambda)+\pi(A-\lambda)^{\prime}=0
$$

Mulitply on the right by $Q$ to find

$$
\pi^{\prime}(I-\pi)=-\pi(A-\lambda)^{\prime} Q
$$

Use (10) and simplify using $\pi Q=0$ to find

$$
\pi \pi^{\prime}=-\pi A^{\prime} Q
$$

Adding this to (7) completes the proof.

## Chapter 3, Appendix II. The stationary phase inequality.

Definition. A point $\underline{x}$ in an open subset $\Omega \subset \mathbb{R}^{d}$ is a stationary point of $\phi \in C^{\infty}(\Omega ; \mathbb{R})$ when $\nabla_{x} \phi(\underline{x})=0$. It is nondegenerate when the matrix of second derivatives at $\underline{x}$ is nonsingular..

When $\underline{x}$ is a nondegenerate stationary point the map $x \mapsto \nabla_{x} \phi(x)$ has nonsingular Jacobian at $\underline{x}$. It follows that the map is a local diffeomorphism and in particular the stationary point is isolated.
Taylor's Theorem shows that

$$
\nabla_{x} \phi(x)=\frac{1}{2} \nabla_{x}^{2} \phi(\underline{x})(x-\underline{x})+O\left(|x-\underline{x}|^{2}\right) .
$$

Therefore if $\omega \subset \subset \Omega$ contains $\underline{x}$ and no other stationary point, nondegeneracy implies that there is a constant $C>0$ so,

$$
\begin{equation*}
\forall \underline{x} \in \omega, \quad\left|\nabla_{x} \phi(x)\right| \geq C|x-\underline{x}| . \tag{1}
\end{equation*}
$$

We estimate the size of oscillatory integrals whose phase has a single nondegenerate stationary point. These integrals have a complete asymptotic expansion. Proving the estimate is easier than deriving the expansion. The estimate is proved by the method of nonstationary phase. I learned the dyadic proof below from G. Métivier. See [Stein, Harmonic Analysis, Real Variable Methods] for an alternate proof.

Theorem. Suppose that $\phi \in C^{\infty}(\Omega ; \mathbb{R})$ has a unique stationary point $\underline{x} \in \Omega$. Suppose that $\underline{x}$ is nondegenerate and let $m$ denote the smallest integer strictly larger than $d / 2$. Then for any $\omega \subset \subset \Omega$ there is a constant $C$ so that for all $f \in C_{0}^{\infty}(\omega)$, and $0<\epsilon<1$,

$$
\begin{equation*}
\left|\int e^{i \phi / \epsilon} f(x) d x\right| \leq C \epsilon^{d / 2} \sup _{|\alpha| \leq m}\left\|\partial^{\alpha} f(x)\right\|_{L^{\infty}(\omega)} . \tag{2}
\end{equation*}
$$

Lemma. There is a nonnegative $\chi \in C_{0}^{\infty}\left(\mathbb{R}^{d} \backslash 0\right)$ so that for all $x \neq 0, \sum_{k=-\infty}^{\infty} \chi\left(2^{k} x\right)=1$.
Proof of Lemma. Choose nonnegative $g \in C_{0}^{\infty}\left(\mathbb{R}^{d} \backslash 0\right)$ so that $g \geq 1$ on $\{1 \leq|x| \leq 2\}$. Define the locally finite sum

$$
G(x):=\sum_{k=-\infty}^{\infty} g\left(2^{k} x\right), \quad G\left(2^{k} x\right)=G(x)
$$

Then $G \in C^{\infty}\left(\mathbb{R}^{d} \backslash 0\right)$, and $G \geq 1$. The function $\chi:=g / G$ has the desired properties.

Proof of Theorem. Translating coordinates we may suppose that $\underline{x}=0$. Choose $\chi$ as in the lemma and write

$$
\int e^{i \phi / \epsilon} f(x) d x=\sum_{k=-\infty}^{\infty} \int \chi\left(2^{k} x\right) e^{i \phi / \epsilon} f(x) d x:=\sum_{k=-\infty}^{\infty} I(k)
$$

The half sum $\sum_{k<0} \chi\left(2^{k} x\right)$ is a smooth function on $\mathbb{R}^{d}$ which vanishes on a neighbhorhood of the origin and is identically equal to 1 outside a large ball. The nonstationary phase Lemma 1.2.2 implies that

$$
\left|\int e^{i \phi / \epsilon}\left(\sum_{k<0} \chi\left(2^{k} x\right)\right) f(x) d x\right| \leq C \epsilon^{m} \sup _{|\alpha| \leq m}\left\|\partial^{\alpha} f(x)\right\|_{L^{1}(\omega)}
$$

The sum $\sum_{2^{k} \epsilon^{1 / 2} \geq 1} \chi\left(2^{k} x\right)$ is a bounded function supported in a ball $|x| \leq C \epsilon^{1 / 2}$ so

$$
\left|\int e^{i \phi / \epsilon}\left(\sum_{2^{k} \epsilon^{1 / 2} \geq 1} \chi\left(2^{k} x\right)\right) f(x) d x\right| \leq C \epsilon^{d / 2}\|f(x)\|_{L^{\infty}(\omega)} .
$$

There remains the sum over $1 \leq 2^{k}<\epsilon^{-1 / 2}$. The change of variable $y=2^{k} x$ yields

$$
I(k)=2^{-k d} \int \chi(y) e^{i \phi_{k}(y) /\left(2^{2 k} \epsilon\right)} f\left(2^{-k} y\right) d y, \quad \phi_{k}(y):=2^{2 k} \phi\left(2^{-k} y\right)
$$

It follows from (1) that there is a constant $c>0$ so that on the support of $\chi$,

$$
c^{-1} \leq\left|\nabla \phi_{k}\right| \leq c .
$$

In addition there is are constants $C(\alpha)$ independent of $k \geq 0$ so that $\left|\partial^{\alpha} \phi_{k}\right| \leq C_{\alpha}$. The method of nonstationary phase shows that there is a constant independent of $k \geq 0$ so that

$$
\left|\int \chi(y) e^{i \phi_{k}(y) /\left(2^{2 k} \epsilon\right)} f\left(2^{-k} y\right) d y\right| \leq C\left(2^{2 k} \epsilon\right)^{m} \sup _{|\alpha| \leq m}\left\|\partial^{\alpha} f(x)\right\|_{L^{1}(\omega)}
$$

Therefore

$$
\sum_{1 \leq 2^{k}<\epsilon^{-1 / 2}}|I(k)| \leq C \epsilon^{m} \sum_{1 \leq 2^{k}<\epsilon^{-1 / 2}} 2^{-k d} 2^{2 k m} \sup _{|\alpha| \leq m}\left\|\partial^{\alpha} f(x)\right\|_{L^{1}(\omega)}
$$

The finite geometric sum has ratio $r=2^{2 m-d}>1$. If $K$ is the largest index,

$$
r^{K} \leq 1+r+r^{2} \ldots+r^{K}=\frac{r^{K+1}-1}{r-1}<\frac{r}{r-1} r^{K}:=C(r) r^{K}
$$

The sum is comparable to the last term. Therefore, with $C=C(m, d)=r /(r-1)$,

$$
\epsilon^{m} \sum_{1 \leq 2^{k}<\epsilon^{-1 / 2}} 2^{-k d} 2^{2 k m} \leq C \epsilon^{m}\left(2^{K}\right)^{2 m-d} \leq C \epsilon^{m}\left(\epsilon^{-1 / 2}\right)^{2 m-d}=C \epsilon^{d / 2}
$$

This completes the proof.
Corollary. Suppose that $\phi(x, \zeta)$ is a family of phases depending smoothly on $\zeta$ on a neighborhood of $0 \in \mathbb{R}^{q}$ and that $\phi(x, 0)$ satisfies the hypotheses of the preceding Theorem. Then there is a neighborhood $0 \in \mathcal{O}$ so that the hypotheses are satisfied for $\zeta \in \mathcal{O}$ and the estimate (1) holds with a constant independent of $\zeta \in \mathcal{O}$.

Proof. The first assertion follows from the implicit function theorem applied to the system of equations $\nabla_{x} \phi(x, \zeta)=0$. The estimates of the proof are all locally uniform which proves the second assertion.

## Chapter 3, Appendix III. Hadamard's Ovaloid Theorem.

Theorem (Hadamard). If $d \geq 3$ and $\mathcal{M}$ is an oriented connected compact immersed hypersurface of $\mathbb{R}^{d}$ whose Gaussian curvature is nonzero at all points, then $\mathcal{M}$ is the boundary of a strictly convex set.

Proof. Consider the Gauss map $\mathcal{N}$ from $\mathcal{M}$ to $S^{d-1}$ which takes a point to its unit normal consistent with the orientation.

The nonvanishing curvature is equivalent to the differential of $\mathcal{N}$ being invertible at all points. The inverse function theorem shows that this is equivalent to $\mathcal{N}$ being a local diffeomorphism.
For any $\xi \in S^{d-1}$ the point(s) $\underline{x} \in \mathcal{M}$ where $x . \xi$ is maximal have normal equal to $\xi$ so $\mathcal{N}$ is surjective.
The number of preimages of points is finite and locally constant, hence constant. Therefore $\mathcal{N}$ is a covering map.
Since $S^{d-1}$ is simply connected, it follows that $\mathcal{N}$ is a homeomorphism and therefore a diffeomorphism. We recall the proof.
It suffices to show that $\mathcal{N}$ is injective. If $\mathcal{N}\left(m_{1}\right)=\mathcal{N}\left(m_{2}\right)=p \in S^{d-1}$ choose a curve $\gamma_{0}:[a, b] \rightarrow \mathcal{M}$ connecting $m_{1}$ to $m_{2}$. The image $\mathcal{N} \circ \gamma$ is a closed curve $\mu_{0}$ in $S^{d-1}$ beginning and ending at $p$.
Simple connectivity implies that there is a homotopy of closed curves $\mu_{t}$ for $0 \leq t \leq 1$ beginning and ending at $p$ with $\mu_{1}$ reducing to the constant path $p$.
Since $\mathcal{N}$ is a covering, the homotopy lifting lemma shows that there is a homotopy $\gamma_{t}$, $0 \leq t \leq 1$ so that $\mathcal{N} \circ \gamma_{t}=\mu_{t}$.
The point $\gamma_{t}(a)$ is a point of $\mathcal{M}$ depending continuously on $t$ with $\mathcal{N}\left(\gamma_{t}(a)\right)=p$. It follows that $\gamma_{t}(a)$ is constant and therefore equal to $m_{1}$. Similarly $\gamma_{t}(b)=m_{2}$. In particular this holds for $t=1$.
But $\gamma_{1}$ is a lifting of the constant map $\mu_{1}$ and is therefore constant. Therefore

$$
m_{1}=\gamma_{1}(a)=\gamma_{1}(b)=m_{2}
$$

proving injectivity.
Thus each vector in $S^{d-1}$ is normal to $\mathcal{M}$ at exactly one point. This shows that $\mathcal{M}$ is strictly convex in the sense that it intersects each tangent plane in exactly one point.
That it is strictly convex in the stronger sense of osculating ellipsoids, then follows from the nonvanishing Gaussian curvature.

Example. A curve in $\mathbb{R}^{2}$ with positive curvature and looping twice about the origin shows that the result is not true when $d=2$.

## References

S. Alinhac, Equations Differentielles: Etude Asymptotique; Application aux Equations aux Dérivées Paritelles, Lecture Notes, University of Paris Sud, Orsay, 1980.
S. Alinhac, Blowup for Nonlinear Hyperbvolic Equations Birhäuser, 1995.
[Ba] S. Banach, Thorie des Operations Linaires, reprint, Chelsea Publishing Co., New York, 1955.
[Be] M. Beals, Propagation and Interaction of Singularities in Nonlinear Hyperbolic Problems, Birkhäuser, Boston, 1989.
[BR] R. Benedetti and J.-L. Risler, Real Algebraic and Semialgebraic Sets, Actualités Mathématiques, Hermann (1990).
[ Br ] P. Brenner, The Cauchy problem for symmetric hyperbolic systems in $L_{p}$, Math. Scand. 19(1966), 27-37.
[BvW] P. Brenner and W. von Wahl, Global classical solutions of nonlinear wave equations, Math. Zeit. 176(1981), 87-121.
[Bo] J.-M. Bony, Calcul symbolique et propagation des singularités pour les équations aux dérivées nonlinéaires, , Ann. Sci. Ecole Norm.Sup. 14(1981), 209-246.
[Bl] N. Bloembergen, Nonlinear Optics, Benjamin, New York, 1964.
A. Cauchy, Mémoire sur l'intégration des équations differentiells, Exercises d'analyse et de physique mathématiques, vol. 1, Paris 1840.
Y. Choquet-Bruhat, Ondes asymptotiques et approchées pour les systèmes d'équations aux dérivées partielles non linéaires, J. Math. Pures. Appl. 48(1969)117-158.
R. Courant, Methods of Mathematical Physics, vol. II., Interscience Publishers, 1962.
R. Courant, K.O. Friedrichs, and H. Lewy, Über dei partiellen differenzengleichungen der physik, Math. Ann. 100(1928-1929)32-74.
[D] P. Donnat, Quelque contributions mathématiques in optique non linéaire, Ph.D. thesis, Ecole Polytéchnique, Paris, 1994.
P. Donnat and J. Rauch, Modelling the dispersion of light, in Singularities and Oscillations, eds J. Rauch and M. Taylor, IMA Volumes in Mathematics and its Applications, Springer Verlag, 1997.
P. Donnat and J. Rauch, Dispersive nonlinear geometric optics, Jour. Math. Phys. 38(3)(1997)1484-1523.
[Du] J.J. Duistermaat, Fourier Integral Operators Courant Institute of Mathematical Sciences Lecture Notes, 1973.
[F] K.O. Friedrichs, Symmetric hyperbolic linear differential equations, Comm. Pure Appl. Math. 7(1954)345-392.
[Ga] A. Gabor, Remarks on the wave front set of distribution, Trans. AMS 170(1972) 239-244.
V. Georgiev, S. Lucente, and G. Zilotto, Decay estimates for hyperbolic systems, Hokkaido Math. J. 33(2004) 83-113.
[GV1] Ginibre and G. Velo, Generalized Strichartz inequalities for the wave equation. J. Funct. Anal. 133 (1995)50-68.
[GV2] J. Ginibre and G. Velo, The global Cauchy problem for the nonlinear Klein-Gordon equation. Math. Z. 189 (1985) 487-505.
[Gu1] O. Gues, Ondes multidimensionelles epsilon stratifiées et oscillations, Duke Math. J. 1992
[Gu2] O. Gues, Dévelopment asymptotique de solutions exactes de systèmes hyperboliques quasilinéaires, Asympt. Anal. 6(1993)241-269.
[GR] O. Guès and J. Rauch, Hyperbolic $L^{p}$ multipliers are translations, to appear.
[Нö 1] L. Hörmander, Fourier intergral operators I, Acta Math. 127 (1971)79-183.
[Hö 2] L. Hörmander, The Analysis of Linear Partial Differential Operators vols.I,II, Springer-Verlag, Berlin, 1983.
L. Hörmander, Lectures on Nonlinear Hyperbolic Differential Equations, Springer-Verlag, Berlin-New York-Heidelberg, 1997.
J. Hunter, and J. Keller, Weakly nonlinear nonlinear waves, Comm. Pure Appl. Math. 36(1983)547-569.
J. Hunter, A. Majda, and R. Rosales, Resonantly interacting weakly nonlinear hyperbolic waves II, Stud. Appl. Math. 75(1986)187-226.
[I] N. Iwasaki, Local decay of solutions for symmetric hyperbolic systems with dissipative and coercive boundary conditions in exterior domains, Publ RIMS Kyoto U. 5(1969)193218.
F. John, Partial Differential Equations $4^{\text {th }}$ ed., Springer-Verlag, New York, 1982.
J.L. Joly and J. Rauch, Justification of multidimensional single phase semilinear geometric optics, Trans. A.M.S. 330(1992)599-623.
J.L. Joly, G. Métivier and J. Rauch, Formal and rigorous nonlinear high frequency hyperbolic waves, in Nonlinear Hyperbolic Waves and Field Theory M.K. Murthy and S. Spagnolo eds, Pitman Research Notes in Math no. 253, 1992, pp. 121-143.
J.L. Joly, G. Métivier and J. Rauch, Resonant one dimensional nonlinear geometric optics, J. Funct. Anal. 114(1993)106-231.
J.L. Joly, G. Métivier and J. Rauch, Coherent nonlinear waves and the Wiener algebra, Ann. Inst. Fourier, 44(1994)167-196.
J.L. Joly, G. Métivier and J. Rauch, Generic rigorous asymptotic expansions for weakly nonlinear multidimensional oscillatory waves, Duke Math. J. 70(1993)373-404.
J.L. Joly, G. Métivier and J. Rauch, A nonlinear instability for $3 \times 3$ systems of conservation laws, Comm. Math. Phys. 162(1994)47-59.
J.L. Joly, G. Métivier and J. Rauch, Diffractive nonlinear geometric optics with rectification, Indiana U. Math. J. 47(1998) 1167-1242.
J.L. Joly, G. Métivier and J. Rauch, Dense oscillations for the compressible two dimensional Euler equations, in Nonlinear Partial Differential Equations and their Applications, College de France Seminar 1992-1993, eds. D. Cioranescu and J.L. Lions, Pitman Research Notes in Math. no. 391, Longman publs. 1998.
J.L. Joly, G. Métivier and J. Rauch, Dense oscillations for the Euler equations II, in Hyperbolic Problems: Theory, Numerics, Applications, eds. J. Glimm, M.J. Graham, J.W. Grove, B.J. Plohr, World Scientific, 1994.
J.L. Joly, G. Métivier and J. Rauch, Diffractive nonlinear geometric optics, in Séminaire Equations aux Dérivées Partielles, Ecole Polytéchnique, Palaiseau, 1995-1996. Available on the homepage www.math.lsa.umich.edu/~rauch
J.L. Joly, G. Métivier and J. Rauch, Hyperbolic domains of dependence and HamiltonJacobi equations, JHDE to appear.
K. Kajitani and A. Satoh, Time decay estimates for linear symmetric hyperbolic systems with variable coefficients and its applications, in Phase Space Analysis of Partial Differential Equations, eds. F. Colombini and L. Pernazza, Centro De Giorgi, Scuola Normale Superiore, Pisa, 2004.
M. Keel and T. Tao, Endpoint Strichartz estimates, Amer. J. Math. 120 (1998), 955-980.
J. Keller, On solutions of nonlinear wave equations. Comm. Pure Appl. Math. 10(1957)523 530.
P.D. Lax, On Cauchy's problem for hyperbolic equations and the differentiability of solutions of elliptic equations. Comm. Pure Appl. Math. 8 (1955), 615-633.
P.D. Lax, Hyperbolic systems of conservation laws II, Comm. Pure. Appl. Math. 10(1957)537-566.
P.D. Lax, Asymptotic solutions of oscillatory initial value problems, Duke Math. J. 24((1957)627-646.
P. D. Lax, Lectures on Hyperbolic Partial Differential Equations, Stanford University Lecture Notes, 1963.
P.D. Lax, Shock waves and entropy, in Contributions to Nonlinear Functional Analysis, eds E. Zarantonello, Academic Press, NY, 1971.
J. Leray, Lectures on Hyperbolic Partial Differential Equations Institute for Advanced Study, 1953.
[Lud1] D. Ludwig, Exact and asymptotic solutions of the Cauchy problem, Comm. Pure Appl. Math. 13 (1960)473-508.
[Lud2] D. Ludwig, Conical refraction in crystal optics and hydromagnetics, Comm. Pure Appl. Math. 14(1961)113-124.
[Lun] R.K. Luneburg, The Mathematical Theory of Optics, Brown Univ. Press, 1944.
A. Majda, Nonlinear geometric optics for hyperbolic systems of conservation laws, in Oscillations Theory, Computation, and Methods of Compensated Compactness eds. C. Dafermos, J. Ericksen, D. Kenderlehrer, and I. Muller, IMA Volumes on Mathematics and its Applications vol. 2 pp. 115-165, Springer-Verlag, New York, 1986.
A. Majda and R. Rosales, Resonantly interacting weakly nonlinear hyperbolic waves, Stud. Appl. Math. 71(1986)149-179.
A. Majda, R. Rosales, and M. Schonbek, A canonical system of integro-differential equations in nonlinear acoustics, Stud. Appl. Math. 79(1988)205-262.
[MT] T.S. Motzkin and O. Tausky, Pairs of matrices with property L, Trans. AMS 73(1952), 108-114.
D. McLaughlin, G. Papanicolaou, and L. Tartar, Weak limits of semilinear hyperbolic systems with oscillating data, pp. 277-298 in Macroscopic Modeling of Turbulent Flows, Lecture Notes in Physics vol 230, Springer-Verlag, 1985.
[Me] G, Métivier, ref for quasilinear
[Mey] Y. Meyer, Remarque sur un théoreme de J.-M. Bony, Suppl. de Rendiconti del Circ. Mat. di Palermo, Atti. del Seminario di Analisi Armonica, Pisa, Série 2, 11981.
G. Métivier and J. Rauch, Real and complex regularity are equivalent for hyperbolic characteristic varieties, Diffewrential and Integral Equations, to appear.
G. Peano, Sull'inegrabilità della equazioni differenziali di primo ordine, Atti. Acad. Torino, 21A(1886) 677-685.
[Ral] J. Ralston, thesis article
J. Rauch, An $L^{2}$ proof that $H^{s}$ is invariant under nonlinear maps for $s>n / 2$, in Global Analysis - Analysis on Manifolds, ed. T, Rassias, Teubner Texte zur Math., band 57, Leipzig, 1983.
J. Rauch, Partial Differential Equations, Graduate Texts in Math 128, Springer-Verlag, New York, 1991.
J. Rauch, Lectures on Geoemtric Optics, in Nonlinear Wave Phenomena, eds. L. Caffarelli and W. E., IAS/Park City Math Series Vol. 5, 1998.
J. Rauch, Finite speed with bare hands,
[RR1] J. Rauch and M. Reed, Nonlinear microlocal analysis of semilinear hyperbolic systems in one space dimension, Duke Math. J. 49(1982)379-475.
[RR1] J. Rauch and M. Reed, Bounded, stratified, and striated solutions of hyperbolic systems, in Nonlinear Partial Differential Equations and Their Applications Vol. IX, H. Brezis and J. L. Lions, eds., Pitman Research Notes in Math., (181)(1989), 334-351.
[RT] J. Rauch and M. Taylor, Exponential decay of solutions to hyperbolic equations in bounded domains, Indiana U. Math. J. 24(1974)79-86.
A. Satoh, Scattering for nonlinear symmetric hyperbolic systems, Preprint, 2004.
J. Shatah and M. Struwe, Geometric Wave Equations, Courant Lecture Notes in Mathematics, New York University, Courant Institute of Mathematical Sciences, New York; American Mathematical Society, Providence, RI, 1998.
[Scha] Schauder, quasilinear exisitence
S. Schochet, Fast singular limits of hyperbolic equations, J. Diff. Eq. 114(1994)476-512.
C. Sogge, Lectures on Nonlinear Wave Equations, International Press, 1995.
E. Stein, Singular Integrals and Differentiability Properties of Functions, Princeton University Press, 1970.
W. Strauss, Nonlinear Wave Equations, CBMS Regional Conf. Series, no. 73, AMS, 1989 [Ta 1] M. Taylor, Pseudodifferential Operators, Princeton Math Series 34, Princeton University Press, 1981.
[Ta 2] M. Taylor, Partial Differential Equations, Basic Theory, Springer-Verlag, 1997.
T. Wagenmaker, Analytic solutions and resonant solutions of hyperbolic partial differential equations, Ph,D Thesis, University of Michigan, 1993.
G. Whitham, Linear and Nonlinear Waves, Wiley-Interscience, New York, 1974.
[Y] Yudovich

